H

HABER-BOSCH PROCESS. See **Ammonia.**

HABIT. As used by the mineralogist, this term denotes the sum of the external characteristics of a mineral. It is also, but more rarely, applied to rocks.

HABITAT. See **Ecology.**

HABIT PLANE. Many phenomena, such as twinning and martensite transformations, occur in metals where plate-like structures develop inside crystals. The crystallographic plane or planes of the parent phase parallel to the sides of these plates are called the habit plane or planes of the phenomena.

HACHURE. A short line drawn parallel to the slope as a means of illustrating topography on a map.

HACKBERRY AND ZELKOVA TREES. These trees are members of *Ulmaceae* (elm family). The hackberry tree (*Celtis occidentalis*) is found in the United States, principally in the eastern part—from the coast west to Indiana. Other concentrations are found in Colorado and New Mexico. The hackberry is a medium-to-large tree, on the average attaining a height between 50 and 100 feet (15 to 30 meters). As shown by the accompanying table, some specimens attain greater heights. The twig is red-brown, having leafy scars that are small and oval. The bud is small, approximately $\frac{1}{8}$ inch (0.3 centimeter) long, pointed, and somewhat flattened. The leaf is from 4 to 6 inches (10 to 15 centimeters) long. It is individual, alternate, and simple. It features sharp teeth with deep veins. The fruit is a drupe, small, and deep purple in color. It ripens in September and October. The taste is bitter. The bark is an ash gray, rough, near wartlike as the tree grows old. The wood is heavy, compact, and pale-yellow in color. It weighs approximately 40 pounds per cubic foot (643 kilograms per cubic meter). Other species of hackberry include the Georgia hackberry (*Celtis tenuifolia*); the Lindheimer hackberry (*C. lindheimer*); and the netleaf hackberry (*C. reticulata*).

Zelkova trees are ornamental trees of attractive habit and handsome foliage. They are deciduous with alternate leaves and polygamous flowers. They have a 1-seeded drupe. Five species are found in Crete, the Caucasus, and eastern Asia. *Zelkova serrata* is an important timber tree, having very durable wood and considered one of the best of building materials in Japan. Young wood is yellowish-white; old wood is a dark brown and known for a beautiful grain. Zelkovas appear much as small-leaved elms, but sometimes take on a shrubby appearance. Small greenish flowers and the fruits are inconspicuous. *Z. serrata* and *Z. Davidii* are also found in North America. They are hardy and can withstand northern climates. The *Z. ulmoides* is less hardy and usually is not found north of Massachusetts. Because of its upright stems, *Z. Davidii* makes an excellent shrub. *Z. serrata* can attain a height up to 100 feet (30 meters), featuring a broad, round-topped head and slender branches.

RECORD HACKBERRY TREES IN THE UNITED STATES[1]

Specimen	Circumference[2]		Height		Spread		Location
	Inches	Centimeters	Feet	Meters	Feet	Meters	
Common hackberry (1972) (*Celtis occidentalis*)	248	630	126	38.4	112	34.1	Michigan
Georgia hackberry (1991) (*Celtis tenuifolia*)	17	43	28	8.5	17	5.2	Virginia
Georgia hackberry (1991) (*Celtis tenuifolia*)	15	38	28	8.5	24	7.3	Virginia
Lindheimer hackberry (1975) (*Celtis lindheimeri*)	72	183	43	13.1	46	14.0	Texas
Netleaf hackberry (1989) (*Celtis reticulata*)	180	457	69	21.0	75	22.9	New Mexico

[1]From the "National Register of Big Trees," The American Forestry Association (by permission).
[2]At 4.5 feet (1.4 meters).
Note: If two trees of same species are listed, they are cochampions.

HACKLY. A term used by mineralogists to describe a jagged fracture.

HADDOCK. See **Codfishes.**

HADFIELD STEEL. See **Manganese.**

HADRONS. These are subatomic particles, the strong interactions of which are manifested by the forces that hold neutrons and protons together in the atomic nucleus. Hadrons include the proton, the neutron, and pion, among others. These particles show signs of an inner structure, i.e., they are made up of other particles, which has led over a period of the last several years to consider the hadrons as combinations of constituents known as *quarks*. See also **Particles (Subatomic).**

HAFNIUM. Chemical element symbol Hf, at. no. 72, at. wt. 178.49, periodic table group 4, mp 2207–2247°C, bp 4601–4603°C, density 13.3 g/cm^3. The alpha form of elemental hafnium has a close-packed hexagonal crystal structure; the beta form, a body-centered cubic struc-

D. M. Considine et al. (eds.), *Van Nostrand's Scientific Encyclopedia*

ture. Metallic hafnium, like zirconium, exhibits passivity in air due to formation of adherent coatings of oxide or nitride. Urbain reported evidence of the element in 1911, but hafnium was not fully identified until 1923 by D. Coster and G. C. de Hevesy. The remarkable similarity between hafnium and zirconium accounts mainly for its late isolation, as compared with the majority of elements. In terms of abundance, there is an average of about 4 ppm hafnium in the earth's crust. The element occurs with zirconium in certain varieties of zircon, including malacon, cyrtolite, and alvite. One mineral found in Scandinavia, thortveitite, contains more hafnium than zirconium. Pegmatite, monazite, baddeleyite, and zerkelite also contain hafnium. First ionization potential 5.5 eV. Oxidation potentials $Hf + H_2O \rightarrow HfO^{2+} + 2H + 4e^-$, 1.68 V;$Hf + 4OH^- \rightarrow HfO(OH)_2 + H_2O + 4e^-$, 2.60 V. Electron configuration $1s^22s^22p^63s^24d^{10}4s^24p^64d^{10}4f^{14}5s^25p^65d^26s^2$. Ionic radius Hf^{+4}, 0.75Å. Other important physical properties of hafnium are given under **Chemical Elements.**

Hafnium usually is extracted from ores along with zirconium. In one process, zircon sand is broken down by carbiding or carbonitriding, followed by chlorination. The mixture formed is dissolved with a complexing agent, after which it is introduced into a liquid-liquid extraction process. The final product is $HfCl_4$. Fractional crystallization of the fluorides of hafnium and zirconium also is practiced. Metallic hafnium is made by the Kroll process in which the $HfCl_4$ is reduced in an inert atmosphere by magnesium. The hafnium sponge and magnesium chloride resulting is vacuum-distilled to accomplish the final separation. In a modified Kroll process, sodium or sodium amalgam may be used. The latter requires less rigid temperature and pressure control during processing, costs less, and introduces fewer impurities into the process. For further purification of hafnium metal, a number of methods have been used, including electrorefining, arc and induction melting, zone refining, and the hot-wire or van Arkelde Boer process.

Uses. Compared with most metals, the annual production of hafnium is low. Mainly produced in the United States, France, and Russia, the combined production is in the range of 100 metric tons annually, or less. Several uses have been found for hafnium: (1) as a control material in water-cooled nuclear reactors. Also hafnium is an effective flux-depressor in a reactor for absorbing neutrons to decrease the peaks in neutron flux; (2) as a filament in gas-filled incandescent light bulbs; (3) as an alloying ingredient to add strength to tungsten and molybdenum filaments and electrodes used in high-pressure discharge tubes; (4) as a cathode in x-rays tubes; (5) as a getter material in vacuum tubes and systems; (6) as a minor alloying ingredient in nichrome heating elements where hafnium appears to significantly increase the lifespan of the elements; and (7) usually with zirconium, as an ingredient of several alloys.

Chemistry and Compounds. Hafnium metal dissolves in HCl (warm) and slowly in H_2SO_4, more rapidly if fluoride ion F^- is present, forming compounds of HfO^{2+}, or fluoro complexes in the latter case. The metal resists the attack of weak acids and their salts.

Due to its $5d^26s^2$ electron configuration, hafnium forms tetravalent compounds readily, although the Hf^{4+} ion does not exist as such in aqueous solution except at very low pH values, the common cation being HfO^{2+} (or $Hf(OH)_2^{2+}$) and many of the tetravalent compounds are partly covalent. There are also less stable Hf(III) compounds. There is close similarity in chemical properties to those of zirconium due to the similar outer electron configuration ($4d^25s^2$ for zirconium) and the almost identical ionic radii (Zr^{4+} is 0.80Å) the relatively low value for Hf^{4+} being due to the Lanthanide contraction.

With improved means to separate the compounds of these two elements, future research will yield more details of specific hafnium compounds. The methods of separation used effectively include ion exchange techniques, a particularly effective one using a column of silica gel, with a solution of the tetrachlorides in methanol as feed and a 1.9 *N* HCl solution as eluant for zirconium. Separations also have been accomplished through the distillation of the phosphorus oxychloride addition products.

See list of references at end of entry on **Chemical Elements.**

HAGFISHES (*Agnatha*). A jawless fish of the family *Myxinidae*, is an aggressive scavenger usually averaging less than 30 inches (76 centimeters) in length. The hagfish is characterized by the primitive features of jawless fishes—no scales, no sympathetic nervous system, a cartilage skeleton, and single nostril. The hagfish is elongate, rather wormlike, and blind. Because the fish can exude large quantities of a slimy mucus, it is sometimes called a "slime eel." Among species of hagfishes are the Japanese *Paramyxine*, the *Eptatretus*, the Atlantic *Myxine glutinosa*, and the Pacific *Heptatretus stouti*. The latter species has been used in medical research, particularly in studies of the hag heart (no heart nerves or symphathetic nerves). Generally, hagfishes prefer cold to temperate marine waters from shallow levels down to about 3,000 feet (900 meters). They cannot tolerate fresh or brackish waters.

See also **Cyclostomata;** and **Fishes.**

HAIDINGER FRINGES. Optical interference fringes seen with thick, flat plates near normal incidence. The fringes of the Fabry-Perot interferometer are of this type. They are also known as constant angle or constant deviation fringes.

HAIL. See **Precipitation and Hydrometeors.**

HAIR. There are several kinds of hair on the human body. The appearance depends on age and body location. The so-called *lanugo* is that hair which develops on the unborn child. Usually, it is shed before birth, or within the first few months after birth. The lanugo is immediately replaced by secondary hair which is fine and soft and is often called "baby hair." The coarser hair of later life is called *tertiary hair*. Hairs are continually lost from all parts of the body throughout life, and up to a certain age, those which replace them often are coarser than their predecessors.

There are about 125,000 hairs on the scalp of the average person. Darker persons usually have fewer scalp hairs than blonds. Scalp hair usually grows from 3 to 5 inches (7.5 to 12.5 centimeters) per year and, if permitted, can become as long as 2 to 3 feet (0.6 to 0.9 meter), or even longer.

The hairs of the body originate from hair follicles embedded in the skin. The lower part of the follicle extends into the dermis where it is supplied with blood vessels. Generally, only one hair grows from a single follicle. That part of the hair beneath the surface of the skin is termed the *root*, while that part extending outward from the skin is called the *shaft*. The sebaceous glands of the skin have their openings in the hair follicles. These glands secrete a substance (sebum) which is responsible for the oily appearance of the skin or scalp. Persons with oily skin possess overactive sebaceous glands. When the hair follicle becomes plugged, the sebum collects within it, turns dark at the surface, and becomes a "blackhead."

Minute muscles (*erectors pilorum*) are connected to the hair follicle. When these muscles contract, they temporarily displace the entire follicle, causing the hair to "stand on end." The skin surrounding the hair is also elevated by the contraction of these muscles, giving the skin a prickled appearance, sometimes called "goose pimples." Contraction of the muscles also exerts pressure on the sebaceous glands, causing the emission of extra amounts of sebum. Thus, this set of reactions aids in protecting the body from sudden cold, the hairs forming better insulation when standing erect, and the sebum coats the skin with a further barrier against the cold.

The partial or complete absence of hair from the body is called *alopecia*.

The use of hair analysis and examination in the forensic sciences has been known for a number of years. Regarding the analysis of Sir Isaac Newton's hair for mercury, see **Mercury.**

HAIR (Abnormal Growth). See **Hirsutism.**

HAIR HYRROMETER. See **Hygrometer.**

HAIRSTREAK (*Insecta, Lepidoptera*). Small butterflies, those of the temperate zone dull-colored and those of the tropics often brilliant.

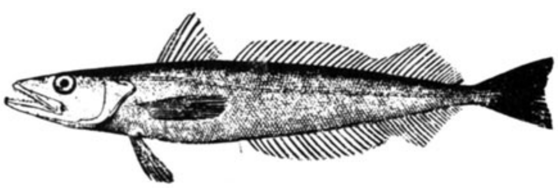
The hind wings of most species bear hairlike tails. With the coppers and blues they make up the family *Lycaenidae*.

HAIRTAIL. See **Cutlassfishes.**

HAIRWORM (*Nematomorpha* or *Gordiacea*; formerly placed in the phylum Nemathelminthes with the *Nematoda*). Long slender roundworms of small size, which live as parasites in the bodies of invertebrates, chiefly insects.

HAKE (*Osteichthyes*). Of the family *Merluccidae*, the hakes are closely related to the codifishes, but have a special systematic position due to their unusual distribution. The family has just one genus, *Merluccius*. The slender body, skull structure, and the large-toothed mouth give this carnivorous fish a garlike appearance. There are two dorsal fins and one long anal fin, which is almost the mirror image of the second dorsal fin in shape, size, and position.

The hake (*Merluccius merluccius*) is found in the northeastern Atlantic Ocean off the western and southwestern coasts of Europe, along the continental shelf. The northern border of the distribution is formed where branches of the Gulf Stream meet masses of polar waters. This is also the northern limit of the *American hake* or *silve hake (Merluccius bilnearis*). See accompanying figure. Living in deep water has enabled the hakes to penetrate the tropical Atlantic Ocean and inhabit oceanic regions in the southern hemisphere with temperate to subtropical conditions. This accounts for the large South Atlantic populations of *stockfish (Merluccius capensis*) off southwestern Africa; and *Merluccius hubbsi* from the coasts of southern Brazil and Argentina. There are also Pacific Ocean species: *Merluccius gayi* and *M. productus*, off the western coasts of North and South America. Their presence has been explained by a presumed migration around Cape Horn. The New Zealand species, *Merluccius australis*, may also have come by this route.



Silver hake.

Hakes can be over 3 feet (1 meter) in length, but there are small- and medium-sized species as well. They are predators, feeding chiefly on herring and other schooling fishes. The European hake seeks its prey at night in the upper water levels. During the day, it is less active and stays near the floor, at which time it can be caught easily, even with a dragnet. This species spawns in spring, apparently without preferred spawning sites. The floating eggs then drift within the hake distribution region. The commercial importance of hake has increased since the early 1960s.

The *Cape hake* in South African waters is a whitefish (*M. capensis*) and it appears that there are two distinct populations on the trawling grounds. About half of the population attains sexual maturity at an age of 3 to 4 years. Peak spawning occurs during spring and early summer. Preliminary studies indicate that in the case of the Cape hake, diurnal vertical migration is much less pronounced than in the case of the European cod. The diet of the adult hake is comprised of rattails, maasbanker, and squid, and cannibalism is quite common.

See also **Fishes.**

HALF-ADDER. A circuit having two output points, S and C, representing sum without carry and carry, and two input points, A and B, representing addend and augend, such that the output is related to the input according to the following table:

Input		Output	
A	B	S	C
0	0	0	0
0	1	1	0
1	0	1	0
1	1	0	1

Two half-adders and an Inclusive-OR circuit, properly connected, can provide a Full-Adder having two inputs (augend and addend) and a carry input which produces a sum output (without carry) and a carry output.

HALF-CELL. An electrochemical system consisting of a single electrode and an electrolytic solution, with usually a (reversible) ionization process in progress between electrode and electrolyte. See also **Galvanic Cell.**

HALF-LIFE (Biological). The time of survival of half the individual members of an unstable system. The half-life $t_{1/2}$ of the system is related to the decay constant λ and the mean life τ by the relation:

$$t_{1/2} = \frac{\ln 2}{\lambda} = \frac{0.693}{\lambda} = 0.693\tau$$

The term half-life is most commonly applied to systems of radionuclides but may also be applied to other systems that decay.

The biological half-life of a substance is the time in which a living tissue, organ or individual eliminates, through biological processes, one-half of a given amount of a substance which has been introduced into it. The effective half-life is a term usually applied to a radioactive substance in a biological organism. It is defined in terms of the half-life of the radioactive substance itself, and its biological half-life in the organism, by the following expression:

$$\text{effective half-life} = \frac{\text{radioactive half-life} \times \text{biological half-life}}{\text{radioactive half-life} + \text{biological half-life}}$$

HALF-LIFE (Elements). See **Chemical Elements.**

HALF-SILVERED SURFACE. A surface coated with a metallic film of such thickness that it transmits approximately half of the light falling on it at normal incidence and reflects approximately half.

HALF-THICKNESS (Absorber). The thickness of a particular absorber that will reduce the intensity of a beam of radiation to one-half its initial value. If the absorption is exponential, the half-thickness is related to the linear or mass absorption coefficient and the mean free path as follows:

$$d_{1/2} = \frac{\ln 2}{\mu} = \frac{0.693}{\mu} = 0.693l$$

where $d_{1/2}$ is the half-thickness, μ is the absorption coefficient and l is the mean free path.

HALFWIDTH OF A SPECTRAL LINE. The intensity within a spectral line may be expressed as $I(x)$, where x is a measure of wavelength, frequency or wave number, and where $I(x)\,dx$ is a measure of the contribution to the intensity between x and $x + dx$. The halfwidth of the line is the halfwidth of the function $I(x)$.

HALIBUT. See **Flatfishes.**

HALIDES. A compound made up of a halogen (astatine, bromine, chlorine, fluorine, or iodine) and another element or radical may be termed a *halide*. Fundamentally, there are three classes: (1) the *ionic* (saline) halides, (2) the *covalent* (acid) halides, and (3) the *complex* halides. The ionic halides are most sharply characterized by the halides of the alkali and alkaline earth metals, plus those of certain Lanthanide and Actinide metals. They form ionic or semi-ionic crystals in the solid state, have high boiling points and melting points, and are soluble in polar solvents. Their bonding is electrovalent, varying in degree with the difference between the electronegativities of the halogen and the metal. Potassium iodide and silver fluoride are ionic, but silver iodide is essentially covalent. The fluorides exhibit a primarily ionic character for most of the metals, but the other halogens form fewer ionic compounds. The degree of ionicity varies down as well as across the periodic table.

The covalent (acid) halides have low boiling and melting points, are soluble in nonpolar solvents and insoluble in polar solvents, although they often react with the latter. The degree of covalence generally is greatest for the nonmetals. For a given nonmetal, the boiling point depends upon both the number of atoms of the halogen with which it is combined and the symmetry of the molecule. For example, the boiling points of bromine(I) fluoride, bromine(III) trifluoride, and bromine(V) pentafluoride, BrF, BrF_3 and BrF_5, are 20, 135, and 40.5°C, respectively.

The complex halides are very numerous, because of the readiness with which halide ions form coordination compounds with metals. In general, stability of these complexes depends upon the size and electronic structure of the metal ion—the smaller cations from their more stable compounds with the smaller halide ions, notably with fluoride, while with larger cations the order of stability is that of polarizability of the halide, i.e., decreasing from iodide to fluoride. The more electronegative transition elements form especially stable complexes; e.g., those of palladium, platinum, etc., $PdCl_4^{2-}$, PtF_6^{2-}, etc. The most common halo complexes have four or six halogen ions coordinated with the cation, although such complexes as those of copper, gold and mercury, e.g., CuI_2^-, $AuCl_2^-$, $HgCl_3^-$, etc., are notable exceptions.

See also **Bromine; Carbon; Chlorine; Chlorinated Organics; Fluorine;** and **Iodine.**

HALITE (Rock Salt). The mineral halite (rock salt) is naturally occurring sodium chloride, NaCl, common salt. It is isometric with cubic habit and cleavage. It is brittle; hardness, 2.5; specific gravity, 2.168; luster, vitreous; colorless when pure, but usually white, yellow, red, or blue. It is soluble in water. Halite occurs interbedded with sedimentary rocks in all parts of the world and in all but the very oldest rocks. It frequently occurs in association with anhydrite and gypsum. In the United States this type of "salt beds" has been exploited in Michigan, New York, Ohio, and Pennsylvania. Louisiana produces salt from great subsurface dome-shaped masses, often 2,000–4,000 feet thick. The salt domes of the Gulf Coastal Plain are particularly important as subsurface structures, on the flanks of which are apt to occur large and important pools of petroleum. Poland, Saxony, Austria, and France possess well-known deposits of salt, as well as the former U.S.S.R., England, Algeria, India, and China. Salt is chiefly used in cooking and as a preservative; in the manufacture of soda ash for the glass industry; and as a source of many sodium compounds. It derives its name from the halogen group of elements to which chlorine belongs.

See also **Sodium Chloride.**

HALL EFFECT AND QUANTIZED HALL EFFECT. In 1879, Edwin H. Hall (Johns Hopkins University), discovered that if a strip of gold leaf, carrying an electric current longitudinally, was placed in a magnetic field with the plane of the strip perpendicular to the direction of the field, the points directly opposite each other on the edges of the strip acquired different electric potentials; and that if such points were joined through a sensitive galvanometer, a feeble current would be indicated. In other words, the equipotential lines, ordinarily running across at right angles to the edges, were skewed into an oblique position, and the electric lines of flow in the plane of the strip were deflected to one side.

If one looks along the strip in the direction of the current, with the magnetic field directed downward, then, with strips of antimony, cobalt, zinc, or iron, the electric potential drop is toward the right and the effect is said to be positive; while with gold, silver, platinum, nickel, bismuth, copper, and aluminum, it is toward the left, and the effect is called negative. The transverse electric potential gradient per unit magnetic field intensity per unit current density is called the "Hall coefficient" for the metal in question. Thus, the Hall coefficient R_H is defined as

$$R_H = \frac{E_y}{j_x H_z}$$

where E_y is the electric field developed in the y direction when a current of current density j_x flows in the x direction through a magnetic field H_z in the z direction. According to the free electron theory of metals, the Hall coefficient should be given by

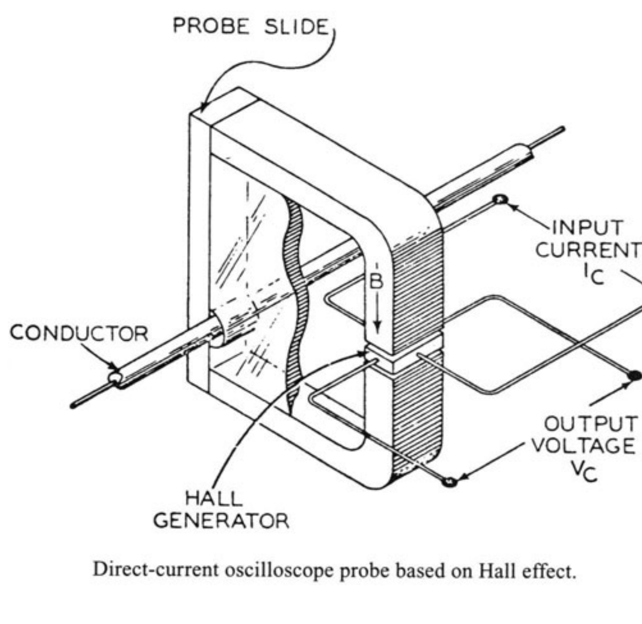
$$R_H = \frac{B}{ne}$$

where N is the number of free electrons per unit volume, of charge e (in esu), and c is the velocity of light. The observed result that for some metals the carriers would seem to have positive charges is explained by the band theory of solids. In a nearly filled band, the wave functions of the electrons near the top of the band are so modified that it is the holes in the band that behave like particles. Since a hole represents the absence of negative charge, it behaves as if positively charged. The Hall angle is the ratio of E_y (defined above) to the field E_x, generating the current in the magnetic field H_z. The Hall mobility is the mobility of the electrons or holes in a semiconductor as measured by the Hall effect.

A number of transducers utilize the Hall effect. Shown in the accompanying diagram is a direct-current oscilloscope probe based on the effect. A steady direct current I_c is applied to one axis of the Hall generator and a magnetic field B, proportional to the current through the conductor, is applied to a second axis. An output voltage V_c is taken across the third axis of the Hall generator. The output voltage can be calculated from:

$$V_c = \frac{10^{-5} R_H}{t} I_c B$$

where Vc = Hall voltage, volts
R_H = Hall corfficient, cm 3/coulomb
t = thickness, cm
B = magnetic field density, kilogauss



Direct-current oscilloscope probe based on Hall effect.

Exploring the Complexities of the Hall Effect

Over the intervening century since Hall's discovery and notably since the advent of semiconductor technology, the Hall effect has inspired research. A number of related effects have been observed. One of these, for example, is the widely studied galvanomagnetic effect, referred to as *transfer magnetoresistance*. By shorting the Hall field or by choosing a disk geometry so that such a field does not exist, one obtains a "magnetoresistance" (more strictly, a *magnetoconductivity*) which does not saturate. This is called the Corbino magnetoresistance or Corbino effect. There are several thermal effects in a magnetic field which can produce transverse voltages or temperature gradients. These result from the velocity separation of charge carriers by the Lorentz force—the energetic ones going to one side, the slower ones going to the other. Temperature gradients are produced, and also electric fields. In the Righi-Leduc effect, a longitudinal temperature gradient produces a transverse temperature gradient (thermal analog of the Hall effect). In the Nernst effect, it produces a transverse electric field. In the Ettingshausen effect, a longitudinal electric current produces a transverse temperature gradient. This latter effect, if large, can disturb the Hall field, since the potential probes and leads are seldom made of the same material as the specimen. Therefore, the Ettingshausen temperature gradient can produce a thermoelectric voltage which adds to the Hall voltage.

Analysis of Hall-effect data has been one of the most widely used techniques for studying conduction mechanisms in solids, especially semiconductors. For the single-carrier case, one readily obtains carrier concentrations and mobilities, and it is usually of interest to study these as functions of temperature. This can supply information on the predominant charge-carrier scattering mechanisms and on activation energies, i.e., the energies necessary to excite carriers from impurity levels into the conduction band. Where two or more carriers are present, the analysis becomes more complex, but much more information can be obtained from studies of the temperature and magnetic field dependencies.

Unlike, for example, the magnetoresistance, the Hall effect is a first-order phenomenon. A weak magnetic field, it depends linearly on the magnetic field intensity and it does not vanish in isotropic solids if all the carriers have essentially the same velocity or if the scattering is characterized by a relaxation time which is independent of the carrier energy. As previously indicated, the Hall effect forms the basis of a number of devices used in isolating circuits, transducers, multipliers, converters, rectifiers, and gaussmeters (for measurement of magnetic fields). The fundamental component of such devices is a slab of material (often called a "Hall generator") possessing favorable Hall characteristics.

The Quantized Hall Effect

In 1980, Klaus von Klitzing[1] (High Magnetic Field Laboratory, Max Planck Institute), made some unusual findings while studying the Hall effect in devices in which the electrons free to carry current are confined within a *thin layer* of material. The researcher found that by cooling an experimental device to within a degree of absolute zero and by placing the device in a *very strong* magnetic field, the behavior of the ordinary resistance and the Hall resistance differed dramatically from that expected of a traditional Hall-effect device. Instead of increasing steadily and linearly as the strength of the magnetic field was increased, the Hall resistance increased in a *series of plateaus*. There were intervals observed in which the Hall resistance did not vary at all when the strength of the magnetic field was varied. Between the plateaus, the Hall resistance increased smoothly with increasing magnetic field. It was also found that during the same intervals of magnetic field strength during which the Hall resistance exhibited plateaus, the voltage drop parallel to the current was noted to disappear completely (no electrical resistance in sample and current flows without dissipating any energy).

The vanishing electrical resistance and the plateaus in the Hall resistance are remarkable phenomena. It is even more remarkable, as pointed out by Halperin (see reference), that on each plateau the value of the Hall resistance satisfies a remarkably simple condition, i.e., the reciprocal of the Hall resistance is equal to an integer multiplied by the square of the charge on the electron and divided by Planck's constant (the fundamental constant of quantum mechanics). Each plateau is characterized by a different integer. Essentially, in such a system, the Hall resistance is reduced to the formula

$$R_H = \frac{1}{Nce}$$

where n = density of electrons (per square meter) in the sample. If the two-dimensional system is connected to an external reservoir of electrons and the magnetic field B is allowed to vary, then the density of electrons in the layer will vary with B in such a way as to minimize the combined energy of the layer and reservoir.

The degree of precision of the quantized Hall effect has amazed even the experts. Measured values of the Hall resistance at various integer plateaus are accurate to about one part in six million. The effect can be used to construct a laboratory standard of electrical resistance that is much more accurate than the standard resistors currently in use. Authorities also observe that, if the quantized Hall effect is combined with a new calibration of an absolute resistance standard, it should be able to yield an improved measurement of the fundamental dimensionless constant of quantum electrodynamics, the fine-structure constant α.

In his original experiment, von Klitzing used a silicon field-effect transistor (MOSFET) of exceptional quality and of the type used on integrated circuit chips. In the device, electrons are trapped in a so-called inversion layer near the surface of a silicon crystal that is covered with a film of insulating silicon oxide, on top of which is deposited a metal "gate electrode," used to control the density of conduction electrons in the inversion layer.

A somewhat similar Hall effect phenomenon, known as the *fractional quantized Hall effect*, was observed at the National Magnet Laboratory (Cambridge, Massachusetts) by Tsui, Störmer, and Gossard (AT&T Bell Laboratories) a couple of years after von Klitzing's finding. The fractional quantized Hall effect was first noted in a heterojunction (an interface of crystals made of two different semiconducting materials). As pointed out by Halperin, in a heterojunction, electrons from one semiconductor are attracted to more energetically favorable locations in the other semiconductor. The positive charge thereby created in the "donor" semiconductor provides a force attracting the electrons back, however, and they become trapped in a thin layer at the interface of the two crystals.

Additional Reading

Beer, A. C.: "Hall Effect and Related Phenomena," in *The Encyclopedia of Physics* (R. M. Besancon, Ed.), 3rd Ed., Van Nostrand Reinhold, New York, 1980.
Betalin, I. A., Isham, C. J., and G. A. Vileovisky, Editor: "Quantum Field Theory and Quantum Statistics," Hilger, Bristol, U.K., 1987.
Davies, P., Editor: "The New Physics," Cambridge Univ. Press, New York, 1989.
Eisenstein, J. F., and H. L. Stormer: "The Fractional Quantum Hall Effect," *Science*, 1510 (June 22, 1990).
Ellis, P. J., and Y. C. Tang, Editors: "Trends in Theoretical Physics," Addison-Wesley, Redwood City, California, 1990.
Halperin, B. I.: "Theory of the Quantized Hall Conductance," *Helvetica Physica Acta*, 56(4603), 1241–1246 (June 17, 1983).
Halperin, B. I.: "The Quantized Hall Effect," *Sci. Amer.*, 52–60 (April 1986).
Halperin, B. I.: "The 1985 Nobel Prize in Physics," *Science*, **231**, 820–822 (1986).
Nicholas, R. J., Ed.: "Proceedings of the Fifth Int. Conf. on Electrical Properties of Two-Dimensional Systems," in *Surface Science*, **142** (1984).
Schwartzschild, B.: "Von Klitzing Wins Nobel Prize for Quantum Hall Effect," *Physics Today*, **38**(12), 17–20 (December 1985).
von Klitzing, K., Dorada, G., and M. Pepper: *Phys. Rev. Lett.*, **45**, 494 (1980).

HÄLLEFLINTA. A Swedish term for hard, dense, metamorphic rocks composed chiefly of microscopic crystals of quartz and feldspar with occasional phenocrysts. Accessory minerals may be hornblende, chlorite, hematite or magnetite. The texture and composition of hälleflinta suggests that it is the metamorphosed equivalent of acid lava flows or tuffs.

HALLEY'S COMET. This is probably the most famous of all the comets. It is the brightest periodic comet, and so was the first to have its return predicted. In 1705, Edmund Halley (whose name almost certainly rhymed with *valley*) computed the orbit of the great comet that he and others observed in 1682, and found the elements to be almost

[1]Nobel Prize winner for Physics, 1985.

identical with those that he derived for the prominent comets observed by Kepler and Longomontanus in 1607 and by Peter Apian in 1531. He noted that the intervals between the three dates were not quite identical, and correctly attributed the difference to gravitational perturbations of the comet's motion by the planets. Celestial mechanics had not yet advanced to the stage at which planetary perturbations could be readily evaluated, either to prove the aforementioned conclusion about the slight inequality in the intervals of the comet's return, or to enable a completely accurate prediction of the next return after 1682. Nonetheless Halley's predicted time for the next return, late 1758 or early 1759, did partially allow for the effect of Jupiter, the most important of the perturbing planets. The comet was actually first seen again on Christmas night of 1758. By that time, with the aid of improved mathematical methods, perihelion passage had been computed beforehand by Clairaut, Lalande, and Madame Lepaute, and predicted to be within a month of mid-April, 1759. The comet actually passed perihelion on March 13 of that year. With each successive return since 1682, Halley's comet has had its position determined with methods of increasing precision. This, coupled with increasingly sophisticated computational procedures, has led to increasingly accurate predictions for subsequent returns: in 1835 it passed perihelion within a few days of the predicted time, and for the return of 1910 the agreement was better still. A 1971 ephemeris for the 1986 return made some allowance for nongravitational forces (see **Comet**), namely, the partially known effects of asymmetric emission of materials from the nucleus. This ephemeris apparently predicted the February 9, 1986 perihelion correctly within $1\frac{1}{2}$ hours.

Following the apparition of 1910, it was possible to compute the dates of perihelion passage backward for many centuries. Examination of ancient records—which, prior to the fourteenth century, are mainly Chinese—has enabled observations of Halley's comet to be identified with confidence as far back as 87 B.C. This is no small achievement, in view of the planetary perturbation problem on the one hand, and, on the other, of the fact that many comets as bright as or brighter than Halley's have been recorded.

Fig. 1. Visual aspect of Halley's comet on the morning of May 12, 1910. The Square of Pegasus is on the left; the planet Venus is on the right. The comet tail is about 32 degrees long. The observer was in Mexico.

Fig. 2. Lowell Observatory photograph of Halley's comet on May 15, 1910. Venus, its image greatly enlarged by overexposure, is on the right.

At its return in 1910, Halley's comet was first picked up by Wolf at Heidelberg, on Sept. 11, 1909, at 5×10^8 km from the sun, intermediate between the distances of Mars and Jupiter. It approached the sun in the evening sky as a telescopic object, passing within the earth's orbit on the far side from the sun, and passed perihelion, at 0.59 au, on April 20. It emerged into the morning sky in the first weeks of May as a beautiful naked-eye object for those possessing a dark sky. See Figs. 1 and 2. It then turned eastward and passed between the earth and sun; according to computation the head of the comet actually transited the solar disk on May 19 (universal time), although it was undetectable in doing so. Around this time, experienced observers with good skies traced the tail to a distance of 120° from the below-horizon head. The earth grazed the comet's tail and probably passed through it, and patent-medicine vendors advertised concoctions for warding off the effects of the comet's tail, which had, however, no detectable terrestrial effects. A few days later the head of the comet was visible to the naked eye in the evening sky, despite interference by bright moonlight (except for a total lunar eclipse on May 23). It remained a naked-eye object, for experienced observers with good skies, throughout most of June. It was followed photographically, at that epoch already a more sensitive method than visual detection, until July 1, 1911, when it was 8.3×10^8 km from the sun or slightly more distant than Jupiter. It should have reached aphelion, more distant than the planet Neptune, in 1949.

Return of the Comet in 1986. On its next return, the comet was seen on October 16, 1982, almost $3\frac{1}{2}$ years before perihelion, by observers using a charge-coupled device on the Mt. Palomar 200-inch telescope. The comet was magnitude 24.2 and 1.6×10^9 km from the sun, or slightly more distant than the planet Saturn. It was within 8 seconds of arc of its predicted position.

At perihelion in 1986, Halley's comet was almost directly behind the sun. See Fig. 3. As had been foreseen, this circumstance made the 1985/86 return the poorest for visual observation in many centuries. This was especially true for dwellers of the north temperate zone, since, as the comet moved out from behind the sun, it moved for about six weeks in the skies of the far south. Experienced observers agree that, especially for northern observers, the past two decades have provided several comets that were visually more impressive than Halley's 1985/86 display; certainly comet Bennet (1969) in April, 1970 far outperformed the recent Halley. Nevertheless, even northern observers saw comet Halley with the naked eye, in dark skies. Moreover, from within northern cities the comet was readily visible in binoculars and small telescopes in December and January of 1985/86, and again in mid-March, late April, and the first days of May. The far-southern history occurred from the second half of March through most of April. The urban appearance of the comet was much like the central bulge of the Andromeda nebula, which, as was the comet at its brightest, is a naked-eye object in reasonably dark skies. A naked-eye tail was visible in March and April of 1986 (on dark skies), estimates of length ranging from a few degrees to a few tens of degrees.

Fig. 3. Warner and Swasey Observatory photograph of Halley's comet on March 8, 1986.

Undoubtedly the most important thing about the recent visit of Halley's comet to the inner solar system is the fact that it was met by no less than five spacecraft, sent by the European Space Agency (ESA), Russia, and Japan. The ESA flyby, called Giotto, made much the closest approach, penetrating deeply within the comet's head and coming within about 500 km of the solid nucleus on March 14, 1986. An excellent view was also made by the aging *Pioneer Venus* spacecraft.

Since Giotto was launched in the direction of the Earth's motion around the sun while Halley's comet moves in the opposite sense, the spacecraft met the comet at the terrific relative velocity of 68 km sec^{-1} Consequently the imaging optics were badly sand-blasted by cometary dust long before closest approach to the nucleus, and the best view of the nucleus was transmitted at 11,000 km separation. The nucleus appeared oblong, about 15×7 km in size. Outgassing and dust emission were occurring at only a few areas on the sunlit portion of the nucleus, whose albedo is estimated at 2–4% (close to indirect ground-base estimates). See Fig. 4.

Both Giotto and the Russian flybys were instrumented to analyze the comet dust, which proved to be rich in H, C, N, and O. Hence it appears not unlikely that comet dust includes organic materials. The flybys also greatly strengthened the evidence that a major and probably principal molecule present in the nucleus is H_2O. The "dirty snowball" model of the cometary nucleus, proposed by Fred Whipple in the 1950s, is thus vindicated.

Fig. 4. This picture is a composite of 60 of the images of Comet Halley that were taken by the Halley Multicolour Camera during the GIOTTO flyby of the nucleus. The very successful GIOTTO mission was ESA's first interplanetary mission. The spacecraft flew to within 600 km (375 miles) of the nucleus at 00:03:02 UT on 14 March, 1986. The Halley Multicolour Camera (HMC) was built by an international team led by the Max Planck Institut für Aeronomie, FRG. West Germany provided the electronics, structure and mechanisms, France the optics, Italy the optical baffle and deflecting mirror, and Belgium the optical simulator. The United States' participation was directed by NASA and performed by Ball Aerospace Systems Division, which provided preliminary engineering design, baffle design, program management services and image processing of the camera data, and is participating in the analysis of the images.

In this picture the sun is to the left and north is up. The bright areas are the source regions for the active dust jets. The dark night side of the nucleus is silhouetted against light scattered from dust which lies on the far side of the nucleus. Sunlight is illuminating the nucleus from an angle of 107 degrees from the viewing direction, and the sunrise terminator can be clearly seen in the picture. The complex surface structure can be seen at the foot of the dust jets. The bright spot in the night side of the nucleus is caused by the top of a "hill" which extends up into the morning sunlight. The surface has very low reflectivity (about 4 percent), suggesting that surface is covered by a dark mantle which is porous and traps the light. The surface is a good insulator, separating the surface at a temperature (in sunlight) of 300K to 400K (100F to 250F) from the icy core with a temperature of 50 to 100K (-400F to -300F). The active regions are associated with cracks in the surface mantle caused by thermal stress. The long dimension of the nucleus is about 15km (9.25 miles) and the short dimension about 8km (5 miles). The images were taken at distances of 20,000km to 4000km (12,500 to 2500 miles) from the nucleus at a relative velocity of 68 km/sec (153,000 miles per hour). The resolution varies from 800 meters ($\frac{1}{2}$ mile) at the lower right to 100 meters (330 feet) at the foot of the bright jet. The nucleus rotates about an axis perpendicular to the long dimension with a period of about 54 hours, and there is some evidence for a nutation period of about 7.3 days.

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References

Halley's summary of his comet work was published in *Synopsis Astronomiae Cometicae*, Oxford 1705, and *Tabulae Astronomicae*, London 1749. See also Shapley and Howarth, *A Source Book in Astronomy*. Contemporary English-language accounts of the 1910 apparition of Halley's comet are given in the journals *The Observatory*, Vol. 33, and

Popular Astronomy, Vol. 18. All the securely identified perihelion passages of the comet are listed in Marsden's *Catalogue of Cometary Orbits* (Smithsonian Astrophysical Observatory, 5th ed., 1970). The 1971 calculation of the 1985–86 ephemeris is by Brady and Carpenter, published in *The Astronomical Journal*, **76**, 728 ff. Calculations by Yeomans, leading to the date of the 2061 return, are discussed in *The Astronomical Journal*, **82**, 435 ff.; this paper also lists in detail the early basic papers. The spacecraft Giotto was named for the painter Giotto; see also "Giotto's Portrait of Halley's Comet," by R. J. M. Olson, *Sci., Amer.*, **240**(5), 160–170 (1979).

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Additional Reading

Balsiger, H. K., Fechtig, H., and J. Geiss: "A Close Look at Halley's Comet," *Sci. Amer.*, 96 (September 1988).

Belton, M. J. S.: "P/Halley: The Quintessential Comet," *Science*, 1229 (December 13, 1985).

Brandt, J. C., and M. B. Niedner, Jr.: "The Structure of Comet Tails," *Sci. Amer.*, 49 (January 1986).

Cowen, R.: "Frozen Relics of the Early Solar System," *Science News*, 248 (April 21, 1990).

Gore, R.: "Much More Than Met the Eye: Halley's Comet '86," *Nat'l Geographic*, 758 (December 1986).

Reitsman, H. J., Delamare, W. A., and F. L. Shipple: "Active Polar Region on the Nucleus of Comet Halley," *Science*, 198 (January 13, 1989).

Smyth, W. H., et al.: "Analysis of the Pioneer-Venus Lyman-Alpha Image of the Hydrogen Coma of Comet P/Halley," *Science*, 1008 (August 30, 1991).

HALLUCINOGENS. There are many substances which will, if taken in appropriate quantities, produce distortion of perception, vivid images, or hallucinations. Most of these substances will produce powerful peripheral as well as the central effects. Some few agents are characterized by the predominance of their actions on mental and psychic functions. This group of drugs has been called hallucinogens, psychotomimetics, psycholytics, and psychodelics, among several ambiguous terms. None of these names is adequately descriptive of these compounds.

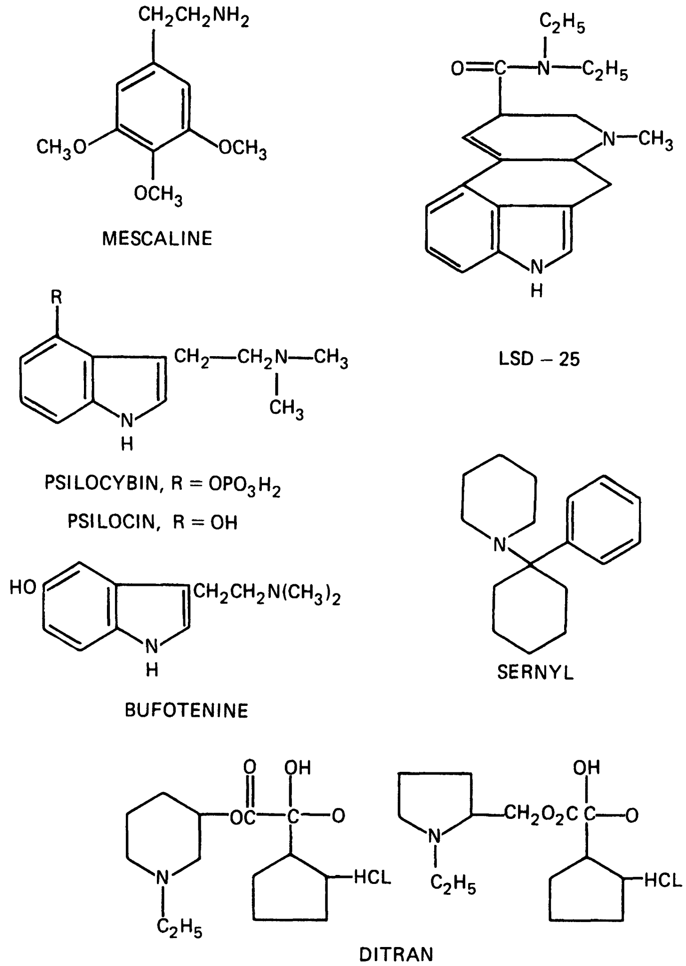
Hallucinogens may be classified into five groups of chemically distinct compounds: (1) lysergic acid derivatives of which lysergic acid diethylamide (LSD-25) is the prototype; (2) phenylethylamines, such as mescaline; (3) indolealkylamines, which include psilocybin, psilocin, and bufotenin; (4) piperidyl benzilate esters, typified by ditran (a 70:30 mixture of N-ethyl-2-pyrrolidymethyl phenylcyclopentylglcolate and N-ethyl-3-piperidyl phenylcyclopentylglycolate), and (5) phenylcyclohexyl piperidines (sernyl). The chemical structures of these compounds is shown in the accompanying figure.

Drugs from the first three groups have been isolated from naturally-occurring sources. LSD-25 is a molecular component of ergot, a fungus which infects cereal grains. Mescaline, historically the oldest hallucinogen, was isolated from a Mexican peyote cactus. Psilocybin and psilocin were isolated from the Mexican mushroom, *Psilocybe mexicana*. Bufotenin is found in some varieties of toadstools. The indole derivatives are chemically closely related to serotonin (5-hydroxytryptamine), a compound which plays an important, yet unknown role in the central nervous system.

The piperidyl benzilate esters and phenylcyclohexyl piperidines are synthetic compounds, and have not been shown to occur naturally. Some authorities do not consider them to be hallucinogens, but active researchers in the field include them among the most active psychotomimetics.

Clinical syndromes from LSD-25, mescaline, and the indoleamines are similar. Somatic symptoms are nausea, dizziness, loss of appetite, blurred vision, paresthesia, weakness, drowsiness, and trembling. These result frequently and are usually associated with sympathomimetic effects, such as increased pulse rate and slight temperature elevation. Perceptual and psychic changes are marked. Visual illusions and vivid hallucinations, decreased concentration, slow thinking, depersonalization, dreamy states, changes in mood, and often anxiety are commonly found.

The clinical syndromes from ditran are different from those produced by the aforementioned drugs in some respects. Disorganization of thought, disorientation, confusion, mood changes, and visual and auditory hallucinations are observed. The piperidyl benzilate esters are central anticholinergics, and mental states produced by them are reminiscent of those from other anticholinergics, such as scopolamine.



Structures of some hallucinogenic drugs.

The effects of phenylcyclohexyl derivatives are also distinctive. Comparatively minor somatic symptoms are evoked. Psychic effects predominate, being typically characterized by feelings of unreality, depression, anxiety, and delusional or illusional experiences. The effects of these drugs are said to be more analogous to natural psychoses than those of the other drugs; however, the same claim has been made for ditran.

When LSD-25 was discovered, it was believed that the drug would provide an extremely useful tool in the investigation of psychoses and mental illness. However, therapeutically, the hallucinogens, including LSD-25, have been of little value to psychiatrists.

HALO. See **Atmospheric Optical Phenomena.**

HALOCARBONS (Ozone Depletion). See **Oxygen.**

HALOGENATED COMPOUNDS. See **Chlorinated Organics; Organic Chemistry.**

HALOGEN GROUP. The elements of group 17 (formerly 7a) of the periodic classification sometimes are referred to as the Halogen Group. The individual elements commonly are called *halogens*. In order of increasing atomic number, they are fluorine, chlorine, bromine, iodine, and astatine. The elements of this group are characterized by the presence of seven electrons in an outer shell, and hence have the ability to gain an electron to form negative ions with a completed octet of valence electrons. The halogens present striking similarities of chemical behavior, all being very reactive and, in particular, readily form substitution compounds with numerous organic compounds. Although these elements also have other valences all have a −1 valence in common.

HAMILTONIAN (or Hamiltonian Function of a System). Generally denoted by the symbol H, the Hamiltonian is defined by the equation

$$H(q_k, p_k, t) = -L(q_k, p_k, t) + \sum_{l=1}^{3n} p_l \dot{q}_t(q_k, p_k, t)$$

L is the Lagrangian function of the system, expressed as a function of the coordinates, momenta and time. $\dot{q}_t$ stands for the generalized velocities, also expressed as functions of the coordinates, momenta and time, where q are the coordinates of position, p, those of momentum, and the dot means the derivative with respect to time. n is the number of particles of the system. If the time does not occur explicitly, the system is called conservative, and H is identical with the total energy of the system.

HAMMER FORGING. See **Forging; Iron Metals, Alloys, and Steels.**

HAMMERHEAD SHARKS. See **Sharks.**

HAMSTER. See **Rodentia.**

HAND. The terminal portion of the pectoral appendage of mammals, developed for grasping and in some species largely freed from locomotor uses. True hands appear only in the primates.

The skeletal structure of the hand includes the series of five bones, the metacarpals, which attach it to the wrist, and the five divergent series of phalanges located in the digits. Of the digits, one, the thumb, is placed and articulated so that it can be opposed to the other four, which are fingers. As a result, the appendage can be used for grasping like a forceps, and also as a prehensile organ by folding the fingers back against the palm. In some of the monkeys, the prehensile method of grasping is more important in moving through the trees, and the thumb has shifted and become smaller so that it can no longer be opposed.

The human hand is the most versatile grasping organ in the animal kingdom.

Since the serious introduction of robotics to industry in the early 1970s, designers have devoted much research to studies of the human hand, with the target of duplicating the manipulative skills of the end effectors (robotic hands and fingers). Scientists at the Department of Energy, Massachusetts Institute of Technology, have been studying the explicit manipulation of the human hand, particularly in the development of robots for handling radioactive materials by telerobotics, in which a robot is controlled remotely by human operators. Essentially, robotic hands currently are inferior to the human hand in terms of what may be called a "sense of touch." Common ground is being explored by robot scientists and neurosurgeons.

Outstanding work also is being conducted at the Johns Hopkins Applied Physics Laboratory, which by December 1992 had developed a tactual simulator consisting of a 20 × 20 arrangement of 400 pins spaced as little as 0.4 mm apart, each controlled by its own microprocessor. The array is reported to vibrate up to 400 times a second at varying amplitudes. In some instances, it is believed that the array matches or exceeds the sensory capabilities of the skin. It is interesting to note, however, that the device will sense an area of touch of about 64 square millimeters, roughly the area of one human fingertip. Like humans, robots require more than one finger in most instances to accomplish their assigned tasks. Robotic manipulators are described further in entry on **Robot and Robotics**.

Carpal Tunnel Syndrome. During the last few years, physicians have received an increasing number of complaints from patients who present pain and numbness in the hands and wrists. Known as CTS, this condition is caused by compression of the median nerve at the wrist. This nerve passes through a relatively firm tunnel made up of the wrist bones and the tough transverse carpal ligament. Several tendons also pass through the carpal tunnel. Anything that increases the pressure inside the tunnel can cause nerve compression, as may be present in rheumatoid arthritis, gout, conditions that may cause inflammation, and swelling of the tendons. Repetitive motion involving the hands also can cause tendinitis. This, in turn, causes compression of the median nerve. Women, particularly those in their mid-forties, appear to be affected most frequently. Often, repetitive actions will result from long periods of office and assembly work requiring the hands or even from hobbies.

Symptoms of CTS include intermittent pain, burning, or numbness in the affected hand. Typically, these symptoms worsen at night. Also, there may be a loss of fine hand control or grip strength. In very advanced cases, there may be complete loss of sensation in the affected hand. A series of tests applied to the nerve and muscles may be required to confirm diagnosis of CTS.

A relatively simple surgical procedure usually corrects the situation, but before opting for surgery, the treating physician may suggest that a splint on the wrist be tried for a period or that an injection be made into the carpal tunnel that will relieve the manifestation of the condition at least temporarily. Early diagnosis and intervention most often will prevent chronic weakness, pain, and disability that can result from long-term nerve compression.

Some physicians may prefer to use arthroscopy, a procedure that enables carpal tunnel release without an incision. See also **Arthroscopy.**

HANDEDNESS (Right- and Left-). Defined in terms of the motion of a screw. A *right-handed* screw, when rotated in the sense of Fig. 1(a) (counterclockwise looking down at the page), will move out of the page; when rotated in the sense of Fig. 1(b), a right-handed screw will move into the page. A *left-handed* screw will move into the page in (a) and out of the page in (b). The mirror image of a right-handed screw is left-handed and vice versa.

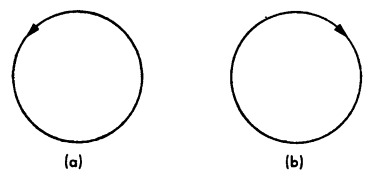


Fig. 1. Screw rotations.

The vector product is also defined in terms of a right-handed screw. Thus **A** × **B** = **C** where the magnitude of **C** is |A| |B| sin θ, and the direction of **C** is given by the direction of progression of a right-handed screw rotating in the sense of rotating **A** into **B** through the smaller angle (θ). **C** is thus a vector pointing into the page in Fig. 2.

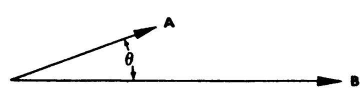
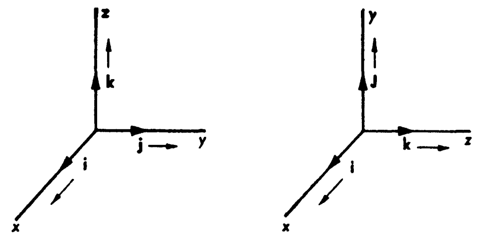


Fig. 2. Vector product.

Coordinate systems are also classed as right or left handed. In Fig. 3, coordinate system (a) is right-handed, since rotation of the unit vector **i** into the unit vector **j** would make a right-handed screw progress in the direction of **k**, **i** × **j** = **k** and also **j** × **k** = **i**, **k** × **i** = **j**. Thus in a right-handed coordinate system, the above cyclic relations among the unit vectors hold. Coordinate system (b), on the other hand, is left-handed, i.e., it would take a left-handed screw to carry the unit vectors into each other in cyclic order of vector multiplication.

Circular polarization of electromagnetic waves is described as right-handed or left-handed depending on whether the direction of rotation of the electric vector and the direction of progression of the electromagnetic wave are related to a right-handed or a left-handed screw.

Handedness is evident throughout many natural phenomena. For example, twining vines, as they grow, wind from left to right (Trumpet



(a) (b)

Fig. 3. Coordinate system.

honeysuckle, *Lonicera sempervirens*) or from right to left (blindweed, *Convolvulus arvensis*). Although snail shells (*Liguus virgineus*) usually are right-handed, some left-handed varieties are known, presumably because of mutations. The bacterium *Bacillus subtilis* also is asymmetric with respect to left and right. Normally, the bacterium forms right-handed spiral colonies, but, when heated, the colonies change to left-handedness.

The broad concept of left and right is widely manifested in the isomeric compounds of chemistry. This effect was discovered by Louis Pasteur in 1848 in connection with his work with tartaric acid. Pasteur found that there are two versions of tartaric acid, even though they are of identical chemical composition. Dextro- and leveotartaric acids became a primary example of optical isomerism. See also **Isomerism,** and **Tartaric Acid**.

Humans are seldom ambidextrous. Most people are right-handed, a fact that does not indicate any relationship to races or cultures. Reliable information on the possible genetics of this characteristic preference remains to be developed.

An elementary particle, such as an electron or proton, is spherically symmetrical (i. e., it is *achiral*). However, once in motion, the spinning particle becomes chiral (i.e., it acquires the characteristic of handedness (chirality)). Some researchers have proclaimed that the universe is asymmetric with respect to chirality.

The foregoing subject is elegantly developed by Hegstrom and Kondepudi (see reference listed).

Additional Reading

Gardner, M., Editor: "The Ambidextrous Universe: Mirror Asymmetry and Time-Reversed Worlds," 2nd Edition, Charles Scribner's Sons, New York, 1979.

Hegstrom, R. A., Chamberlain, J. P., Seto, K., and R. G. Watson: *Amer. J. of Physics*, **56**, *12*, 1086 (December 1988).

Hegstrom, R. A., and D. K. Kondepudi: "The Handedness of the Universe," *Sci. Amer.*, 108 (January 1990).

Kondepudi, D. K., and G. W. Nelson: "Weak Neutral Currents and Origin of Biomolecular Chirality," *Nature*, **314**, *6010*, 438 (April 14, 1985).

Kondepudi, D.: "Parity Violation and the Origin of Biomolecular Chirality," in *Entropy, Information, and Evolution: New Perspectives in Physical and Biological Evolution* (Bruce H. Weber, D. J. Depew, and J. D. Smith, Editors), MIT Press, Cambridge, Massachusetts, 1988.

Ruthen, R.: "Quantum Pinball: A Device That Can't Tell Left from Right," *Sci. Amer.*, 36 (November 1991).

HANGING VALLEY. Under normal conditions a tributary stream enters the main stream at grade, that is, at the same level. Under certain circumstances the tributary valley may be at a greater elevation than the main valley into which the tributary stream will plunge, forming a waterfall. In such cases the tributary valley is called a hanging valley, and the stream in it is said to be out of adjustment with the main stream.

Hanging valleys originate in the following ways: by glacial action, the main glacier cutting down its valley faster than a tributary glacier; by river action, the main stream eroding its bed faster than the tributary stream; by faulting, the tributary stream flowing off the upthrown block. A fourth type of hanging valley, much less common, may result from a stream plunging over wave-cut cliffs or other escarpments into a lake or ocean basin.

HAPLOID. See **Cell (Biology).**

HAPTEN. See **Immune System and Immunology.**

HARDENABILITY OF STEEL. The hardenability of steel refers to the ease with which it can be hardened rather than the maximum hardness value attainable. For example, a 1-inch diameter bar of a certain 0.20% carbon alloy steel can be hardened to 50 Rockwell "C" in the center by quenching in oil. A similar bar of plain carbon steel requires a drastic quench in brine to attain the same hardness, and therefore, has a lower hardenability. Neither bar can be quenched to a greater hardness because 50 Rockwell "C" is the maximum attainable for a 0.20% carbon steel. A 0.40% carbon steel can be hardened to a maximum of about 60 Rockwell "C" and the maximum for high-carbon steel is about 65 Rockwell "C."

Of the several methods for determining the relative hardenability of steels, the Jominy test is the most widely used. A cylindrical specimen 1 inch (2.5 centimeters) in diameter and about 3 inches (7.6 centimeters) long is heated to the hardening temperature and quenched in a special fixture which holds the specimen in a vertical position and directs a stream of water on the bottom surface. The stream takes an "umbrella" shape and does not wet the sides. Cooling occurs progressively from the bottom to the top of the cylinder and the cooling rate at any distance from the bottom is known and reproducible from one sample to another. the hardness along the length of a quenched Jominy bar decreases from bottom to top. The distance from the bottom, expressed in sixteenths of an inch, to the point where the hardness is 50 Rockwell "C" is one method of reporting the hardenability.

One of the principal functions of alloying elements in steel, such as manganese, chromium, nickel, molybdenum, etc., is to increase the hardenability. Whereas prodigious amounts of expensive alloys were formerly used to insure full hardening, especially in medium and heavy sections, wartime shortages focused attention on the use of as little alloy as possible within the hardenability requirements. A large number of steels were developed containing relatively small additions of a number of elements, and a number of these steels have continued in use.

HARDENING OF METALS. There are three principal methods of hardening metals and alloys: cold working (see **Cold-Worked Metal**) by plastic deformation, precipitation hardening, and quench hardening as applied to steel. The last two methods involve heating and cooling operations. A pure metal may also be hardened through the addition of alloying elements. When a solid solution is formed it is normally harder than the pure metal. If additional phases are formed by alloying, these may also be harder than the pure metal and contribute to the hardness of the metal.

HARDENING (Precipitation). See **Precipitation Hardening.**

HARD FACING. Deposition of a hard wear-resistant alloy on a metal surface. The material to be deposited is generally in the form of a welding rod and may be applied by gas or arc welding. Such surfaces are usually finished by grinding.

While hard facing or hard surfacing is usually a maintenance operation, it is also used in new production. The surfacing material may be cemented carbides, nonferrous Stellite-type alloys, or iron-base alloys with alloying additions such as chromium, tungsten, manganese, silicon, nickel, and carbon. While hard facing is most often applied to steel, cast iron and some of the nonferrous alloys such as Monel metal can also be coated. Typical applications are metal-working dies, oil well drilling tools, excavating equipment, shafting, and rolling mill rolls.

HARDNESS. The significance of this term as applied to solids has various interpretations. Commonly, it refers to the resistance of the substance to surface abrasion, so that of two solids, the one that will scratch the other, as diamond scratches glass, is the harder. Again, it may de-note rigidity, or lack of plasticity, or even strength; in some cases a combination of several such properties. The original Mohs' Scale of Hardness is delineated in Table 1 and further described under **Mineralogy.**

TABLE 1. HARDNESS SCALES

Moh's Scale	Ridgway's Extension of Mohs' Scale	Metal Equivalent	Others
1. Talc			
2. Gypsum			
			2.5. Finger Nail
3. Calcite			
4. Fluorite			
5. Apatite			
			5.5. Window Glass
6. Feldspar (Orthoclase)	6. Orthosclase or Periclase		
			6.5. Steel (Knife Blade; File)
7. Quartz	7. Vitreous Pure Silica		
8. Topaz	8. Quartz	8. Stellite	
9. Corundum or Sapphire	9. Garnet		
	10. Topaz		
	11. Fused Zirconia	11. Tantalum Carbide	
	12. Fused Alumina	12. Tungsten Carbide	
	13. Silicon Carbide		
	14. Boron Carbide		
10. Diamond	15. Diamond		

1. In the above scales each abrasive is capable of scratching all others above it in each scale and may be scratched by all abrasives below it.

2. The gap between 9 and 10 in the original Mohs' scale is much greater than that between 1 and 9 in the same scale.

3. Various additional hardness scales have been devised by different investigators; in general, different materials maintain the same order of hardness in all these scales.

In metallurgy and engineering, hardness is determined by methods based on resistance to penetration by an indenter of greater hardness than the material being tested. Aluminum, copper, lead, magnesium, tin, and their alloys, as well as plastics are generally indented by hardened steel balls ranging in size in the various tests from $\frac{1}{16}$ inch to 10 millimeters in diameter. The same methods may be used for soft steels and irons, but for heat-treated steels and all other alloys which develop high hardness special diamond indenters, or in some cases sintered tungsten carbide balls, are used. In all of the technological tests, the indenters are impressed into the test material under carefully regulated loads; thus, the relative size of the resulting indentation becomes a measure of hardness. (See Table 2.) The operating principles of the instruments most widely used in this country follow:

Brinell. The indenter is a 10-millimeter diameter hardened steel ball. A sintered tungsten-carbide ball is also coming into use, especially for testing hard metals. The load applied is generally 500 kilograms for soft metals and 3,000 kilograms for steels and hard metals. Brinell hardness is equal to the load (kilogram) divided by the surface area (square millimeter) of the impression made in the test material. Tables are available for direct conversion to hardness from the diameter of the indentation as measured with a calibrated magnifier after removal of the piece from the testing machine.

Rockwell. Indenter is $\frac{1}{16}$-, $\frac{1}{8}$-, or $\frac{1}{4}$-inch-diameter (1.6, 3.2, or 6.4 millimeter) steel ball or a conical diamond having an apex angle of 120° and a slightly rounded point. The various scales used are designated by letters. Rockwell "B," for example, indicates a 100-kilogram load on a $\frac{1}{16}$-inch (1.6 millimeter) diameter ball. Rockwell "C" indicates a 150-kilogram load on the diamond indenter. Rockwell "30T" designates a load of 30 kilograms on a $\frac{1}{16}$-inch (1.6 millimeter) diameter ball. (An instrument of higher sensitivity known as the Rockwell Superficial Tester is used for loads of 15, 30, and 45 kilograms.) The size of the indentation is measured by a dial gauge as the final depth minus a small preliminary penetration produced by a minor preload of 10-kilograms. The Rockwell hardness values are arbitrary numbers having an inverse relationship to the depth of the indentation.

TABLE 2. TYPICAL HARDNESS VALUES

Material	Brinell 500 kg	Brinell 3000 kg	Rockwell	Vickers 50kg
Aluminum, annealed	23		H 45	25
Magnesium alloy	63		B 21	63
Armco iron	66	73	B 31	71
Yellow brass, annealed	72	82	B 40	77
Copper, cold rolled	99	83	B 55	110
Mild steel, annealed	107	117	B 70	123
Aluminum alloy, 24st	130	144	B 78	146
Stainless steel, annealed	121	145	B 80	153
Yellow brass, cold rolled	174	178	B 91	189
Ni-Moly steel, quenched in water, tempered at 1200°F (649°C)		241	C 23	255
Same, 1000°F (538°C)		293	C 31	310
Same, 800°F (427°C)		363	C 38	380
High-speed tool steel		684	C 62	740

Vickers. Also known as Diamond Pyramid Hardness. Indenter is a square-based diamond pyramid with included angle between faces of 136°. Loads may vary from 1 to 120 kilograms with 10, 30, and 50 kilograms in common use. Hardness is equal to load (kilograms) divided by surface area (square millimeter) of the permanent indentation. It is determined directly from optical measurements of the diagonals of the indentation which appears square at the surface of the metal.

Tukon. A highly sensitive instrument for determining hardness under very light loads down to 25 grams. The small indentations are measured at high magnifications up to 1,000 times. The indenter is a diamond pyramid that makes an elongated impression, one diagonal being 7 times the other in length.

Eberbach. Also used for very light loads. Consists of a spring-loaded, Vickers-type diamond pyramid indenter arranged for use on a metallurgical microscope.

Scleroscope. Depends on the height of rebound of a diamond-tipped body falling under the force of gravity from a fixed height. The instrument is relatively small and is portable. One type reads directly on a graduated dial.

While there is overlapping in the field of useful application of the various hardness tests, each has certain special qualifications. The Brinell test makes a large indentation, giving an average hardness value for several grains even in rather coarse-grained metals; however, it cannot be used on small or thin specimens. The various Rockwell tests are widely used, especially for rapid production inspection of parts. The Vickers test, which originated in England, is less rapid than the Rockwell but has the advantage of a single scale covering the hardness of all metals from lead to the hardest tool materials. The Tukon test makes it possible to determine the hardness of very thin sheets and of thin metallic coatings such as chromium plate, or zinc on galvanized steel. The Scleroscope test is used principally on heavy forgings or castings which cannot be placed in an indentation-type instrument, or for field tests where a portable instrument is required.

HARDNESS (Mineral). See **Mineralogy.**

HARDNESS (Water). See **Feedwater (Boiler).**

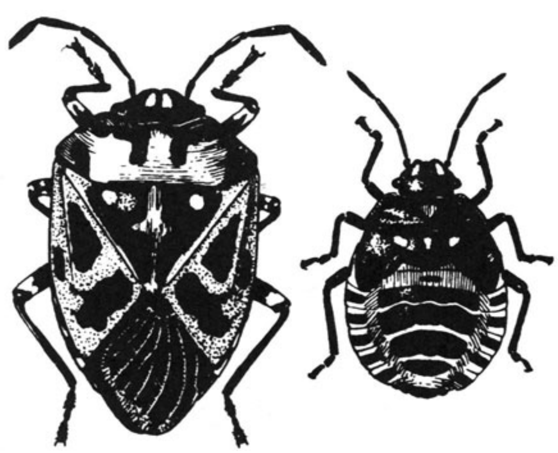
HARDPAN. The term which prospectors and miners give to the subsurface or basal layers of placer deposits in which the gold-bearing gravels have been cemented and hardened. The same term is also used to designate till or boulder clay which has been cemented by limonite.

HARDWOODS. See **Wood.**

HARE. See **Rabbits and Hares.**

HARELIP. A congenital deformity in which there is a failure of fusion of the maxillary and median nasal processes, resulting in a cleft in the upper lip. This is part of the same defect associated with cleft palate. In some cases, the harelip may be double, in which case there is a division on either side of the mid-line of the lip. Correction of the deformity must be by surgery, and this is best accomplished at a very early age. Usually, correction of the lip is done first, thus enabling the infant to suck. Surgery on the palate normally is undertaken just as soon as there is sufficient tissue to cover over the bony palate after repair. This usually occurs at the age of 18 to 24 months, well before abnormal speech habits are formed. In some cases, the operation must be performed in several stages. Cosmetically and functionally, the surgical results usually are excellent.

HARLEQUIN BUG (*Insecta, Hemiptera*). Of the family *Pentatomidae* (stink bug), also known as the "fire bug" and the "calico black" bug, this insect has a major economic impact on cabbage and cruciferous crops in the southern United States. When not controlled, an entire crop can be destroyed. The bug kills plants by sucking sap from the underground portions of the plant. In addition to cabbage, the harlequin is injurious to Brussels sprouts, collard, cauliflower, horseradish, kohlrabi, mustard, radish, and turnip. If these preferred sources of nourishment are not immediately available to the insect, it will attack asparagus, bean, eggplant, okra, potato, and tomato. The insect also can damage numerous other garden crops, certain weeds, small fruit trees, and field crops.



Harlequin bug. Adult at left; nymph at right. (*USDA*.)

The harlequin bug is colorful, with red and black spots, of a shield shape, and about $\frac{3}{8}$ inch (9 to 10 millimeters) long. The nymphs, somewhat smaller, have a similar appearance. All stages of the insect may be found from the early to the late months of the year. In the United States, the pest is found in the southern portion in all areas from the Atlantic to the Pacific coasts. The insect is believed to be native to Mexico. In more northern areas, the bug winters as an adult. During the first warm spell of spring, eggs are deposited on the underside of leaves. They appear something like tiny white beer kegs, standing on end and usually in a double row. There are two black bands around each "keg." Hatching of the eggs occurs within 1 to 4 weeks, depending upon temperature. Immediately, the nymphs commence feeding and destroying target plants. During a period ranging from 4 to 9 weeks, the insect passes through five instars, after which it is ready to mate and lay eggs for the next generation. Normally there are three to four generations per year.

In addition to control chemicals, populations can be controlled by destroying weeds that attract the insects. These include *Amaranthus* and wild mustard. Advantage of the insect's preference for certain crops, such as kale, mustard, radish, and turnip, can be taken by planting a small area to such plants either very early in the season or after harvest. Large concentrations of the insects thus can be killed with relatively small amounts of insecticide. The debris from such decoy crops should be burned.

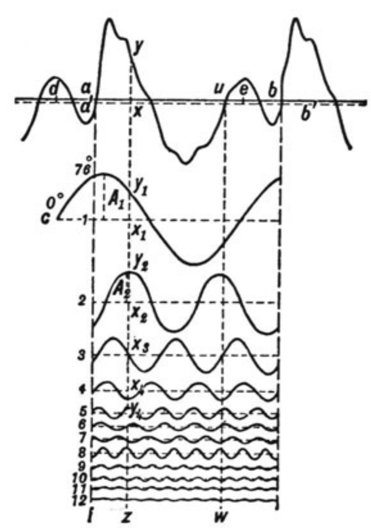
HARMONIC. A sinusoidal frequency component of a waveform. The harmonic has a frequency that is an integral multiple of the fundamental frequency. The frequency of the second harmonic will be double that of the fundamental frequency (first harmonic).

Harmonic distortion is nonlinear distortion characterized by the appearance in the output of harmonics other than the fundamental component when the input wave is sinusoidal. Harmonic distortion is sometimes called amplitude distortion.

HARMONIC ANALYSIS. Not only is it possible to combine two or more simple harmonic motions of different period, amplitude, and phase to form a complex motion, but there are also means of analyzing the resultant motion, when the latter is given, to find its component harmonics. For example, if the wave form of such a complex tone as that produced by a bell or a saxophone is accurately graphed by means of a phonodeik the equation of the vibratory motion can be deduced in such form as to show the separate components. Fourier showed that the same analysis is possible for any periodic motion, however complicated. The equation, called Fourier's series, may be written

$$y = a \sin 2\pi nt + b \cos 2\pi nt + c \sin 4\pi nt + d \cos 4\pi nt + e \sin 6\pi nt + f \cos 6\pi nt + \cdots$$

in which y is the displacement of the vibrating particle and t is the time. The fundamental frequency n and the constants $a, b, c, d,$ etc., must be calculated from the given wave form or the data from which it is plotted. There is a type of instrument, called a "harmonic analyzer," which automatically computes the coefficients; or it may be done mathematically, though the process is very laborious. The accompanying figure shows the wave form and the twelve components of complex tone, analyzed by Professor D. C. Miller.



Records of a complex sound and twelve of its components.

HARMONIC MEAN. See **Average.**

HARMONIC MOTION. A distinct type of periodic motion, or vibration, characteristic of elastic bodies; illustrated by a bird-cage bobbing up and down at the end of a spiral spring, or (approximately) by the piston of the steam engine. It may be either simple, with only one frequency and amplitude, or made up of two or more simple components and consequently of more complex character. The essential feature of simple harmonic motion is that, with its range extending to equal distances on both sides of an equilibrium position or origin, the acceleration is always toward the origin and directly proportional to the distance from it. With elastic vibrations this is easily seen to follow from Hooke's law, since the force tending to restore the deformed body to equilibrium is proportional to the deformation. See **Elasticity.** The motion is called "harmonic" undoubtedly because the vibrations of bodies emitting musical sounds are of this character. Any simple harmonic motion may be represented by the equation.

$$y = a\cos(2\pi nt + \phi)$$

in which y is the distance at time, t, a is the amplitude, n is the frequency or number of vibrations per unit time, and ϕ is the phase constant, such that when $t = 0$, $y = a\cos\phi$.

It is interesting to note the relationship between harmonic and circular motion. If a peg is inserted in the face of a circular disk or wheel and the latter uniformly rotated, the motion of the peg, as viewed with the wheel seen edgewise, is simple harmonic. In fact, uniform circular motion is made up of two simple harmonic components of the same period and amplitude at right angles, one being a quarter-period ahead of the other in phase. If the two harmonic components have a phase difference other than a quarter-period, the resultant in general is motion in an ellipse; while if they have unequal periods, the path is one of a class of more or less complicated loci called "Lissajous' curves."

HARMONIC OPERATION. *Impeded* harmonic operation is constrained magnetization or forced magnetization. It is the type of operation which takes place in a magnetic amplifier in which the impedance of the control circuit and any circuit closely coupled to it is so great as to substantially prevent the flow of all harmonic currents in such circuits.

Unimpeded harmonic operation is natural magnetization or free magnetization. It is the type of operation that takes place in a magnetic amplifier in which the impedance of the control circuit or any circuit closely coupled to it is so small as to permit substantially unimpeded flow of all harmonic currents in such circuit.

See also **Amplifier.**

HARMONIC PROGRESSION. See **Progression.**

HARMONIC SYNTHESIZER. A machine which combines elementary harmonic constituents into a single periodic function. A machine performing the opposite function is called a harmonic analyzer.

HARMOTOME. The mineral harmotome is a zeolite, composition approximately (Ba, K)(Al, Si)$_2$Si$_6$O$_{16}\cdot$6H$_2$O; it is monoclinic but often forms double twins giving the effect of a square prism. It is a brittle mineral; hardness, 4.5; specific gravity, 2.41–2.50; luster, vitreous; color, white to gray or perhaps yellow, red or brown; white streak; translucent. Harmotome like other zeolites is found in cavities in basalts and similar rocks, sometimes in trachytes or in gneisses, occasionally as a gangue mineral in veins of metallic minerals. Some well-known localities are in Bavaria; the Harz Mountains; Norway; and Scotland. Harmotome occurs in the United States with stilbite, near Port Arthur, Lake Superior. The name harmotome comes from the Greek meaning joint and to cut, referring to the division of the pyramid formed by the prismatic faces of the mineral when in the twinned position.

HARPY. See **Eagle.**

HARRIER. See **Eagle.**

HARTEBEEST. See **Antelope.**

HARTLEY. In information theory, a unit of logarithmic measures of information equal to the decision content of a set of ten mutually exclusive events expressed by the logarithm with the base ten. For example, the decision content of a set of 8 characters equals $\log_{10} 8$, or 0.903 Hartley. Synonymous with information content decimal unit. (*American National Dictionary for Information Processing*.) See also **Shannon.**

HARTLEY OSCILLATOR. See **Oscillator.**

HARTLEY PRINCIPLES (Transmission). The amount of information that can be transmitted is proportional to the width of the frequency range, and the time it is available. Information content is equated to the total number of code elements, multiplied by the logarithm of the number of possible values a code element may assume. Information content is independent of how the code elements are grouped. By quantizing, the continuous magnitude-time function used in ordinary telephony may be transmitted by a succession of code symbols such as are employed in telegraphy. To obtain the maximum rate of transmission of information, the signal elements need to be spaced uniformly.

Time-Frequency Duality. As implied by the Fourier integral, a time function cannot be confined within a small region on the time scale when the steady-state transmission characteristic is confined to a narrow range on the time scale. For example, it is well known that, if a telegraph dot is made narrower and narrower, its corresponding significant-frequency spectrum becomes broader and broader until, in the limit when the dot becomes an impluse, its significant-frequency spectrum is of infinite intent.

HARTMANN TEST. Hartmann devised various optical tests, including the following: (1) Hartmann test for telescope mirrors. For a perfect mirror, light from all points on the mirror should come to the same focus. By covering the mirror with a screen, in which regularly spaced holes have been cut, and then permitting the reflected light to strike a photographic plate placed near the focus, the failure of dots on the plate to be regularly spaced indicates a fault of the mirror. (2) Hartmann test for spectrometers. Light is passed through different parts of the entrance slit. Any change in the spectrum as different parts of the slit are used indicates a fault of the instrument. A "Hartmann diaphragm" is one device for using only one part of the entrance slit at a time.

HARTREE-FOCK APPROXIMATION. Also called Hartree-Fock-Slater approximation. A method for the solution of a many electron problem, e.g., that which arises in considering the band theory of solids or an atom with more than one electron. The antisymmetric wave function for the N-electron system is expanded as a linear combination of determinants of order N, having as elements one electron wave functions. This procedure introduces exchange terms in the Hamiltonian, of the form:

$$e^2 \int \left[\frac{\psi_i(r_1)\psi_j^*(r_2)}{r_{12}} d\tau_2 \right] \psi_j(r_1)$$

where r_{12} is the separation of the points defined by the vectors r_1 and r_2.

HARVESTMAN (*Arachnida, Phalangida*). Spider-like animals, most species with small oval bodies and extremely long slender legs. Those with shorter legs are more easily confused with the true spiders but all may be recognized by the segmented abdomen. Daddy longlegs.

HASTELLOY. See **Nickel.**

HATCHET FISHES (*Osteichthyes*). Of the order *Isospondyli*, family *Sternoptychidae*, hatchet fishes are small, rarely exceeding $3\frac{1}{2}$ inches (9 centimeters) in length. They are silvery and are so named because of their hatchet-head appearance. They possess photophores (light organs) on their sides and undersurfaces. The genus *Argyroplecus* features telescopic eyes which are aimed in an upward direction. They are a food source for tuna, but are not nearly so abundant as their relatives, the bristlemouths. Some species of hatchet fishes have been favorites among tropical-fish fanciers.

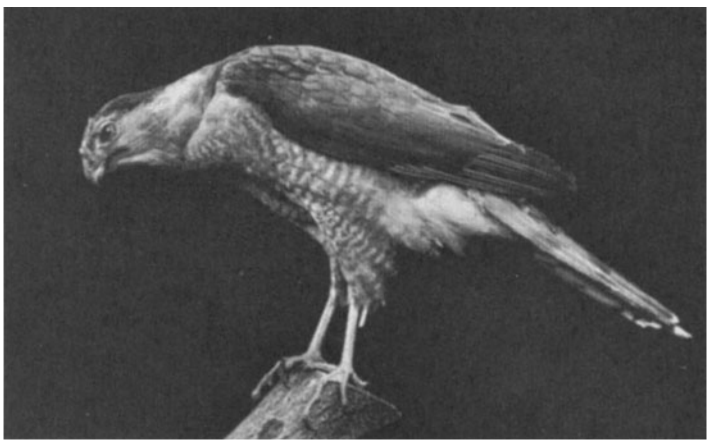
The flying hatchet fishes of South America of the order *Ostariophysi*, family *Characidae* are the only fishes credited with performing true flight. See also **Characids (Osteichthyes).**

Hatchet fishes are fully adapted to a life near the water surface. Like speedboats, the front of which rises off the water at high speed, these fishes can also rise off the water surface. They literally fly several yards (meters) through the air, after a starting movement of a few yards (meters). The initiation of the movement has been seen in nature, but the actual flying is difficult to observe. Laboratory investigations, however, indicate that the process is accompanied by a humming sound.

HAUSDORFF SPACE. See **Topological Space.**

HAVERSINE. See **Trigonometric Function.**

HAWK (*Aves, Falconiformes*). Birds of prey with hooked beaks and large curved claws, closely related to the eagles, falcons, harriers, and others and not sharply distinguished as a group. Hawks are found on all continents. North America has many species, including buzzards, harriers, goshawks and other forms. Most of them are beneficial as destroyers of vermin but the sharp-shinned (*Accipiter velox*), and Cooper (*A. cooperi*) hawks destroy too many birds, including poultry, to be regarded as friends.



Cooper's hawk. (*American Museum of Natural History.*)

There are numerous species of hawks, at least 25 of these occurring in North America, particularly north of Mexico. The hawks are swift in flight, seek their prey by day, have remarkable vision, and eat only what they kill. They are very bold, pouncing upon their prey in a rapid swoop, using claws and talons, firmly clinching the victim. The hawk prefers to take its victim to a private location for consumption.

In the African rain forest, three species are known as darters.

The osprey is a large bird of prey of almost worldwide distribution. It is a skillful fisher and is known in North America as the fish hawk, *Pandion haliaëtus*.

Buzzards are birds of prey of several species that belong to the genus *Buteo*. The North American representatives are commonly called hawks, as Swainson's hawk. The same may be said of the nearly related rough-legged buzzards; American representatives of the genus are the rough-legged hawks. The name is incorrectly, although commonly, applied to the turkey buzzard, which is a vulture. See also **Falconiformes.**

Size Dimorphism. It has been known since medieval times among hawk and falcon fanciers that the male hawk is considerably smaller than the female, whereas the norm for birds in general is the reverse. In a 1985 paper, E. Temeles (University of California, Davis) points out that in addition to size dimorphism among these birds of prey, there is a correlation between the type of diet and degree of dimorphism—the degree of size difference between females and their mates increasing as diet moves through carrion, insects, fish, and mammals to birds. Could it be postulated that the faster the prey moves, the greater will be the size dimorphism in the pursuers? I. Newton (Institute of Terrestrial Ecology, Monks Wood, England) observes that the striking link between diet and size dimorphism possibly may be misleading. Because so much of the biology of birds correlates with their diet in some way, there are many factors that will also correlate with size dimorphism, but not be the cause of it. Mueller (University of North Carolina), in reviewing the various size dimorphism hypotheses, has suggested that a behavioral explanation may be the most reasonable, namely, the female dominance hypothesis. First advanced in the 1960s by T. Cade (Cornell University) and later by S. Smith (Mount Holyoke College), the concept involves some form of protection mechanism for the females, which enhances pair-bonding and pair-bond maintenance.

In nonpredatory birds, where the male usually is physically and behaviorally dominant to the female, courtship often involves something of a role reversal, where the male offers food and essentially amuses the female in his efforts to attract her. Some authorities argue that because the predatory bird is equipped with dangerous talons, beaks, and killer instincts, such social interactions during courtship would be potentially hazardous to the females. Thus, the larger, better equipped female could better protect herself during contact with the male. It is observed that if female predatory birds chose mates smaller than themselves, then the reverse dimorphism among them would emerge. Others suggest that as small males are better equipped for hunting, then female choice for that skill would produce the observed pattern of size dimorphism. Or, if small body size were important in aerial competition for females, this also could produce the size relationships as observed.

Temeles, in proposing a prey vulnerability hypothesis, pointed out that the greater agility of prey not only narrows the potential size range that a predator can exploit, but it also reduces the likely success during each hunting exploit. In essence, this determines the amount of energy captured per amount of energy put out in hunting. In field tests with raptors, Temeles did indeed find that hunting success varied according to the nature of the prey—in the following approximate order: invertebrates, 82%; fish, 58%; mammals, 23%; and birds, 13%.

It can be further postulated that the larger female may be better suited for capturing some prey, the male for others. Perhaps the size dimorphism is not a matter of competition between sexes, but rather a balancing of advantages in insuring the success of the species as a whole. It is expected that many years will be required to produce a reasonably complete answer for the size dimorphism question.

HAWK MOTH (*Insecta, Lepidoptera*). Large moths composing the family *Sphingidae*, one of the largest of the order. These moths have a long, rather stout body projecting beyond the narrow wings. The front wings are much longer than the hinder pair, and because of their limited surface they are vibrated rapidly in flight. The moths have long tongues and visit deep-throated flowers. From their habit of hovering as they probe the flower for nectar they are also called hummingbird moths. Another common name is sphinx moth.

HAWTHORN TREES AND SHRUBS. See **Rose Family.**

HAY BRIDGE. See **Bridge Circuits (Electrical).**

HAY FEVER. See **Allergy.**

HAZE. See **Precipitation and Hydrometeors.**

HAZELNUT SHRUBS. Of the family *Corylaceae*, genus *Corylus*, hazelnut shrubs (rarely trees) are deciduous, and characterized by male catkins that hang from the tree during most of the winter months, and by their edible and tasty fruit, an ovoid nut in a toothed container, simply known as the hazelnut of commerce. The term *filbert* is generally reserved for use with reference to the fruits of two European hazelnut plants, *Corylus avellana pontica* and *C. maxima*. The American hazelnut (*C. americana*) is a shrub ranging from 3 to 8 feet (0.9 to 2.4 meters) in height. It is commonly found in thickets and hedgerows. The leaves are narrow, heart-shaped or sometimes ovate, with abrupt points. They are of a lackluster dark green color and from 3 to 5 inches (7.6 to 12.7 centimeters) in length. The stems are short. The staminate catkins are from 3 to 4 inches (7.6 to 10.1 centimeters) in length. This shrub ranges from Maine westward to Alberta and Kansas and southward to Florida. The record American hazelnut growing in the United States and selected in 1989 is located in Mississippi. As compiled by the American Forestry Association, this specimen has a circumference (at 4.5 feet (1.4 m) above ground level) of 12 inches (30.5 cm) a height of 34 feet (10.4 m), and a spread of 24 feet (7.3 m). The beaked hazelnut (*C. rostrata*) ranges from 3 to 8 feet (0.9 to 2.4 meters) in height and commonly occurs along the road in thickets, ranging throughout Canada from Quebec westward to the Pacific slopes and south into the United States to Missouri, Michigan, and Ohio, and Delaware in the east. It is found in the mountains as far south as northern Georgia. The fruit is edible and sweet and in the form of an ovoid nut. The nut is enclosed in a bristly cup which has a beak-like termination, hence the name.

The California hazelnut (*C. cornuta* or *californica*) is well known for its velvety leaves and makes an attractive garden shrub. For purple coloration in gardens, the purple hazel (*C. maxima purpurea*) is sometimes used. The Turkish hazel (*C. colurna*) can be classified as a tree, in that it can attain a height up to 75 feet (22.5 meters), but in most respects it is similar to the lesser hazelnut shrubs. Another species is the corkscrew hazel or Harry Lauder's walking stick, a shrub which attains a height up to 10 feet (3 meters). It makes an attractive shrub, particularly in winter months when the catkins are on display.

Fig. 1. Filbert ochard in a valley of the Pacific Northwest (United States). (*Oregon Filbert Commission.*)

The record hazelnut (California hazelnut, *Corylus cornuta var. californica*) growing in the United States and selected in 1984 is located in Seattle, Washington. As compliled by the American Forestry Association, this specimen has a circumference (at 4 $\frac{1}{2}$ feet; 1.4 meter above ground level) of 22 in. (56 cm), a height of 47 feet (14.3 meters), and a spread of 42 feet (12.8 meters).

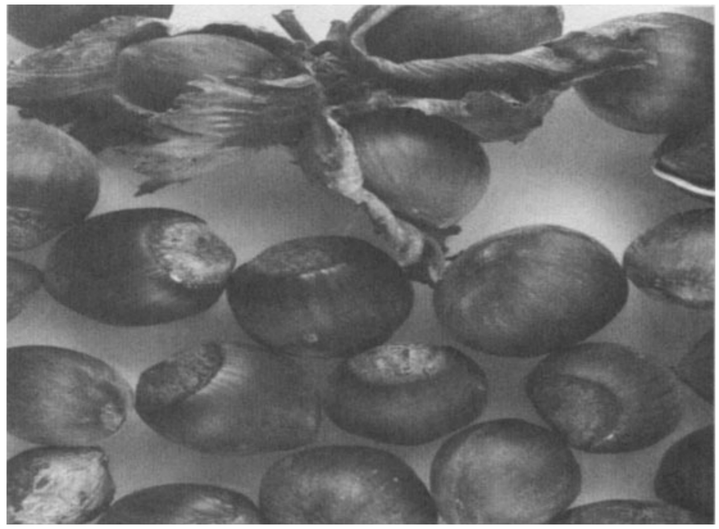


Fig. 2. Filberts (hazelnuts). Upper left, note nuts in husk; upper right, shell opened to expose kernel. (*U.S. Dept. of Agriculture photo.*)

The familiar filbert (hazelnut) of commerce is from transplantation of European varieties several decades (1860) ago, mainly in Oregon and Washington, as well as in a few other western parts of the United States. European species now growing in these areas are *Corylus avellana* and *C. maxima*. Varieties of these include Barcelona, Daviana, and Du Chilly. A filbert orchard in a valley of the Pacific Northwest is shown in Fig. 1. A grouping of filberts is shown in Fig. 2.

HEADACHE. Head pain is a symptom and not a disease. Headache is one of the most common symptoms of a disorder, not only of the nervous system, but of other parts of the body as well. Consequently, discovery of the primary cause of headache is often difficult. The degree of pain associated with headache does not necessarily correlate with seriousness of a cause, a violent headache sometimes being associated with a relatively minor injury. Diagnosis of headache complaint can be facilitated by providing accurate information to the physician—events occurring before the headache, such as emotional stress, exertion, eating, and so on; the time of day or night when headache usually occurs; and other symptoms that may accompany headache, such as nausea, flashes of light, ringing in the ears, rapid or slow onset of the headache, as well as how the headache usually ceases.

The large veins (venous sinuses) and their tributaries that drain the surface of the brain are sensitive to pain, as are the arteries. The brain substance itself apparently is not sensitive to pain, but the coverings of the brain are. The sinuses, teeth, ears, and muscles in the area of the head may be affected so that pain from them, at first local, later covers a wider area.

At least eight pain mechanisms have been identified as causative factors in headache: (1) dilation of the cranial arteries; (2) pulling or traction upon pain-sensitive intracranial structures; (3) traction on and dilation of intracranial blood vessels; (4) inflammation of structures within the skull; (5) contraction of skeletal muscles over the head and neck; (6) spread of pain from stimulation elsewhere in the head; (7) pain from allergenic reaction; and (8) mentally-produced (*psychogenic*) pain. The majority of headaches for which medical attention is sought arise either from dilation of the cranial arteries or contraction of the muscles of the head and neck, or by combinations of these factors. Fortunately, headaches of this type arise from conditions that usually are easy to correct.

Vascular Headache

This term is applied to a type of headache caused by dilation of the cranial arteries. It is associated with general infections, migraine headaches, or those resulting from taking certain drugs; and is largely responsible for so-called hunger and hangover headaches. The headaches of suddenly increased blood pressure are in this group, as well as headaches which follow convulsive seizures or head injury. Headaches of this type usually have a throbbing quality, but this may not be present if the headache is prolonged.

Treatment of vascular headache is generally directed to the underlying cause. The inhalation of high concentrations of oxygen are particularly helpful to persons whose headaches are caused by lack of oxygen. Headaches caused by traction or pressure on intracranial structures are associated with expanding intracranial masses, with brain tumors, abscesses, and hematomas, as examples. Such headaches are aggravated by coughing or straining and are not relieved by drugs which constrict the arteries. Headache associated with brain tumor may be intermittent and mild to moderate in severity and usually does not interfere with sleep.

The headache produced by a hematoma (swelling or tumor filled with blood) is dull, steady, and felt throughout the head. The pain from brain abscess is similar to that of tumor. However, the abscess must be of sufficient size to cause traction before pain is felt.

Headaches caused by traction upon and dilation of the intracranial vessels are typified by the headache which frequently follows lumbar spinal puncture. Despite precautions, at times there may be slow leakage of the spinal fluid through the hole made by the needle. This results in headaches which are ordinarily mild, but can be severe. Once the headache develops, bed rest is about all that is required. The condition heals spontaneously.

Headaches resulting from inflammation of cranial structures are experienced if the patient has any infection within the skull, such as meningitis or encephalitis. Such a headache also occurs as a result of the inflammation that follows brain hemorrhages. These headaches may be intense and require narcotics.

Migraine Headaches

Headaches of this type have been reported since ancient times. Migraine has been termed one of the most common complaints of civilized people. The onset of migraine headaches usually occurs between the ages of 12 to 25, but they can begin at any age. Persons who perform mental work are more likely to be affected than blue collar workers. Also, urban dwellers seem to be more affected than people in rural areas. Often, the migraine victim will be an ambitious, hard driving, meticulous, and exceptionally intelligent individual.

Women who suffer attacks of migraine usually do not have any episodes during pregnancy; and the attacks may disappear entirely after the change of life. The disease may disappear in men and women at all ages, but most frequently attacks cease at around 50 years of age, when the elasticity of the blood vessels has diminished, so that the dilation previously described in the etiology of migraine has decreased.

An outstanding feature of migraine headache, thus differentiating it from other types, is that it affects one side of the head. Other distinctions are the periodic recurrence. There is some evidence that migraine may be hereditary. In most instances, the headaches occur about once every two weeks. In women, it may be associated with the menstrual period. Attacks in some persons, however, do not show this regularity, with headaches being separated by months or even years. A migraine headache may last from a few hours to more than a week. In any individual, the characteristic pain, accompanying symptoms, and length of time are usually about the same for each attack. Some sufferers can generally predict such experiences.

The typical migraine headache commences in the temple, eyeball, or forehead, and soon spreads to include either the left- or right-half of the head. The pain may involve the face and neck and sometimes the arms. Sometimes the headache is preceded by disturbances of vision (dullness of vision, blinding flashes of light, sensitivity to light or sound, or dizziness). As the attack begins, the patient may notice a blind spot, that is, several words in a printed sentence may not be seen. This spot, in rare instances, increases in size until vision in one field is fully gone. The patient may regain the ability to see in the later stages of the attack, but he may still be troubled with dazzling white flashes of light. During the attack, the victim's face usually is pale and sallow and the skin may be sweaty and clammy. The arms and legs may feel cool to the patient, even though there may be fever. Nausea and violent vomiting often mark the climax of the attack. After the attack has run its course, if there has been no vomiting, the patient usually feels relaxed and relieved and may be filled with energy and tend to be overactive, although a dull headache may persist for a day or two.

Many migraine sufferers report that attacks seem to occur in relation to periods of let-down or of exhilaration. Many have noted that their headaches commence on weekends, the first day of a holiday, or on days of planned social engagements or travel. Often, on the eve of onset, the victim may be in high spirits, with an unusually increased appetite. However, on the following morning, the victim may arise with a very depressed or melancholic attitude. The victim may become restless, irritable, and confused, with an inability to concentrate on routine tasks, or to make decisions.

Early Studies of Migraine. Many theories over the years have been offered as to the cause of migraine headaches. Such headaches have been associated with distention of the cranial arteries in the scalp as an immediate cause, but the cause of such distensions has not been well understood. Some authorities have attempted to develop a relationship between the personality traits of migraine patients and those having high blood pressure. For some time, an approximate relationship between children of a migraine-prone parent was considered. Some researchers have attributed migraine headaches to food allergies, asthma, eyestrain, and imbalances of the endocrine system.

Traditional Treatment of Migraine. Until quite recently, physicians have utilized the following procedures:

> Generally, the patient should be left alone in a quiet darkened room because most migraine patients are extremely sensitive to light and odors. An ice bag on the head and hot water bottle at the feet may provide some relief from pain. In some patients, sitting in an upright position rather than lying down reduces the intensity of the pain. Ergotamine tartrate, to be prescribed only by a physician, has been found helpful in terminating the headache in many instances, if given at the beginning of an attack. Inhalation of 100% oxygen may alleviate pain. Strong drugs should not be taken for a migraine attack unless prescribed by a physician, for it is too easy for migraine sufferers to develop a drug habit. The victim should make every effort to determine the factors associated with attacks and to avoid wherever possible such factors. Avoidance of fatigue, late hours, strain, and worry tend to reduce frequency or severity of migraine attacks. In most persons, physical or mental tension is often the immediate cause of an attack.

Breakthrough in Understanding and Treatment. Migraine attacks may persist from 4 to 72 hours and sometimes are preceded by transient focal neurologic symptoms. The conventional treatment as just described has been directed at acute symptomatic relief and sometimes involves prophylactic treatment to reduce the frequency of attacks. It is generally agreed, however, that the efficacy of such treatments is rather poor.

A hypothesis was developed in the 1960s that serotonin [a phenolic amine (5-hydroxytryptamine) and powerful vasoconstrictor found in the blood serum] is important in the pathogenesis of migraine. This finding created much attention because it possibly could become the first drug capable of preventing or reducing the intensity and frequency of migraine. Subsequent tests, however, were disappointing, and only recently has the interest in serotonin been revived.

Recently, a new drug, sumatriptan, was introduced and is reported to be effective in the treatment of migraine and cluster headaches. It is reported that this *design drug* was developed for a particular subpopulation of 5-hydroxytryptamine receptors. The question still posed is where are these receptors? Are they vascular or neural?

In the past, many researchers have attributed migraine to vasodilation, but evidence to this effect has been unconvincing. In fact, the exact cause of migraine-induced pain still remains uncertain, but a central pain mechanism is strongly suggested because so many migraine sufferers also note upper- and lower-limb pain concurrent with head pain. This has focused attention on the probable neural component of the illness. This led researchers to tentatively associate migraine (and cluster headache) with abnormal serotoninergic transmission, possibly at different loci. This could be the central factor to the variable symptoms of both migraine and cluster headaches.

In 1991, The Subcutaneous Sumatriptan International Study Group reported on a study of 639 patients with migraine attacks in a randomized, double-blind, placebo-controlled parallel-group clinical trial in which sumatriptan was tested. Conclusions reported: "We conclude that a single 6-mg dose of sumatriptan given subcutaneously is a highly effective, rapid-acting, and well-tolerated treatment for migraine attacks. The administration of a second dose 60 minutes later to patients not responding well to an initial dose affords little additional benefit."

Cluster Headaches. These headaches are characterized by recurrent, unilateral attacks of great intensity and brief duration. They may be accompanied by local signs and symptoms of autonomic nervous system dysfunction. The attacks occur in series lasting weeks or months, thus the name "cluster" headaches. The pain is reported as severe and reaches a maximum within a comparatively short period of time. The cluster headache syndrome occurs with unusual periodicity and, for that reason, sometimes is referred to as the "alarm clock" headache. As explained by Rankin, "The autonomic symptoms are bilateral but are more severe on the same side as the pain. The hypothalamus may be an activation site in this disorder. The posterior hypothalamus contains cells that regulate autonomic functions, and the anterior hypothalamus contains cells (the suprachiasmatic nuclei) that serve as the principal circadian pacemaker in mammals. Activation of both is necessary to explain the symptoms of cluster headache. The pacemaker is modulated by a 5-hydroxytryptamine-mediated (serotoninergic) system." Thus sumatriptan appears to be effective in the treatment of cluster headaches as well as migraines. With this recent knowledge of the biological mechanism, new designer drugs may be developed to treat both migraine and cluster headaches for patients who do not respond fully to sumatriptan. A study group known as The Sumatriptan Cluster Headache Study Group conducted a clinical trial in 1991 (similar to the previously mentioned trial involving migraine sufferers) and concluded: "Sumatriptan is an effective and well-tolerated treatment of acute attacks of cluster headache."

Additional Reading

Diamond, S., Editor: "Migraine Headache Prevention and Management," Marcel Dekker, New York, 1990.

Ekbom, K., et al.: "Treatment of Acute Cluster Headache with Sumatriptan," Department of Neurology, Söder Hospital, Stockholm, Sweden. [Abstract published in *New Eng. J. Med.*, 322 (August 1, 1991).]

Ferrari, M. D., et al.: "Treatment of Migraine Attacks with Sumatriptan," Department of Neurology, University Hospital, Leiden, the Netherlands. [Abstract published in *New Eng. J. Med.*, 316 (August 1, 1991).]

Raskin, N. H.: "Serotonin Receptors and Headache," *New Eng. J. Med.*, 353 (August 1, 1991).

HEADER. Any pipe, conduit, duct, or channel, which acts as a central point of distribution of a fluid flow to several branch lines, is a header.

HEADING. The direction of the forward end of the keel of a ship (either airborne or seaborne) is known as the heading of the ship. Unless a qualifying adjective is used with the term heading, it means direction with reference to true north. Compass heading, or magnetic heading, may be converted to heading by applying the compass corrections.

See also **Compass (Navigation); Course;** and **Navigation.**

HEAD (Hydrostatic). See **Hydrostatic Pressure.**

HEAD WIND. See **Jet Streams.**

HEAD (Zoology). The region of a bilaterally symmetrical animal body lying at the front end in relation to the ordinary direction of locomotion, or, in bipedal vertebrates like humans and some of the birds, at the highest level.

The development of a head is indicated in animals which are without sharply separated body regions, such as the flatworms. This process of cephalization is closely correlated with bilateral symmetry. The portion of a bilateral animal which goes first inevitably is the first to encounter new sources of stimuli, and shows some concentration of sense organs. Usually the chief nerve center, a cerebral ganglion or brain, also develops here. The concentration of sense organs and nervous control in the head remains characteristic of the region throughout the animal kingdom and in most groups is accompanied by the location of the mouth in the head, together with associated structures for securing food.

HEARING AND THE EAR. The role of the sense organ of hearing (the ear) is to code acoustic disturbances into neural signals suitable for transmission to the brain. The study of this process necessarily involves anatomy and physiology of the ear, the nature of auditory pathways and central nervous system activity in hearing, properties of acoustic signals that elicit auditory responses, and observed phenomena of auditory behavior. These aspects serve to define and delineate areas for investigations of hearing.

The truly phenomenal aspects of hearing can be observed in such behavior as localization of sounds, speech perception and particularly the understanding of one voice in the noisy environment of many, and the recognition of acoustic events that only last a few milliseconds. These and other behavioral phenomena remain to be fully accounted for in theories of hearing.[1]

Structure and Function of the Human Ear

What the human ear can do in processing auditory signals has been established for several years and in rather exquisite quantitative detail. How sounds are conducted to the inner ear is relatively well understood, as is the manner in which signals move from the cochlea to the brain along the eighth cranial nerve. The *total hearing process*, however, continues to elude researchers because of the extreme complexity of the ear's transducer, the *cochlea*. About the size of a pea and containing the organ of Corti, the cochlea incorporates well over a million essential moving parts. These are *hair cells* which, with remarkable subtlety, combine mechanical, hydrodynamic, electrical, and biochemical phenomena in their processing and measuring of incoming acoustic signals. They do this with amazing sensitivity and excellent frequent discrimination. Studies of the inner ear in recent years have taken advantage of advanced technology, including scanning electron microscopy. As mentioned by Hudspeth (University of California School of Medicine), a central goal of current auditory research is the elucidation of the cellular and molecular bases for the active process in the organ of Corti. If the present models are correct, the contribution of this active process must occur every few microseconds or tens of microseconds to facilitate high-frequency hearing. To date, these have been demonstrated to occur on a time scale of seconds to minutes, not of microseconds. This is but one gap that hopefully will be ultimately explained in further biophysical studies of the hair cells.

General Structure of the Ear. Traditionally, a description of the ear is based upon three regions: (1) external, (2) middle, and (3) inner ear. From a functional standpoint, however, the ear may be divided into the outer and inner regions, as indicated in the highly schematic diagram (Fig. 1).

External Ear. This includes the auricle (that part of the ear which can be seen) as well as the *auditory* or *ear canal*, which extends to the *tympanic membrane* (eardrum). The outer ear performs the process of transforming acoustic energy into mechanical energy. The space between the auricle and the ear drum is called the *external auditory meatus*. The meatus is an irregularly shaped tube approximately 27 mm long (adult), with a diameter of about 7 mm, and terminated by the tympanic membrane. The ear canal is an acoustic resonator. Frequencies in the

[1]Theories pertaining to the human hearing process date back a century and a half. Various hypotheses have been proposed. In the early years of study, two contradictory concepts were proposed. First, Seebeck (1843) suggested that the pitch of complex tones composed of higher harmonics (integral multiples of the fundamental frequency) corresponds to that of a pure tone whose frequency equals that of the fundamental frequency, and that the pitch does not change, even when the fundamental frequency (missing fundamental) is removed. The perceived pitch is assumed to be related to the temporal structure of the auditory stimulus (*periodicity* or *virtual* pitch). In the second theory, Helmnoltz (1862) suggested a systematic spatial representation of pure tones in the auditory system according to their frequency (*tonotopic* organization). This was supported by later invasive physiological methods. Helmholtz suggested that frequency information is encoded as *place* information and that perceived pitch is related to the *place* of cortical excitation.

In 1956 Licklider attempted to unify the two hypotheses (*place* versus *periodicity* pitch).

The refinement of biomagnetic measurements made it possible to test Licklider's unified hypothesis in human hearing systems. Further theoretical developments are beyond the scope of this article, but are elegantly described in the Pantev reference listed. See also **Acoustics.**

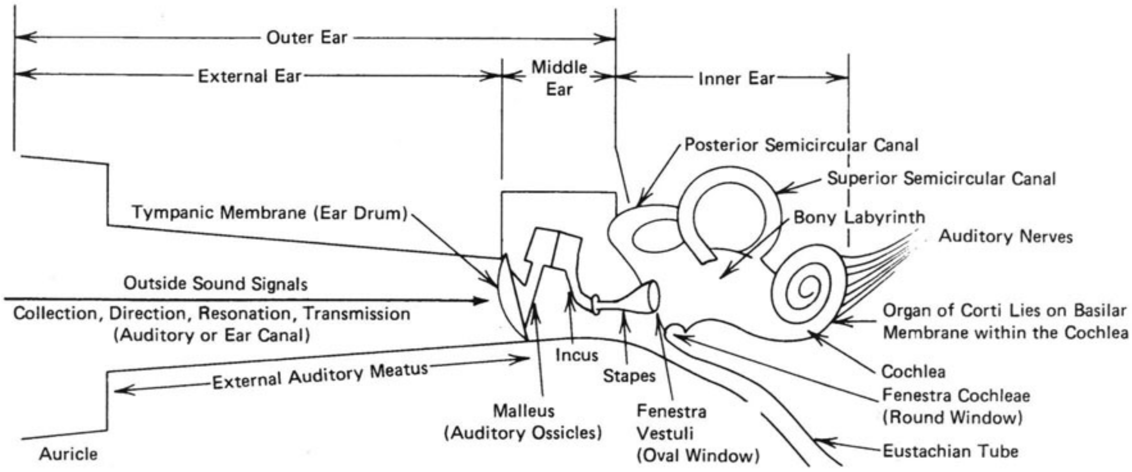


Fig. 1. Highly schematic representation of human auditory system.

range of 3 to 4 thousand Hz are increased in pressure at the eardrum, as compared with the pressure at the entrance to the canal. The eardrum is in a protected position at the end of the canal and thus humidity and temperature conditions at the drum are relatively independent of those external to the ear.

Middle Ear. This is an irregularly shaped, air-filled space in the pestrous portion of the temporal bone. The three auditory ossicles of the middle ear, (a) the *malleus*, (b) the *incus*, and (c) the *stapes*, provide mechanical linkage between the tympanic membrane and the *fenestra vestuli*, an opening in the vestibule of the inner ear, commonly referred to as the oval window. The auditory ossicles are shown greatly enlarged in Fig. 2. The handle of the malleus attaches to the tympanic membrane, and the footplate of the stapes attaches to the oval window. Two important functions are provided by the middle ear. The first is to amplify and deliver sound vibrations from the drum to the inner ear, and the second is that of protecting the inner ear from very loud sounds. The amplification of sound waves is accomplished by apparent lever action of the ossicles that produces a greater force at the oval window than the force at the drum, and because of the gain in force that results from the relationship between the larger drum area to the smaller stapedial footplate area. The area of the drum is approximately 25 times that of the oval window. The amplification gain of these two factors is approximately 25 dB. The effectiveness of the middle ear action in increasing hearing sensitivity is evidenced in middle ear pathologies where the ossicular chain is disrupted. A hearing loss of 25 dB or more occurs. The second function of the middle ear, that of protecting the inner ear from loud sounds, is accomplished by reflex action of the middle ear musculature, the tensor tympani, and the stapedius. The action of the muscles is to retract the eardrum, draw the stapes away from the oval window, and change ossicle vibrations in such a way as to decrease the transmitted pressure. Latency of muscle contraction and possible muscle fatigue limit protection of the inner ear by these mechanisms. Middle ear air pressure is equalized by virture of the Eustachian tube which connects the middle ear and the nasopharynx. The pressure equalization is necessary for normal ear drum movement.

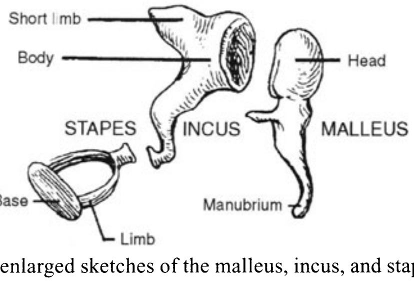


Fig. 2. Greatly enlarged sketches of the malleus, incus, and stapes. (*Anatomy of the Ear, Grace Hewitt.*)

Inner Ear. This is a system of cavities in the dense petrous portion of the temporal bone. One of the cavities is the cochlea, a bony labyrinth that is approximately 35 mm in length, coiled around a central core for $2\frac{3}{4}$ turns.

The Cochlea. Hudspeth, a contemporary researcher in the field, describes the mammalian cochlea as an extraordinarily complex structure that operates in a manner that is fundamentally simple. Nevertheless, the details of the cochlea are still rather poorly understood. Sound made up of a pattern of pressure changes at the eardrum is mechanically conducted through the chain of bones within the middle ear. The stapes, the last of the three bones, is mounted like a piston in contact with fluid within the cochlea. As the stapes moves back and forth in response to stimulation, pressure changes are transmitted into the cochlear fluids. The cochlea is comprised of three fluid-filled chambers, two bony and one membraneous. See Figs. 3 and 4. These chambers are separated from one another by two elastic partitions, which are helically coiled, one top another, about a common axis.

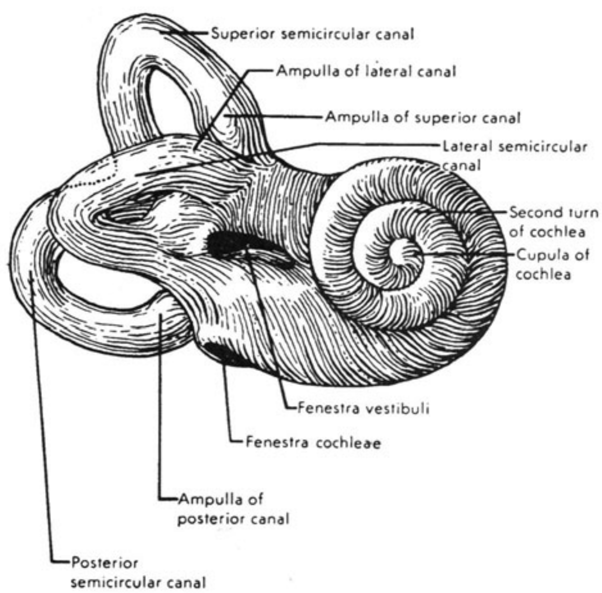


Fig. 3. The bony labyrinth. (*Anatomy of the Ear, Grace Hewitt.*)

When the stapes compresses the fluid within one chamber (basilar membrane), one of the partitions between the cochlear chambers is deflected. It has been found that even when stimulated with a simple sound, such as a pure tone, nevertheless the basilar membrane moves in

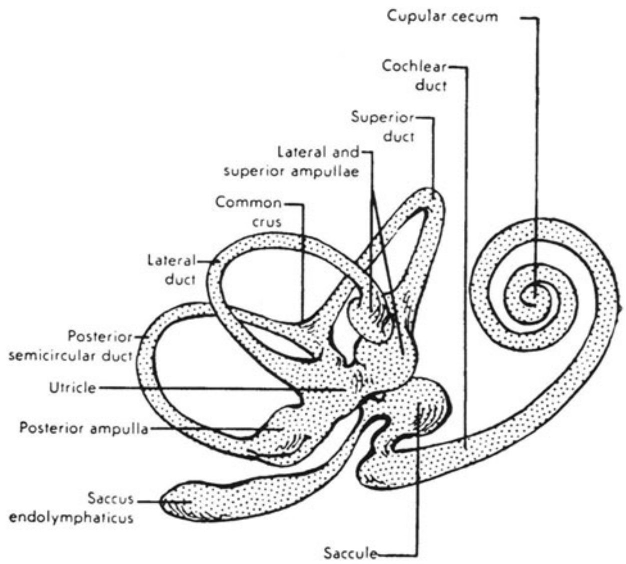


Fig. 4. The membranous labyrinth. (*Anatomy of the Ear, Grace Hewitt.*)

a complex fashion. As explained by Hudspeth, because the dimensions and mechanical properties of the membrane vary from its base to its apex, the membrane does not act like a homogeneous string on a plucked musical instrument. Rather, the basilar membrane develops a traveling wave in a region along its length that depends upon the stimulus frequency. It has been determined that low frequencies (down to 20 Hz in humans) excite motions near the apex of the cochlea. High frequencies (up to 20 kHz in humans) deflect the basal parts of the partition.

For persons with profound sensorineural hearing loss in both ears that cannot be helped by hearing aids, cochlear implants may be useful. Initially announced and approved for use in the United States in 1985, these implants have been used in a comparatively few thousand patients thus far, but estimates indicate that one-quarter million people could benefit from them. A wire electrode that is 1 millimeter in diameter at its widest point is "threaded" into the cochlea to electrically stimulate the auditory nerves. Deafness of this type is caused by damage to the 12,000+ sensory hair cells that line the normal cochlea. Since 1985, the implants have been improved considerably. A 22-channel implant was the first to be approved for use in children in 1990. Tests have shown speech comprehension improvement of from 5 to 160%. In original devices, signals were sent at a maximum rate of 300 times per second. This has been increased to 1000 times per second. The response ranges widely from one individual to the next. As observed by Skinner (Cochlear Implant Program, Washington Univ. School of Medicine, St. Louis), where the stimulation occurs in the cochlea may be just as important as the intensity of the stimulation.

Organ of Corti. This helical structure, which is about 34 mm in length, rests on the basilar membrane in humans. This organ incorporates many thousands of hair cells as well as other types of cells—it has been estimated to have about 16,000 hair cells, in four parallel rows. Each hair cell has about a hundred sterocilia and thus the receptive organelles in each ear exceed 1 million. It is known that each frequency moves a specific zone of the basilar membrane; thus any given tone influences a particular group of hair cells most strongly. Hudspeth observes that one of the cochlea's main virtues ensues from this arrangement, that is, the basilar membrane functions as a *spectral analyzer*, decomposing a complex sound, such as the human voice, into its pure tonal constituents. The hair cells that receive information about a particular tonal input act in some way (still defying an exact description) to transduce mechanical motions of the basilar membrane into electrical signals that are suitable for analysis by the nervous system.

A number of models of the organ of Corti have been constructed in an effort to explain how the hair cells and basilar membrane accomplish transduction so efficiently and within such tight performance parameters of discrimination, accuracy (fidelity), and repeatability.[2]

Neural Responses. The auditory pathways provide for the neural impulses from the ear to be transmitted to the cerebral centers of the auditory cortex. Processing of the neural signals probably occurs at synaptic connections as well as in the cortex. The cell bodies of the receptor neurons are located in the spiral ganglion. Neurons of the auditory nerve make synaptic connections with the hair cells of the cochlea. Nerve fibers typically innervate many hair cells, and more than one nerve fiber may make a connection with the same hair cell. There is recent evidence to indicate that there are also descending neural pathways as well as ascending ones. The central nervous system may thus be involved in auditory processing at the cochlea. Spiral ganglion axons make synaptic connections with cells of the central nervous system at the cochlear nucleus. At this point, there is interconnection between the pathways for the two ears. Other synaptic stations between this point and the auditory cortex include the inferior colliculus and the medial geniculate body. Evidence from pathological auditory systems is of particular interest with respect to the auditory pathways. An impaired cochlea, for example, may result in a better than normal response to small amplitude changes in a sound. A lesion of the eighth cranial nerve is frequently manifested by a rapid decrease in the ability to respond under sustained stimulation. The ability to process speech is markedly affected when there is an involvement of the lower central nervous system. Cortical involvement does not affect usual speech or pure tone inputs.

Diseases and Disorders of the Auditory System

Common earache may arise from many causes and occurs in numerous forms. The most frequent cause of pain, aside from mechanical injuries, arises from some kind of bacterial infection. A physician should be consulted when an earache persists over several hours.

Otitis Media—Acute. An infection of the middle ear. Normally, the middle ear is sterile. The problem occurs rather commonly in early childhood, but the incidence decreases with increasing age. The disease take a number of forms. When bacteria ascend from the nose and throat to the middle ear, the condition is referred to as *purulent otitis media*. The predominant symptoms are pain, fever, and often diminished hearing. Perforation of the tympanic membrane (See Fig. 1) and otorrhea (discharge) may occur. One key to diagnosis is a bulging tympanic membrane with accompanying obscuration of the bony landmarks. The most frequent cause of purulent otitis media is pneumococcus, followed in order by *Haemophilus influenzae*. Anaerobic bacteria, although prominent in the normal flora of the upper respiratory tract, rarely cause acute otitis. Staphylococci are rarely involved. *H. influenzae* is a common cause in young children and is seen in about one-third of the patients between 5 and 9 years of age. A number of other microorganisms can be involved (streptococci, *Neisseria catarrhalis*, and *S. epidermis*, among others).

[2]One end of the hair cells rests on the basilar membrane; the other ends of the hair cells are the cilia, very fine hairlike processes, which make contact with the tectorial membrane, a membrane that overlaps the organ of Corti and that functionally behaves as if it were hinged at the cochlear wall. There are three rows of outer, and one row of inner, hair cells along most of the length of the basilar membrane. When vibrations are introduced into the inner and cause displacement of the basilar membrane, a shearing of the action of the cilia occurs that results in neural activity. It is assumed that amplification occurs in the inner ear in that small pressures on the basilar membrane result in a shearing force of considerably greater magnitude that distorts the hair cells. The result is increased sensitivity of the hearing system. Physical properties of the cochlea are such that different frequencies tend to localize at different points along the basilar membrane. The basilar membrane is narrowest and stiffest at the basal end, and most lax and widest at the apical end of the cochlea. High-frequency sounds result in the greatest disturbances near the basal end, and low-frequency sounds tend to localize near the apical end. When the role of the cochlea in pitch and loudness analyses is considered it is now realized that more is involved in pitch perception than the place of localization on the basilar membrane, although the particular neural fibers involved are probably relevant. Loudness is probably related to the total number of neural impulses per unit time.

In past years, acute suppurative mastoiditis and, less frequently, meningitis sometimes followed acute otitis media. With current antibiotic therapy, these are rare occurrences. Therapy includes pain relievers (analgesics), decongestants, and antibiotics. Ampicillin is frequently the drug of choice for children, and penicillin for adults. Where ampicillin-resistant strains of *H. influenzae* are encountered and where there is allergic response to penicillin, combinations of erythromycin and sulfisoxazole or trimethoprim and sulfamethoxazole are used. The latter drug alone is sometimes used for the chemoprophylaxis of recurrent otitis in children.

Otitis Media—Chronic. This condition usually results from neglected or recurrent acute otitis media and is seen in all age groups. Pain and fever may be absent, with hearing loss and foul discharge being the major symptoms. The tympanic membrane will be perforated. A number of microorganisms (staphylococci, streptococci, *Pseudomonas aeruginosa*, and enteric gram-negative bacilli, among others) may be cultured from the discharge. Antibiotics generally are ineffective. Surgery may be required in advanced cases. In some tropical and developing countries, where clostridia may be introduced with dirty cloths used for removing ear drainage, otogenous tetanus may develop. Otitis media can be a predisposing cause of bacterial meningitis and also may follow a measles infection.

During recent years, there has been considerable interest in the way otitis media occurring during early life may produce lasting developmental impairment. In the past, disorders of speech, language, cognition, and behavior have sometimes been attributed to early otitis media. The mechanism presumed involved deficits in conductive hearing and corresponding "auditory deprivation" during supposedly critical periods of brain development. As pointed out by Hubbard et al., the question has both practical and public health significance because, at one time or other, otitis media affects a large proportion of children. A study reported in 1985 involving nearly 50 children supported the hypothesis that early, long-standing otitis media may result in impairment of hearing and of speech, but no support was found for the hypothesis that cognitive, language, and psychosocial development are adversely affected.

Serous Otitis Media. Also called *secretory otitis*, this condition is characterized by the collection of fluid in the middle ear. This fluid may be either clear (serous) or gluelike (mucous). The predominant symptom is impaired hearing, which varies from a slight to almost total loss. Children who have serous otitis media may be subject to frequent upper respiratory infections and often have enlarged lymphoid tissue in the nasopharynx. If there is an underlying allergy or infection, appropriate antihistamines, antibiotics, or sulfonamides may be administered. Draining the fluid through an incision in the eardrum may relieve the condition. When there are repeated attacks, tiny plastic tubes can be inserted into the middle ear to provide adequate aeration, a procedure that requires a hospital environment. These tubes may be left in place for 3 to 4 months. Many cases of severely impaired hearing in adults can be attributed to middle ear infections in childhood. In infants and children, the Eustachian tube is shorter and more nearly horizontal than in adults, thus making the tube more likely to be an avenue of infection.

Otitis Externa. This disorder originates from the same causes as all middle ear infections, but it differs in the type of inflammation and the changes that occur in the tissues. A head cold may precede the infection. The attack of inflammation is sudden and causes congestion in the linings of the ear spaces, Eustachian tube, and mastoid cells. The ear itself fills with fluid, which gradually becomes puslike. Pain is the main symptom and can be severe, radiating, and throbbing. In children, early symptoms may include refusal to eat, nausea and vomiting, rolling the head, or tugging at the ear. Temperature generally runs high. A ringing sensation and dizziness may be present. Hearing is impaired as long as pus remains in the middle ear. If the condition is left untreated, after several days the eardrum ruptures spontaneously. For as long as three weeks, fluid seeps through the canal and then subsides. The parts of the middle ear are so intricate and delicate that infection spreads easily. Pain resulting from movement of the external ear assists in distinguishing otitis externa from otitis media. Topical therapy includes polymyxin B and neomycin, usually with excellent results. Where true cellulitis of the external ear develops (infrequently), systemic antibiotics and possibly debridement of infected cartilage may be indicated. Very rare neurologic complications of this condition can be life-threatening and require parenteral therapy with tobramycin and carbenicillin, as well as surgical debridement.

Aero-Otitis Media. In this disorder, the structures of the middle ear are affected by changes of pressure which occur during airplane flights. In milder cases, there is a sensation of stuffiness in the ears, with a slight inflammation of the eardrum, and perhaps some minor hearing impairment. Excruciating pain and hemorrhages in the tympanic membrane may occur in more severe cases. Although the condition still may occur among sensitive individuals, pressurized aircraft cabins have greatly alleviated the problem. If one senses this developing, chewing gum or moving the lower jaw with the mouth open will usually prevent it by opening the Eustachian tube which will equalize the pressure. The problem is more common with persons who have upper respiratory infection or severe nasal allergy.

Mastoiditis. The middle ear is generally involved when there is an infection of the mastoid process of the temporal bone. The acute form of this disease (*acute mastoiditis*) has been practically eliminated since antibiotic drugs became available to combat middle ear infections.

The inflammation in mastoiditis involves the lining of the mastoid cells. The infection may enter the bone, which becomes soft and decayed. The causes of mastoiditis include respiratory infection, abnormal anatomy of the ear in infants and children, improper channels for ear drainage, and lowered resistance to infection. Mastoiditis may occur as a secondary infection to various diseases. The predominating symptom is pain, which may be either continuous or intermittent. If the patient is not treated, the intense pain could persist for 6 or more days, which may not be true for middle ear infection. Also unlike middle ear infection, mastoiditis is characterized by a definite, localized tenderness over the mastoid process.

In *chronic mastoiditis*, which now occurs more often than the acute type, drainage from the ear (*otorrhea*) is the principal symptom. Fever may or may not be present. If acute mastoiditis should occur, the physician may perform a *mastoidectomy*. In this operation, the infected mastoid cells are removed through an incision in the area behind the ear, or in the external auditory meatus.

Punctured Eardrum. The most common cause of a punctured eardrum is the insertion of a sharp object into the ear. Violent explosions near the ear may cause the drum to tear or rupture. Decreased air pressure during or after descent from high altitudes, severe sneezing, diving, and increased pressure frequently are responsible for damaged membranes. Sometimes, diagnosis is difficult. The pain accompanying a puncture is sharp and intermittent. Blood may ooze from the injury, but this is not positive proof of a drum tear, because the same symptom may be present in a skull fracture. Dizziness, ringing sounds, and headaches also are significant symptoms. A tear in the eardrum may heal without treatment within a period of a few weeks, but there may be aftereffects which may not be noticed, even for as long as a year. A grafting operation known as *tympanoplasty* can be employed in cases in which the tear does not close.

Growth on the Eardrum. Following rupture or perforation of the eardrum, small chalky (lime) deposits may form at the site of healing as a result of repeated attacks of middle ear infection. If they form from a healed perforation, they mark the path of least resistance for a future rupture. It is the general opinion of physicians that such deposits do not affect normal hearing. There is no successful way of removing the chalk deposits without injuring the eardrum seriously or depressing the hearing. Hence, it is rarely attempted.

Boils or Furnucles. When present in the external ear, these often produce severe pain because the skin in this region normally adheres closely to the underlying cartilage and bone. If infection is allowed to persist, perforations of the eardrum may occur. Through them, infection may spread to the middle ear, the inner ear, or the mastoid area. An x-ray will assist in determining the nature of any secondary complication.

Fungus Infection. *Otomycosis* is a fungus infection of the outer ear and canal. The inside of the ear appears dirty and crusty, and fluid seeps out continually. When the crusts and scales are removed, the skin beneath is raw and bleeds easily. Itching causes much discomfort. Pain is usually present because of the swelling of the canal; hearing may be impaired. Treatment is by specific solutions and ointments. Home remedies are not recommended.

Tinnitus. Most persons, at one time or another, experience this disorder, a sensation of ear noise which is more noticeable in a quiet environment. Such sounds may seem to be in the head rather than the ear, and may affect one or both ears. The symptom is associated with many conditions, including middle ear infection, Ménière's syndrome, exposure to intense noise, circulatory diseases, otosclerosis, and neutritis of the auditory nerve. The symptom also may be caused by excessive amounts of coffee, tobacco, or alcohol. Quinine, certain antibiotics, or large doses of aspirin also may produce tinnitus. Such sounds occur most often in persons between ages 50 and 70. The reason for the sensation has not been established. Inasmuch as the symptom could be an early warning of hearing damage, it should be investigated.

Cauliflower Ear. Known as *hematoma of the auricle*, this disorder has long been recognized as the badge of the prizefighter. It is caused by injury to the external ear. A hard blow may cause bleeding below the skin. If this accumulation of blood remains for sometime, it becomes fibrous tissue and eventually will be converted into a bone-like or cartilaginous substance. Thus, the ear will be deformed by this irregular mass of extra tissue. For prevention, the blood should be removed before it clots. Plastic surgery also is used for restoration of affected ears.

Congenital Malformations. These occur rather frequently, but generally they are not gross enough to impair hearing. They may be unsightly. Absence of the lobe or the outer rim of the ear (*helix*), large protruding ears, and irregular shapes are among the more common malformations. Plastic surgery can restore most of these conditions to normal appearance. Occasionally, a congenital defect, such as an obstruction in the canal, may have to be removed before hearing improves. In rare instances, the ears may be displaced on the head, and in some extreme cases when the lower jaw is grossly misshapen, they may even be fused together (*synotia* or *otocephaly*).

Vestibular Disturbances. The semicircular canals of the inner ear are partially responsible for adjusting the body to changes in motion. The rate of these changes normally allows sufficient time for the canals to maintain body equilibrium. Rapid, irregular, and continuous waves of motions, when they persist over a period of time, may interfere with the vestibular apparatus of the ear and the result is **motion sickness.** This unpleasant condition may be encountered at sea, in the air, while riding in an automobile, on an elevator, etc. The personal reactions to motion sickness are highly individualistic. Recovery is rapid, once the cause is avoided. A number of oral drugs, such as dimenhydrinate (Dramamine®), meclizine, cyclizine, or promethazine, are used as preventive measures—taken an hour prior to boarding a boat, car, etc. Drowsiness may be a side-effect, thus the drugs should not be taken by persons operating automobiles or other vehicles and dangerous machinery.

Ménière's Syndrome. Prosper Ménière described this malady in 1861 and correctly attributed its origin to the inner ear. Its characteristic symptoms are sudden severe episodes of *vertigo* (dizziness), tinnitus, and fluctuating hearing loss. The term syndrome continues to be used because the exact causes of the disorder have not been fully established. Persons in the middle age group are more commonly affected by the syndrome. The vertigo associated with an attack may be so severe that the simplest activities become impossible. Usually, the patient has a sensation that objects are whirling about. The same type of dizziness occurs with certain cardiovascular disorders and middle ear infections. Attacks may last for minutes or weeks. The tinnitus, usually a roaring noise, sometimes persists between attacks. Nausea and vomiting are also usual symptoms.

The course of the syndrome is unpredictable. Remissions of up to several years often occur. About two-thirds of the patients improve or recover regardless of treatment. No single form of therapy has been fully successful. Certain drugs, such as Dramamine®, often help control the vertigo. Sedatives or tranquilizers are occasionally helpful. If the condition is disabling and unilateral, the diseased parts of the labyrinth may be surgically removed. The procedure stops the vertigo, but balance is impaired and hearing loss in the affected ear is total. Ultrasonic radiation has been used to irradiate the labyrinth with the objective of destroying the diseased portions. For relief of severe vertigo, some surgeons recommend the Tack operation to drain the saccule, which contains endolymph. A tack, a small pointed piece of metal, is placed through the footplate into the sac, thus allowing drainage. According to one theory, this syndrome is related to an imbalance of pressure between the perilymph and the endolymph. Another innovation has been the use of surgical instruments which are maintained at temperatures as low as -140°C. With these instruments a surgical procedure should be less likely to damage the cochlea.

Vestibular Neuronitis. This is a comparatively common syndrome, the manifestations of which are vertigo, vomiting, and imbalance. Some authorities believe the disorder results from irritation of the vestibular portion of the eighth cranial nerve. Although vestibular neuronitis resembles Ménière's syndrome in many respects, there are no audiologic symptoms, and in particular no hearing loss. The disease is benign and there is no specific treatment.

The Ear and the Nature of Sound

Sound as a physical phenomenon is described in considerable detail in the article on **Acoustics.** Sound involves a disturbance in the air that is a forward and backward, rarefaction and compression, movement of air parcels. The unit of force usually used in acoustics is the dyne. Sound pressure is frequently expressed in dynes per square centimeter. Intensities of sounds are usually measured on a decibel scale, a logarithmic ratio scale. The tremendous loudness range of the ear is exemplified by the fact that the most intense sound that can be tolerated is a million million times greater in intensity than a sound that is just audible. This is a range of approximately 120 dB. The frequency range of hearing is frequently given as 16 to 20,000 Hz. The ear is most sensitive in the middle frequency range of 1,000 to 6,000 Hz. In terms of discrimination of frequency and intensity, it is possible for about 1,400 pitches and 280 intensity levels to be distinguished.

Hearing Loss and Deafness

Deafness means nearly complete or total loss of hearing. There are two types: (1) congenital, and (2) acquired. In the congenital type, the person is born deaf or later becomes deaf because of an inborn defect. Hard of hearing is a term that applies to those who lose some of the ability to hear later in life, but who have learned how to speak before the loss occurred.

Causes of deafness are many. Some conditions which may cause deafness or milder hearing difficulties include (1) temporary or chronic infections in one or both ears; (2) secondary complications of disease elsewhere in the body; (3) direct damage or defect in some part of the hearing system; (4) aging; (5) occlusion of the auditory canal; (6) aero-otitis media; (7) Ménières syndrome; (8) ostosclerosis; (9) noise; and (10) certain toxic drugs. Side effects of the loop diuretics, ethacrynic acid and furosemide, include transient hearing impairment. Complete deafness has been reported after intravenous administration of ethacrynic acid and a permanent hearing deficit after chronic use.

Conductive deafness results when sound waves are not transmitted properly through the outer and the middle ear. If the damage is to the inner ear or the nerve pathway to the brain, a *sensorineural* (also called *nerve* or *perceptive*) *deafness* occurs. The latter type is generally a greater handicap and usually cannot be reversed. In *mixed hearing loss*, there are elements of both conductive and sensorineural types of loss. Some deafness is caused by a disorder in the central nervous system.

Otosclerosis. Usually first detected during early adulthood, *otosclerosis* can cause a conductive type of hearing loss. Bony growths form just inside the inner ear where the middle ear's stirrup (*stapes*) enters it. Eventually, the footplate of the stapes becomes anchored and no longer conducts sound waves to the inner ear. About 10% of the population is affected to some extent in this way, although they may have no hearing loss for many years. Experience indicates that the disorder may become arrested at any stage. Heredity appears to be an important factor. Middle ear infections are not a cause. The disorder occurs about twice as often in females as in males.

Noise. Individuals vary in terms of susceptibility to noise-induced hearing loss. If sufficiently exposed to intense noise for extended period, all persons are considered as candidates for loss of hearing. Any noise in excess of 85 dB is considered damaging. Frequently, the hearing loss will be accompanied by a high-frequency tinnitus. Noise-induced hearing loss usually is first noted at about 4 kHz, progressively moving into the lower frequencies with continued exposure. The alterations in the inner ear caused by external noise are not well understood. Recovery from the hearing loss is not to be expected. Avoidance of noise or wearing protective devices, such as ear protectors and plugs, is recommended. The best form of prevention is that of taking measures to reduce the amount of noise radiation that escapes from heavy industrial equipment, vehicles, etc.

Even with an increased awareness of the adverse effects of noise, the environment continues to become noisier. Overamplified music and noisy vehicles, particular favorites of young people, have been implicated in the cause of hearing loss. One of the frustrating effects of noise is the masking of speech. For example, if the speaker and listener are separated by 5 feet (1.5 meters), the levels of noise that will barely permit reliable word intelligibility are 50 decibels for normal conversation; 57 dB for raised speech; 63 dB for very loud speech; and 69 dB for shouting. As shown by the accompanying table, these levels are approached or exceeded in several day-to-day industrial and commercial activities.

NOISE LEVELS FOR VARIOUS SOURCES AND LOCATIONS

Description of Noise	Noise Level (dB)
Threshold of hearing	0
Rustle of leaves in gentle breeze	10
Quiet whisper (distance of 5 feet)	10
Average whisper (distance of 4 feet)	20
House in country (average situation)	30
House in city (average situation)	40
Apartment (average situation)	40
Hotel	42
Theater (between performances)	42
Small retail establishment	52
Commercial garage	55
Medium-size office	58
Residential street	58
Restaurant	60+
Medium-size retail establishment	62
Factory or warehouse office	63
Large retail establishment	63
Ordinary conversation (distance of 3 feet)	65
Large office	65
Traffic on busy street	68
Factory (light-to-medium work)	78
Riveter (distance of 35 feet)	97
Hammer blows on steel plate (distance of 2 feet)	114
Threshold of pain	130

Based upon original data by H. F. Olson ("Acoustical Engineering," Van Nostrand Reinhold, New York, 1957).

Instrumental Methods for Measuring Sound and Hearing

The most common measurement of hearing function is the pure-tone audiogram in which a frequency from 125 to 8,000 Hz is plotted against hearing loss in decibels. The audiogram displays the ability of the ear to hear a pure sine-wave tone at a given frequency compared with a "normal" ear. The unit of loudness is the decibel, defined as $10 \times \log_{10}(P_1/P_2)$, where P_1 is the power of the sound being applied and P_2 is the just-audible power required at the given frequency for the "normal" ear to hear. The standard audiometer contains a frequency-selection knob, an attenuator calibrated in 5-dB increments, and a key which connects the output of the instrument to the earphones placed on the subject's head. The procedure is to increase the amplitude slowly while depressing the key in short pulses until the subject reports that the sound can just be detected.

In addition to pure tones, speech sounds are also used as test signals. Using +9 dB (referred to 0.0002 dyne/square centimeter) as a 0-dB threshold level, it is possible to determine the extent of the hearing loss for speech using specially selected two-syllable words having approximately equal stress on each syllable (called "spondaic" words). The equipment used for this measurement consists of a microphone, audio amplifier, and a pair of headsets, the system having a float frequency response between 125 Hz and 8 kHz. Sensitivity, or gain, of the amplifier is controlled by a step attenuator calibrated in 1-dB steps, and the output is arranged to go into either ear separately, or both ears simultaneously.

In the von Békésy pure-tone audiometer, the amplitude control is run up and down by a motor while the subject operates a key. The amplitude is slowly increased until the subject hears the sound, which reverses the motor. The frequency is similarly increased slowly and automatically. The resulting curve is somewhat sawtooth in form and more accurately brackets the threshold values.

In designing and using audiometers, great care must be given to the elimination of background noise and hum. If more than one tone is presented at a time, "masking effects" may occur, giving different results than would be obtained with each sound separately.

Sound-level meters are widely used throughout industry in an effort to stay within legislatively prescribed limitations. Noise-level dosimeters, which automatically compute cumulative noise exposures (for example, the exposure of a worker to noise over an 8-hour workday) are also available. Allowable noise limits in the United States are monitored for compliance by OSHA (Occupation Safety and Health Administration). These limitations are subject to change from time to time as experience is gained.

Hearing Devices

In addition to portable personal hearing aids which have been available for many years, a few researchers are taking a different approach to the problem with the target of developing implantable prostheses for delivering electrical stimuli directly to the auditory nerves. Such devices would be applicable to individuals whose hearing loss is the result of damage to the hair cells of the inner ear. In one design (experimental), an 8-channel, bipolar solid-state device would deliver stimuli at eight different frequencies to separate groups of auditory-nerve fibers in the cochlea. Eight closely spaced pairs of electrical contacts are distributed along the length of the implanted device. The many problems remaining to be solved with such endeavors are well outlined by Loeb (reference listed).

See also **Voice and Sound Production**.

Additional Reading

Borg, E., and S. A. Counter: "The Middle-Ear Muscles," *Sci. Amer.*, 74 (August 1989).

Brownell, W. E., et al.: "Evoked Mechanical Responses of Isolated Cochlear Outer Hair Cells," *Science*, **227**, 194–196 (1985).

Corwin, J. T., and D. A. Cotanche: "Regeneration of Sensory Hair Cells After Acoustic Trauma," *Science*, 1772 (June 24, 1988).

Dooling, R. J., and S. H. Huklse, Editor: "The Comparative Psychology of Audition," Erlbaum, Hillsdale, New Jersey, 1989.

Erickson, D.: "Electronic Earful: Cochlear Implants Sound Better All the Time," *Sci. Amer.*, 132 (November 1990).

Hudspeth, A. J.: "The Cellular Basis of Hearing: The Biophysics of Hair Cells," *Science*, **230**, 745–752 (1985).

Klinke, R., and R. Hartmann, Eds.: "Hearing: Physiological Bases and Psychophysics," Springer-Verlag, New York, 1983.

Kryter, K. D.: "The Effects of Noise on Man," 2nd Ed., Academic Press, Orlando, Florida, 1985.

Licklider, J. C. R.: in *Information Theory* (C. Cherry, Editor), Butterworths, London, 1956.

Pantev, C., Hoke, M., Lütkenhöner, B., and K. Lehnertz: "Tonotopic Organization of the Auditory Cortex: Pitch Versus Frequency Representation," *Science*, 486 (October 27, 1989).

Romani, G. L., Williamson, S. J., and L. Kaufman: *Science*, **216**, 1339, (1983).

Seebeck, A: *Ann. Phys. Chem*, **53**, 417 (1843). (A classic reference.)

Suter, A. H.: "Noise Wars," *Technology Review (MIT)*, 42 (November–December 1990).

von Helmholtz, H.: "Die Lehre von den Tonempfindungen als physiologische Grundlage für die Theorie der Musik," Vieweg, Braunschweig, 1862. (A classic reference.)

HEARING (Fishes). See **Fishes.**

HEARING ORGANS. See **Sensory Organs.**

HEART AND CIRCULATORY SYSTEM (Physiology). The circulatory or cardiovascular system of the human body is comprised of the heart and the blood vessels (arteries, veins, and capillaries). These organs are highly interdependent. The study, diagnosis, and treatment of diseases and disorders of this system fall under the general classification of *cardiovascular medicine*.

Several other articles in this encyclopedia relate to the heart and circulatory system and include:

Aneurysm	**Cerebrovascular Diseases**	**Endocarditis**
Angiography	**Collateral Circulation**	**Hemorrhage**
Anticoagulants	**Congestive Heart Failure**	**Hypertension**
Aorta	**Diastole**	**Hypotension**
Arrhythmias	**Diuretics**	**Ischemic Heart Disease**
Arteries and Veins	**Echocardiography**	**Pulse**
Blood Pressure	**Electrocardiography**	**Sphygmomanometer**

Heart

The heart is the muscular organ that pumps blood through various conduits to and from all parts of the body. Depending upon the size of the adult individual, the human heart weighs somewhat less than three-quarters of a pound (about 340 grams). The organ essentially is a hollow muscle capable of contraction like other muscles. A contraction of the heart is referred to in general terms as a *heartbeat*. The rate of the heartbeats can be changed by two different sets of nerves: (1) The accelerating nerves are connected to the spinal cord and are a part of the sympathetic nervous system; (2) the *vagus nerve* depresses the rate and is connected to the brain stem. The beating of the heart commences long before birth and continues as long as life continues. Beats occur at the rate of 70–80 times per minute in adults, but may increase to 100 beats per minute during exertion, or in the presence of emotional disturbance. During a 70-year life span, it is estimated that the heart beats some 3 billion times, an average of about 42 million beats per year. Each contraction of the heart moves slightly more than 2 fluid ounces (~59 cubic centimeters) out into the arteries, providing a change of blood over the body about once every minute. During a lifetime of 70 years, a total of 250 million quarts (~236.5 million liters) of blood are moved, almost enough to fill a large football stadium. There are only a little over 6 quarts (~5.7 liters) of blood in the average human body, so that this blood requires not only rapid circulation, but also a fine adjustment of controls to assure the proper and effective distribution required by the body.

The highly schematic diagram of the heart given in Fig. 1 indicates the principal components of the heart structure. The heart is divided into four chambers—two auricles, referred to as the right and the left auricle; and two ventricles, referred to as the right and the left ventricle. The flow of blood through these chambers is controlled by four valves, as numbered in the diagram: (1) the tricuspid valve; (2) the mitral valve; (3) the pulmonic valve; and (4) the aortic valve.

Blood coming from over the body through the large veins (venae cavae) enters the right auricle at A. This blood has been partially depleted of its oxygen. As the lower, thick-muscled ventricles expand, this blood enters the right ventricle through the tricuspid valve. Then, the ventricle contracts and forces the blood into the pulmonary artery toward the capillaries of the lungs and is prevented from running back into the heart by the closure of the pulmonic valve. In the meantime, the purified blood in the left auricle has just arrived from the lungs through the pulmonary veins, at B. From here it passes into the thick-walled left ventricle through the mitral valve. When the right ventricle forces blood out into the pulmonary artery, the left ventricle at the same time contracts and sends blood out into the arteries of the body, passing through the aortic valve into the aorta. The auricles thus act as collecting chambers, while the ventricles serve as pumps. The right side of the heart collects the blood and forces it through the lungs; while the left side collects it from the lungs and forces it through the body as a whole. The four valves between the various chambers of the heart prevent the blood from flowing backward and maintain the pressure between heartbeats because of the closed system that results.

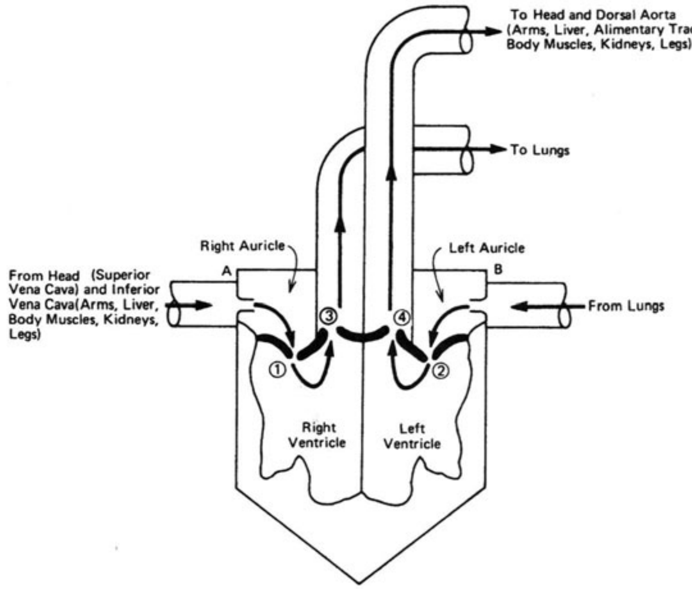


Fig. 1. Highly schematic diagram of major components of human heart: (A) Entrance of blood from venae cavae to right auricle; (B) entrance of blood from lungs; (1) tricuspid valve; (2) mitral valve; (3) pulmonic valve; (4) aortic valve. Diagram is not to scale.

In order that blood can be moved forward in an orderly manner, it is important that the heart muscles expand and contract at just the right time and that all the valves open and close completely at the proper time during the cycle. This control is accomplished by a special structure known as the *sino-auricular node*. This is the pacemaker of the heart. It is not entirely dependent upon the general nervous system, and it has been known to function for some time after breathing has ceased. Sudden changes in temperature, unusual nervous stimuli, fright, a sense of impending danger, or a happy thought can affect this heart center and, thereby cause speeding or slowing of the heart action. All warm-blooded animals have such a fine adjustment that acceleration or retardation may occur within $\frac{1}{100}$th second.

Traditionally, the accepted model of the heart's function was derived mainly from work done near the end of the 19th Century by Otto Frank (Germany) and Ernest H. Starling (England) who postulated that the energy imparted to the blood by the contraction of a ventricle, independent of any control by nerves or hormones, is proportional to the length of the ventricular muscle fibers at the end of the preceding diastole. It was assumed that once systolic contraction was complete, the subsequent diastolic filling becomes a passive function of venous pressure, which stretches the relaxed muscle of the ventricle wall. The Frank-Starling concept is that the energy expended in contraction has no essential role in the diastolic filling of the ventricles.

In 1986, reporting on research of the early 1980s, researchers T. F. Robinson, S. M. Factor, and E. H. Sonnenblick (Albert Einstein College of Medicine, Yeshiva University) describe a new model of the heart, suggesting that some energy from each contraction is stored within the muscle to provide the power for a suction that aids filling. The effect appears to be amplified by the motion of the heart as a whole. They point out that the Frank-Starling model does not reflect the dynamic interplay between systole and diastole. The researchers concede that vast improvements in the instrumentation for accurately measuring the parameters of the heart's performance were not available at the time the original model was proposed. Thus, in the earlier model, the mechanism by which the heart is filled is relatively static. In the new model, the dynamic relation between systole and diastole is critical to the proper function of the heart. The systolic contraction provides much of the energy that drives the process of diastolic expansion. The researchers explain that this energy is stored and recovered in two ways: (1) By the gross motion of the heart itself (when the heart contracts, it propels blood upward and thus, by Newton's law of action and reaction, and thus propels itself downward within the body). Recoil stretches the great elastic vessels and connective tissue that hold the heart in place. Subsequently, as the heart relaxes, it springs upward, thus meeting the inflow of blood head on. Thus, the velocity of the blood with respect to the heart is raised and assists in powering the filling process. (2) The energy of systole is stored in the deformation of the heart itself. In the new model, the systolic contraction compresses the elastic elements of the heart and its muscle fibers so that without any external filling, there is a natural propensity for the ventricles to expand—an expansion that creates a negative pressure (suction) that pulls blood into the ventricles from the atria. In summary, the Frank-Starling model is of a static pressure pump; the new model is of a dynamic suction pump.

Networking in the Heart. The sino-auricular node lies in the wall of the right auricle, embedded within the muscular tissue. A heavy partition extends between the left and right side of the heart, so that there is no direct connection between them except for a group of structures consisting of the auriculo-ventricular node, the *common bundle* and its left and right branches. The auriculo-ventricular node transmits impulses from the common bundle, also known as the *buncle of His*, thence to the two branches, and from there to a network of muscle fibers which covers the inside of each ventricle. The network extends to the outer covering of the heart and is called the *Purkinje system*. This system assures an almost instantaneous response of the muscles of the ventricles once the impulse has passed into it. Although the heartbeat is not entirely independent of the general nervous system, it may carry on for some time without the ordinary nerve impulses. This is illustrated by the fact that the heart of a rabbit, for example, may continue to beat long after the animal has died. This automaticity of the heartbeat allows for cardiac transplantation.

The normal beating of the heart is associated with the production of bioelectric currents in the organ. Although these currents are not strong, they are carried to the surface of the body where they may be measured by a sensitive instrument, the electrocardiograph. The beat of a normal heart shows a characteristic pattern of electrical responses. See also **Electrocardiography.**

There are thousands of small muscle fibers interwoven to make up the walls of the heart. The organ also has its own circulatory system to provide the muscle with nourishment. The whole structure is sheathed with a tough sac, the *pericardium*, containing a small amount of fluid. This provides for lubrication of the rapidly moving heart.

The 70 or 80 normal heartbeats per minute do not allow much time between the expansion and contraction of the four heart chambers. The period of relaxation of the muscles, during which the heart fills, is about equal to that of contraction, when it empties. This period of relaxation permits the heart to recover fully from its work period. The contraction of the heart is called *systole*; the relaxation is called *diastole*.

Valving in Heart and Circulatory System. To make the blood move in only one direction, there not only are the valves inside the heart, but also valves in the veins. In addition, in the small veins there is a constricting type of valve which helps adjust the rate of blood flow and the distribution of blood between the several organs in accordance with need. The capillaries act as the final speed control, by being so small that only one or two rows of blood cells may pass through at a time. Here the speed of the flow is so reduced that time is allowed for rebalancing the mineral content of the area, the exchange of oxygen for carbon dioxide, and soluble food for waste materials.

Heart as a Hormone Source. Since the publication of "Essay on the Motion of the Heart and the Blood in Animals" by William Harvey in 1628, the heart has essentially been considered a *pump*, albeit a complex one. John Peters (Yale University School of Medicine), as early as 1935, speculated about a mechanism that may be located in or near the heart to "sense the fullness of the blood-stream," in essence, some biochemical substance that fine-tunes the regulation of blood volume. Later, during the 1950s and 1960s, some researchers established the overall properties of what they called a "natriuresis" (derived from excretion of sodium and diuresis, excretion of water) hormone. The hormone was also referred to as the "third factor," third because of the other two established regulators of blood pressure and blood volume, namely, (1) the hormone aldosterone and (2) the process of filtering the blood by the kidneys. In the mid-1950s, while studying the constituents of heart-muscle cells, J. D. Jamieson and G. E. Palade (Yale) noted unexplained dense bodies in the cells. In 1974, a research team of M. Cantin and J. Genest (Clinical Research Institute of Montreal) and colleagues noted a similarity between the unknown dense bodies and the storage granules seen in the endocrine (hormone-secreting) cells, such as the pancreas and the anterior pituitary gland. In 1976, Pierre-Yves Hatt (University of Paris) and colleagues, by way of experimentation with laboratory animals, noted a relationship between the number of granules in the atrial cardiocytes (heart-muscle cells) and the amount of sodium in the animal diet. As reported by Cantin and Genest (1986), the breakthrough came in 1981 when A. J. deBold, H. Sonnenberg and associates (Queen's University, Ontario, Canada) injected homogenized rat atria into rats and observed a rapid, massive, and brief diuresis and natriuresis. Thus, they concluded that the atria contained a factor that promotes these effects and named it *atrial natriuretic factor* (ANF). The heart-produced substance is classified as a hormone and subsequently it has been found that ANF exerts its effects on the blood vessels, the kidneys, the adrenal glands, and on a large number of regulatory regions in the brain.

Blood Transporting Vessels

The aggregate length of conduits required to transport blood throughout the human body would be measured in terms of miles or kilometers. The beat of the heart forces a temporarily increased amount of blood into the arteries. The arterial walls are elastic and expand to accommodate this larger volume of blood. Between beats, the walls gradually contract, forcing the blood through the capillaries at an approximately constant rate. In this manner, the arteries act as a reservoir which prevents the blood from flowing through the tissues in gushes. See also **Arteries and Veins.**

Blood passing from the heart through the lungs has only about one-sixth of the pressure of the blood as it is forced out over the body through the *aorta*. See Fig. 2. The pressure is still sufficient, however, to cause flow through the multitude of capillaries in the walls of the lungs. The lungs are composed of innumerable small sacs which have a supply of changing air. In the lung or pulmonary capillaries, the blood releases carbon dioxide and takes up oxygen.

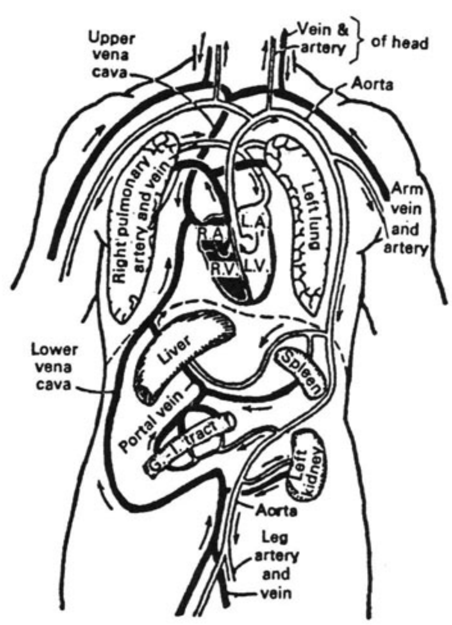


Fig. 2. Highly schematic representation of circulatory system of human body: R.A. = right auricle; L.A. = left auricle; R.V. = right ventricle; L.V. = left ventricle; G.I. = gastrointestinal tract.

The blood continues to flow back through the pulmonary veins and into the left auricle for distribution over the body. The loss of carbon dioxide and the assimilation of oxygen is accompanied by a change of color in the blood, from a dark to a bright red.

Although the liver does not have a special connection with the heart, it acts as a storage organ for blood. Blood is carried to the liver from the stomach and intestinal tract by the portal vein and from the rest of the body by the hepatic artery. It has been estimated that the liver and portal vein drainage system may hold as much as one-third of all the blood in the body. When the body is inactive and requires a smaller amount of blood, the liver and portal vein system relieves the remainder of the system by holding a large part of the excess. Some impurities are removed in the liver and excreted into the digestive tract. The hepatic vein returns the blood from the liver to the larger *vena cava* and heart for distribution over the circulatory system.

The blood supply of the heart itself is by way of special *coronary* arteries. These are necessary to supply the thick heart muscles with the large amounts of food and oxygen necessary for their continuous activity. The walls of the blood vessels themselves contain small canals through which blood is transported to nourish the cells of these tissues.

In addition to its function in the transportation of materials throughout the body, the circulatory system is important in temperature regulation. This arises by virtue of the ability of the muscular walls of the blood vessels to expand or contract, thereby changing the diameter of the vessels. When the capillaries in the skin are expanded or dilated, a larger amount of blood flows through them. If the temperature outside of the body is below body temperature, the blood in these capillaries is cooled. This cooled blood is then transported to the interior of the body where it is able to counterbalance any tendency toward a rise in temperature. On a cold day, these surface capillaries will be constricted so that the blood will not lose undue amounts of heat to the atmosphere.

The size of the various blood vessels thus varies automatically with the particular needs of the body. Drugs which cause a constriction of the blood vessels (*vasoconstrictors*) bring about a rise in blood pressure even though blood content remains fixed. By contrast, *vasodilators* generally bring about a reduction in blood pressure. Physiological changes in the sizes of the blood vessels are in part under the control of vasodilators and vasoconstrictors produced naturally in the body, and partially under the control of the nervous system. Sometimes, a substance that causes a constriction of the blood vessels in one tissue may dilate the vessels in another. The hormone secreted by the medulla of the adrenal glands is one example of a natural vasoconstrictor that aids in regulating the blood pressure in the body.

Cardiac Disorders and Diseases

Generally, three conditions are symptomatic of heart disease: (1) *Myocardial ischemia* (a decrease in blood supply to the heart muscle); (2) disturbances in *cardiac rhythm*; and (3) disorders in the *pumping efficiency* of the heart as may be manifested by increased filling pressure which causes upstream venous circulation, or decreased systolic pumping which results in an inadequate circulation of blood to organs that are located downstream of the heart. Common symptoms of heart problems include chest pain, palpitation, syncope (loss of consciousness), dyspnea (labored breathing), and edema (accumulation of fluid). These conditions, of course, are not exclusive to heart conditions.

Except in emergencies, when time is of the essence, milder symptoms of heart problems will be methodically diagnosed through the use of a number of instrumental techniques. The well-established cornerstone of heart diagnosis remains the electrocardiogram, preferably using twelve leads. Although the interpretation of electrocardiograms has been computerized to a degree and has been found useful in studies of mass populations, the input of an experienced cardiologist is considered mandatory in the analysis of specific patients with possible heart problems. During recent years, the two-step exercise procedure has largely been replaced by treadmill exercise. Ambulatory electrocardiographic measurements also have been emphasized in recent years. See **Electrocardiography.**

The use of ultrasound in a technique known as echocardiography has been available since the late 1960s and is growing in acceptance, being of particular value in the diagnosis of such conditions as pericardial effusion, mitral valve prolapse, and left atrial tumors, among others. See **Echocardiography.**

Another relatively new diagnostic tool is *isotope imaging*. Examples include radionuclide angiocardiography, using radioactive technetium, myocardial scanning with techetium pyrophosphate, and myocardial perfusion scanning with radioactive thallium Positron emission tomography (PET) is also used as a diagnostic tool in heart disorders.

Invasive procedures are still required in the diagnosis of many cardiac problems. These include cardiac catheterization, angiocardiography, coronary arteriography, intracardiac electrophysiological studies, and myocardial biopsy. These methods generally are limited to situations of an advanced, more serious nature and where other diagnostic procedures do not suffice. Principal limiting factors in their use are the risks generally attendant to invasive procedures, patient discomfort, and cost.

The major cardiac disorders and diseases are described in separate entries in this encyclopedia.

Valvular Heart Disease. The function of the valves of the heart has previously been described in this entry. See Fig. 1. At one time, most heart valvular damage was ascribed to rheumatic fever. See **Rheumatic Fever**. It has since been established that there are over twenty forms of nonrheumatic valvular diseases.

In *rheumatic heart disease*, there is fibrotic scarring of the valvular tissue which ultimately produces *stenosis* or *regurgitation*. In nearly all cases, some stenosis is present. With exception of rare congenital causes, *stenosis of the mitral valve* is usually considered of rheumatic origin. Stenosis is defined as the narrowing or contraction of a passage or opening. Regurgitation is the abnormal backward progression of fluids; in the case of the heart, the backward return of blood through the valves of the heart. Stenosis adds an extra load on the heart because of increased pressure required to overcome resistance to flow; regurgitation reduces the efficiency of the heart as a pump.

In *nonrheumatic mitral regurgitation*, there is the *floppy valve syndrome*, a dysfunction related to coronary artery disease as well. In floppy valve syndrome, in what is described as an idiopathic pathologic process, there is a loss of fibrous and elastic tissue; this is sometimes called *myxomatous degeneration*. Mitral regurgitation also may result from rupture of the papillary muscle (*papillae* are conical projections from the walls of the cardiac ventricles attached to the cusps of the atrioventricular valves by the *chordae tendineae*). Rupture may occur as the result of infarction or ischemia and thus contributes to mitral regurgitation.

Aortic valve stenosis in adults (particularly the elderly) is considered of nonrheumatic origin and results from a gradual but progressive degenerative thickening and calcification of the leaflets in the valves. The disease process is considered to be somewhat like that occurring in atherogenesis. See **Arteries and Veins.**

Aortic regurgitation is also generally considered a nonrheumatic disorder and frequently occurs as a secondary manifestation of other diseases (syphilis, ankylosing spondylitis, aortic dissection, aortic aneurysm, and inherited diseases that affect connective tissue).

Prevention and therapy in valvular heart disease include the long-term administration of prophylactic antibiotics to decrease the possibilities of a return of rheumatic fever for persons who previously have had acute rheumatic fever. The length of time during which such prophylaxis should be given is debatable among authorities. In persons with rheumatic heart disease featuring aortic regurgitation or a bicuspid aortal valve, most specialists suggest the administration of antibiotics to prevent the development of bacterial endocarditis after dental and surgical procedures. In valvular disease, the physician will be aware of the risk of systemic embolism which sometimes develops in connection with rheumatic heart disease. Long-term administration of anticoagulants may be indicated in such cases. Cardiac arrhythmias arising from valvular disease will be handled as described in the entry on **Arrhythmias (Cardiac).**

Surgery is frequently indicated in valvular heart disease. This may range from repair of malfunctioning parts to valve replacement. Over the years, over three dozen designs of *artificial (prosthetic) valves* have been used. Designs of preference in recent years have included the Starr-Edwards, the Smeloff-Cutter, and the Björk-Shiley valves. These valves are considered to have ample durability. The principal problems sometimes involved include thrombus formation and embolism, and thus long-term anticoagulant therapy is usually indicated.

In the United States, the natural tissue valve preferred is the porcine aortic valve. In Europe, some valves are configured from dura mater (outermost membrane of the brain and spinal cord), pericardium (membrane enclosing the heart), and fascia Jata (wide, dense sheath of the thigh muscles). The valves from pig hearts make excellent replacements for human heart valves. They are durable, resistant to infection, and not readily rejected by the human body. There has been a shortage of valves of the proper size from this source. The valves are taken from pigs of various sizes, with most of the animals weighing less than 80 pounds (36 kilograms). Since most pigs in the United States are slaughtered at around 200 pounds (90.7 kilograms), the supply of hearts from small pigs is limited. Also, only about one of every ten valves is suitable for placement in the human heart. The cost of raising pigs strictly for their heart valves has proved prohibitive. A number of countries slaughter pigs weighing less than 80 pounds (36 kilograms) and the hearts from these pigs can be obtained rather inexpensively at slaughterhouses. However, they have the potential of introducing exotic diseases of swine into the United States. Such diseases as African swine fever, hog cholera, foot and mouth disease, and swine vesicular disease could devastate the pork industry in the United States. To prevent the introduction of such diseases, scientists at the U.S. Department of Agriculture (Plum Island, New York) have developed a method for inactivating these viruses. They have found that glutaraldehyde, a substance used to stabilize pig heart valves prior to their transplantation in humans, will kill the viruses associated with these diseases. Nevertheless, great care must be exercised in making certain that all porcine valves are fully free of such viruses prior to surgery.

Cardiomyopathies. Dysfunctions of the heart muscle (*myocardium*) that are *not* related to coronary atherosclerosis, hypertension, or valvular problems, fall into four categories which when considered as a group are called *cardiomyopathies*. From the standpoint of hemodynamics (study of movements of the blood), these categories (Goodwin, 1970) are: (1) *congestive*; (2) *hypertrophic*; (3) *restrictive*; and (4) *obliterative*.

In *congestive cardiomyopathy*, the contractility of the heart muscle is subnormal. Common symptoms include dyspnea (labored or difficult breathing) and fatigue. Often pulmonary congestion accompanies the disorder. There is often mild elevation of blood pressure. Cardiac enlargement is common. This condition must be differentiated from acute myocarditis. The usual course of congestive cardiomyopathy is to congestive heart failure, ultimately the cause of death of persons with the condition. The prognosis is variable. Therapy includes salt reduction, digitalis glycosides, and diuretics.

For many years, it has been observed that congestive cardiomyopathy is frequently seen in alcoholics. The term *alcoholic cardiomyopathy* now frequently appears in the literature. Alcohol has not been definitely identified as the cause; possibly the malnutrition usually associated with alcoholism may be the major contributor. In the midwestern United States and Canada in the 1960s, there was an epidemic of cardiomyopathy, but this was ultimately traced to cobalt toxicity derived from an additive used in making the beer consumed in the region.

In recent years, there has been considerable rethinking as regards the possible connection between cardiomyopathy and coronary artery disease; in the past, the presence of cardiomyopathy by definition ruled out coronary artery disease.

A common cause of congestive cardiomyopathy in certain regions of South America is **Chaga's Disease,** which see.

Hypertrophic cardiomyopathy has been known for many years, but possibly well defined for the first time by Teare (1958), who termed the disorder "asymmetrical hypertrophy of the heart." Hypertrophy is an increase in the volume of a tissue or organ caused entirely by enlargement of existing cells. Asymmetry refers to the disproportionate hypertrophy of the left ventricle which effectively reduces the size of the left ventricular chamber. The result is obstruction to left ventricular outflow. In recent years, new names have been given to the disease—*muscular subaortic stenosis*; and *idiopathic hypertrophic subaortic stenosis*. Symptoms include angina, syncope, palpitations, and congestive heart failure. See **Congestive Heart Failure.** Although the symptoms of the disease worsen with time, the process may be slow—a span of years. In some cases, however, sudden death may occur, particularly in children and men with a family history of this condition. About 15% of cases are treatable by surgery. Drug therapy is not universal, but is directed toward the profile of symptoms presented.

In *restrictive cardiomyopathy*, the myocardium loses its resilience and becomes rigid—conditions which offer resistance to ventricular filling and elevate cardiac filling pressures. The condition tends to mimic constrictive pericarditis. Symptoms are those of congestive heart failure. There are no fixed therapies for this disease that have proven effective. Some authorities believe that removal of excess iron in the body by phlebotomy (incision of a vein) may provide some relief.

In *obliterative cardiomyopathy*, there is a massive fibrosis (formation of fibrous tissue) of the endocardium. This reduces the size of the ventricular cavities. Although the disease, of unknown etiology, is frequently seen in eastern Africa, it is seldom encountered in Europe and the Western world.

Pericarditis. Inflammation of the membrane enclosing the heart (*pericardium*) may take three fundamental forms, all of which are generally termed *pericarditis. Acute pericarditis* is usually associated with a viral infection. There is chest pain which increases with inspiration (contrast with myocardial infarction), a low-grade fever, and sometimes tachycardia. The physician will listen for the sounds of a characteristic pericardial friction rub. Where a bacterial infection is diagnosed, antibiotics will be used; for neoplasms, radiation or chemotherapy may be indicated. In *pericardial effusion*, fluids accumulate in the pericardial cavity. Echocardiography is commonly used in diagnosis. In acute forms of cardiac tamponade (compression of heart due to collection of fluid in pericardium), as may arise from an injury, an aortic dissection, or rupture of an aortic aneurysm, prompt surgery may be indicated. In a less severe situation, pericardiocentesis (puncture and aspiration) may be used. In *constrictive pericarditis*, diastolic filling of the heart is impeded, the results of which are an increase in venous pressure and reduced cardiac output. At one time, this condition was almost exclusively attributed to a tuberculous lesion. A majority of cases are classified as idiopathic, but some are related to radiation exposure, to rheumatoid arthritis, or uremia. Surgical removal of the pericardium is sometime indicated.

Sudden Cardiac Death. This term applies to the unexpected cessation of breathing and circulation when the hearts stops pumping, usually caused by an underlying heart disease, such as atherosclerosis of the coronary arteries. If the patient's breathing and circulation are not restored within a few minutes, permanent biological death, precipitated by irreversible brain damage, will result. It is estimated that between 20% and 30% of sudden cardiac deaths result from myocardial infarction; the remainder (statistics not yet reliable) is divided between myocardial ischemia and primary rhythm disturbance. Provided exceptionally effective emergency measures are applied (difficult in many situations), the long-term prognosis for attacks resulting from myocardial infarction or ischemia are good; they are poor in the case of a primary rhythm disturbance. About 25% of heart attacks can be classified as out-of-hospital sudden cardiac detahs, instances in which coronary heart disease has precipitated the attack with very little warning, often no warning whatsoever. Currently, fewer than 5% of sudden cardiac death patients are successfully resuscitated. The persons in the United States with coronary heart disease run into the several millions, of which 1.5 million (approximately) suffer heart attacks each year. It is estimated that 75% of these persons are admitted to a hospital in time (warnings noted hours, weeks. or months in advance), of which 80% are discharged, but usually having to follow some therapeutic regimen. On the other hand, 25% of the 1.5 million persons suffer sudden cardiac death outside of a hospital. Approximately 95% of these attacks are fatal (some 600,000 deaths per year).

In large communities, or exceptionally progressive smaller communities, some progress has been made in getting persons to a hospital barely in time to effect treatment. Emergency medical technician (EMT) teams have been formed. They have been trained for handling cardiac emergencies. The immediate treatment is cardiopulmonary resuscitation (CPR), which is a repeated series of mouth-to-mouth respirations and chest compressions that circulates a small amount of oxygenated blood to the brain, heart, and other vital organs. This is followed by specific medical treatment, designed to restore normal circulation and respiration. This usually includes the insertion of a breathing tube into the trachea, delivery of drugs, and defibrillation. The latter applies an electric shock across the victim's chest to depolarize all heart cells simultaneously and thus reset, so to speak, the pacemaking nodes of the heart. Defibrillation thus interrupts chaotic twitching of the heart muscle. Seattle, Miami, Los Angeles, and Columbus (Ohio) pioneered the EMT program. The concept, however, was first applied in Belfast (U.K.) in the late 1960s.

Coronary Artery Bypass Surgery and Percutaneous Coronary Angioplasty are discussed in article on **Ischemic Heart Disease.**

Congenital Disorders and Anomalies

Most congenital disorders of the circulatory system appear in the embryo as the result of some defect in development, usually between the fifth and eight week of pregnancy. An infection in the mother during pregnancy, or rubella (German measles), may be responsible for the abnormality. In some cases, the heart may be located in the right side of the body, although this seldom causes any difficulty and may not be noticed immediately. More serious defects are those which involve the size and development of the chambers of the heart, its valves, and connecting vessels. In some patients, such congenital defects may manifest themselves only after many years, and cause nothing more than a slight discomfort in breathing. In other instances, the defects may be such as to inhibit seriously the flow of blood through the heart and lungs.

In one of the malformations (*patent ductus arteriosus*), a small duct connecting the aorta and the pulmonary artery fails to close at birth. Since the pressure is higher in the aorta, blood will flow from this vessel to the pulmonary artery and back to the lungs, from which it had just come. This means that even when the lungs are working at full capacity, all of the oxygenated blood is not being circulated to the body. Difficulty in breathing and palpitation are outstanding symptoms. Once it is discovered, this defect can be repaired surgically by tying or dividing and sewing the open ends of the duct.

If defects exist which allow a mixing of arterial and venous blood, the patient frequently has a bluish or *cyanotic* appearance. This condition, if not corrected, may limit the life of the patient to a relatively few years. Best known of the cyanotic congenital heart defects are those that are found in "blue babies." One of the most common conditions causing blue babies is really a combination of four malformations (*tetralogy of Fallot*). In this disorder, the prenatal partition (*septum*) between the two pumping chambers (*ventricles*) of the heart has failed to close at birth. In addition, the major artery (*aorta*) leading from the heart is slightly out of place, and the artery leading from the heart to the lungs is constricted. The right ventricle, therefore, not only must pump blood through the lungs, but also must work directly against pressure from the left, so that the ventricle becomes enlarged because of the extra work. Blood which has been through the lungs becomes mixed with that which has not. An increase in the number of red blood cells may occur to compensate for the circulatory insufficiency. The child's fingers may be club-shaped and there may be a failure on the part of the child to develop physically in a normal manner. Breathlessness is common.

At one time, the treatment of blue babies was limited and consisted mainly in preventing infection and overactivity of the patient. The span of life was short. Now, in a special surgical procedure, one of the arteries—the *aorta, common carotid, subclavian,* or *innominate*—is connected to the pulmonary artery. There is then an increase of the blood flow to the lungs sufficient to permit the patient maximum activity without placing undue strain on the heart. This operation, when needed, is performed during the very early years of childhood. At a later date, the individual can be fully corrected with a second operation, utilizing the heart-lung machine.

Congenital Anomalies. Each of the four valves of the heart may have congenital anomalies. The *tricuspid valve* may have a deformity of the leaflets, known as *Ebstein's malformation of the tricuspid valve*. Or there may be *tricuspid atresia*, in which the valve never forms, preventing the normal flow of blood from the right auricle into the right ventricle. Instead, it flows from the right auricle into the left auricle through a hole in the wall between the two upper chambers of the heart. The *pulmonary valve* cusps are partially fused in some individuals and prevent the proper flow of blood, *pulmonary stenosis*. This condition can be caused by narrowing of the orifice leading to the valve or fusion of the leaves of the valve itself. In the normal heart, the systolic pressure is the same on both sides of the valve. If the pressure is found to be lower in the pulmonary artery than in the right ventricle, the physician knows that *pulmonary stenosis* exists. The mitral valve may have *atresia, incompetence*, or *stenosis*, although isolated cases of these conditions are rare. The aortic valve in the heart may have a congenital narrowing of the orifice or fusion of the cusps, known as *aortic stenosis*. Most of these abnormalities of heart valves can be corrected surgically.

The most common congenital malformation occurring as a single lesion is *ventricular septal defect*, in which there is a hole in the wall between the left and right ventricles. Following diagnosis, this abnormality can be corrected surgically by sewing a patch composed of a tough, resilient plastic material over the opening. A hole between the two auricles, *atrial septal defect*, allows blood to flow from the left side of the heart as the result of pressure differences. This defect can be corrected by directly suturing the edges of the defect.

A more complicated group of defects occurs when there is a hole between both the upper chambers (*atria*) and lower chambers (*ventricula*) with malformed intervening tissue and one or both valves between the atria and ventricula. These most difficult lesions can be corrected with the use of the heart-lung machine and require the use of a patch and sometimes a prosthetic valve.

In some cases, the oxygenated blood from the lungs returns partially or totally to the right side of the heart instead of draining into the left auricle. This type of malformation, *anomalous drainage of pulmonary veins*, is characterized by an abnormal condition—the same amount of oxygen being present in all the chambers of the heart, the pulmonary artery, and the aorta. This condition can be corrected by various surgical procedures in which the anomalous drainage is redirected into the correct left auricle.

In *coarctation of the aorta*, another rather common genital heart defect, the main artery leaving the heart is constricted to such an extent that the flow of blood to all parts of the body is restricted. When the diagnosis of this condition has been confirmed, the constriction can be removed surgically and the ends of the aorta reunited or the defect bridged with a synthetic vessel, thus allowing the blood to flow freely.

Cardiac Transplantation

Since the first human heart transplantation was accomplished by Christian Barnard, a South African surgeon, in December 1967, the practical feasibility of the procedure for extending life in patients with obviously terminal heart disease has been under severe scrutiny, not only by the medical professions, but by government regulators and the lay public. The gamut of technical, social, and economic pros and cons has been discussed extensively but not fully resolved. Moratoriums by governments and by hospital groups have been invoked and revoked. Aside from economic restraints, decisions to use or not use the procedure largely rest with the medical professionals and their institutions (hospital facilities, etc.) and, of course, with the patient.

In assuming that there will be improvements both in postoperative survival and quality of life, some authorities estimate that the number of *technically justifiable* (unrelated to socioeconomic factors) cardiac transplantations in the United States will not exceed 1000 to 5000 per year. This figure will be affected largely by the guidelines used by the medical profession in selecting candidate patients. One set of guidelines suggests that the prospective candidate not be over 50 years old and have no significant systemic disease other than very advanced cardiac malfunction. It is apparent that a major controlling factor will be the availability of donor hearts.

In 1984, it was estimated that the number of persons with irreversible brain death and identified as suitable allograft donors (of heart and other organs) does not exceed 2000 per year.

As pointed out by Austen and Cosimi, important improvements in cardiac transplantation have resulted from several factors: (1) better definition of criteria for selection of appropriate patients; (2) refinement of the use of antilymphocyte serum and T-lymphocyte monitoring for management of immunospression (organ rejection), (3) improved myocardial preservation due to perfection of effective cardioplegic techniques, and (4) the use of the fungal metabolite, *cyclosporine*. With cyclosporine, the episodes of rejection are less dangerous and easier to treat. Hospital stays, hence costs, have been reduced for those patients receiving this powerful and specific immunosuppressive drug.

It is generally felt that cardiac transplantation should be restricted to highly specialized medical centers. From a technical standpoint, the procedure is now viewed with cautious optimism.

Heart-Lung Transplantation. This comparatively new and highly selective procedure has thus far been used in the United States for a very limited number of patients who had terminal and irreversible pulmonary hypertension and were near death. Survival rate has exceeded 50%, but the number of procedures is so small that it is difficult to forecast future survival statistics. In a number of cases, after surgery, the short-term improvement has been excellent. Lung function has been restored to near-normal levels. Improvements in dyspnea, pulmonary parenchymal function, gas exchange, pulmonary vascular function, and cardiac function have been observed.

In this procedure, the lungs and heart are transplanted as a single unit. This simplifies the transplantation procedure. Cyclosporine is used to suppress graft rejection and appears to do so without the toxicity of conventional agents. A principal toxic effect of cyclosporine is impaired renal function in nearly all patients. Although renal function usually returns toward normal after the drug dose is reduced, some patients have required dialysis therapy, and a prolonged moderate impairment of renal function.

Heart-lung transplantation ultimately may be used in patients with other forms of lung disease, including obstructive lung disease, restrictive lung disease, and cystic fibrosis. Some authorities also stress that this procedure may lead to increased knowledge concerning the pathophysiology of diseases of the lung and pulmonary vasculature.

Cardiopulmonary Bypass Technology. This has been a key not only to heart transplantations, but to all procedures that involve "open heart" surgery. This technology (the so-called heart-lung machine) permits surgeons to operate on the heart for long periods of time in a dry, bloodless field, under direct vision. A pump draws blood from the vena cava, through tubes which are connected to these veins before they enter the heart. The blood is pumped under controlled pressure and flows to an "artificial lung" usually a plastic, membranous structure, where it is allowed to contact a steady stream of oxygen. The oxygenated blood is then pumped through another tube into the arterial system. The oxygen content, temperature, degree of alkalinity or acidity, rate of flow, and pressure, among other instrumental variables, must be carefully regulated throughout the entire surgical procedure. Checks on the circulation in the extremities are made continuously during the bypass procedure to prevent death of any tissues because of inadequate blood supply.

Artificial Heart

The concept of an artificial (prosthetic) heart dates back as early as 1812, when Julien-Jean Céesar La Gallois observed, "if one could substitute for the heart a kind of injection (of arterial blood), one would succeed easily in maintaining alive indefinitely any part of the body." Mechanical perfusion experiments with heart and lung organs date back a century ago (1880), exemplified by the work of Henry Martin. Martin's work prepared the foundation for modern cardiopulmonary bypass technology.

From a bioengineering standpoint, it is interesting to note that the total power output of the human heart is about 2.5 watts, of which 80% is required by the left ventricle (the output side of the heart which pumps blood into the arteries and ultimately to the capillaries). The pressure parameters are well established. See the entry on **Hypertension (High Blood Pressure),** which gives diastolic and systolic pressures.

By 1950, well over 30 artificial heart-lung designs had been proposed. Well known is the work of Lindbergh (Charles) and Carrel in the 1930s in connection with their perfusion pump, then reported by the news media as a "robot heart." Prominent in the search for an artificial heart has been the concept of a device that will be implanted in the human body and take over all heart functions. Lindbergh and Carrel added an interesting new dimension to this objective, as suggested by the following quotation from one of their publications: "We can perhaps dream of removing diseased organs from the body and placing them in the Lindbergh pump as the patients are placed in a hospital. There [the organs] could be treated far more energetically than within the organism and, if cured, replanted in the patient."

Working essentially with these criteria in mind, a number of teams have been researching and experimenting with artificial hearts. These include work at the Cleveland Clinic, dating back to the 1950s. In 1957, these researchers were able to keep dogs alive for about 1.5 hours with a plastic polyvinyl chloride heart energized by compressed air. It should be noted that these experiments were conducted at a time before attempts at human heart transplantation had been made, and when open-heart surgery was in the early pioneering stage. Research on artificial heart valves had just commenced. Progressively, these and other workers refined their designs and selection of materials of construction as well as various energy supplies, including electrically driven apparatus. Nuclear power as a source was considered. By the mid-1960s, researchers at the Cleveland Clinic were able to keep calves alive for 1.5 days with an artificial heart.

It should be mentioned that in England in 1928, Dale and Schuster built a pump with the objective of temporarily bypassing the heart during heart surgery. Dodrill (General Motors Corporation), in 1952, developed a mechanical heart which was used for nearly an hour during human heart surgery. Jarvik stresses, however, that open-heart surgery as it is known today requires the heart-lung machine, not simply a pump to replace the heart.

In 1969, for the first time, an artificial heart was installed in a human being. This artifact was designed by Liotta and Hall (Texas Heart Institute) and sustained life for about 64 hours, during which time a natural heart was being sought for transplantation.

In the early 1980s, Jarvik, a present-generation pioneer in the field, listed at least six criteria for what may be termed "the total artificial heart:" (1) *Small size*, to fit into the existing human cardiac cavity; (2) *work output* ample to provide all needs supplied by a natural heart; (3) a *variable output* in accordance with the changing rate of body requirements (range from rest to vigorous exercise); (4) *gentle handling of blood* to avoid hemolysis (disintegration of the elements of the blood); (5) *ease of sterilization*; and (6) *durability*. There are, of course, numerous other criteria, certainly one of which is economics.

The current consensus in the medical profession today targets the prosthetic heart has a means for maintaining life in a patient who is waiting for a suitable donor heart. This may be a period of days or several weeks. The Jarvik heart and other recently conceived designs have been used successfully for this purpose.

The technical principles of the prosthetic heart, namely, those of a pump, are not complex, but the detailed engineering in selecting materials, means of connection to human tissue, size, durability, resistance to rejection, strength, etc. are indeed very complex and unfortunately beyond the scope of this encyclopedia.

Additional Reading

Bashore. T. M., Ed.: "Invasive Cardiology: Principles and Techniques," B. C. Decker, Philadelphia, Pennsylvania, 1990.

Chatterjee, K., et al, Eds.: "Cardiology: An Illustrated Text/Reference," J. B. Lippincott/Gower Medical, Philadelphia, Pennsylvania, 1991.

Chou, Te-Chuan: "Electrocardiography in Clinical Practice," W. B. Saunders, Philadelphia, Pennsylvania, 1991.

Cohn, P. F.: "Silent Myocardical Ischemia and Infarction," 2nd Ed., Marcel Dekker, New York, 1989.

Francis, G. S., and J. S. Alpert: "Modern Coronary Care," Little, Brown, Boston, Massachusetts, 1990.

Jelliffe, R. W.: "Fundamentals of Electrocardiography," Springer-Verlag, New York, 1989.

Julian, D. G., et al., Eds.: "Diseases of the Heart," Bailliere Tindall, London, 1989.

Pepine, C. J., Ed.: "Acute Myocardial Infarction," F. A. Davis, Philadelphia, Pennsylvania, 1989.

Rowlands, D. J., Ed.: "Emergency Cardiology," Wright, Boston, Massachusetts, 1989.

Topol, E. J., Ed.: "Textbook of Interventional Cardiology," W. B. Saunders, Philadelphia, Pennsylvania, 1990.

HEARTBURN. See **Esophagus.**

HEART FAILURE (Congestive). See **Congestive Heart Failure.**

HEARTWORM DISEASE (Dirofilariasis). This is a serious and potentially fatal disease in dogs. It is caused by a worm (*Dirofilaria immitis*) which is found in the animal's heart and large adjacent vessels. The female worm is 6–14 inches (15.2–35.6 centimeters) long and about $\frac{1}{8}$ inch (3 millimeters) wide. The male is smaller. One dog may have as many as 300 worms. Adult heartworms live in the animal up to 5 years and during that period the female produces millions of young *microfilariae*. These microfilariae live in the bloodstream, mainly in the small blood vessels. They cannot grow to adults without passing through an intermediate host (a mosquito). As many as 30 species of mosquito can serve as host. The microfilariae develop for 10–30 days in the mosquito and then enter the saliva of the insect. At this point, the organisms are *infective larvae* because at this stage of development they will grow to adults when they enter a dog. The mosquito bites the dog, mostly on the abdomen where the haircoat is thinnest.

Adult worms cause disease by clogging the heart and major blood vessels leading from the heart. They interfere with the valve action. By clogging the main blood vessels, the blood supply to other organs of the body is reduced, particularly the lungs, liver, and kidneys. Most dogs infected with heartworms do not show external signs of the disease. When symptoms develop after some period of infection, these will include soft dry chronic cough, shortness of breath, weakness, nervousness, listlessness, and loss of stamina. These features are noticed particularly after exercise. The microfilariae circulate throughout the body, but remain mainly in the small blood vessels, which they tend to clog. Ultimately there is destruction of lung and kidney tissue.

An arsenical drug is used in treatment and usually requires a hospital environment. The treatment requires injections of the drug over a period of 2–3 days. About 6 weeks after the adult worms have been eradicated, further injections of other drugs are required to eradicate the microfilariae. Some veterinarians prefer to eradicate the microfilariae first. To prevent heartworms, many veterinarians recommend the use of diethylcarbamazine citrate (e.g., Filarbits®) in the pet's diet during the mosquito season.

HEAT. The agency whose addition to or removal from a physical system is the cause of thermal changes of various types. These include rise and fall of temperature, changes in length and volume, changes of physical states, such as melting, evaporations, etc.

During the eighteenth century heat was assumed to be a subtle fluid called *caloric*, filling the interstices between the ultimate particles of matter and, under conditions of isolation from the surroundings, known to satisfy a conservation law. The production of heat by friction as well as its disappearance during the performance of external mechanical work established its essential physical nature as another form of *energy* and led to the overthrow of the caloric theory. Nevertheless, we still speak of the *flow* of heat as though it were a fluid and have retained the methods of measuring the *quantity of heat* originally devised by the upholders of the caloric view.

Our direct knowledge of heat is provided by the sensation of hotness and coldness when we come in contact with various physical bodies. It is possible to arrange a set of bodies in a sequence such that A feels hotter than B, B hotter than C, etc. We say that A has a higher *temperature* than B, B a higher one than C, and so on. Of course our sensations are qualitative and are considerably influenced by the thermal conductivity of the body we touch. Thus, on a frosty morning, the head of an ax being metal feels considerably colder than the wooden handle though the two are presumably at the same temperature. To obtain a continuous and reproducible physical scale of temperature, various types of thermometers have been devised of which the mercury-in-glass or colored-alcohol-in-glass are familiar examples. The two temperature scales in common use are the Fahrenheit scale and the Celsius scale. The first assigns values of 32° and 212° to the normal freezing and boiling points of pure water, respectively, and divides this interval into 180 equal sub-intervals or degrees. The Celsius, formerly called the Centigrade scale assigns the respective values of 0° and 100° to the above fixed points; the standard interval is then divided into 100 equal degrees.

Temperature changes are produced by the addition or subtraction of heat from a body. Thus, temperature may be regarded as a measure of the concentration or *intensity* of heat. In general, the more heat we add to a given body the more its temperature rises.

Measurement of Heat. Since heat is imponderable and not directly observable, it is necessary to measure the size of a given quantity of heat by its effect on another body. If this effect is the production of a rise in temperature from some initial temperature, t_1, to a final temperature, t, then the rise $(t - t_1)$ is found to vary inversely with the mass of the test body. It is thus natural, following the calorists, to regard the quantity of heat, say Q, as determined by the product of m and $(t - t_1)$. Thus we say

$$Q \text{ is proportional to } m \times (t - t_1)$$

To make this statement into an equation we write

$$Q = \text{constant} \times m \times (t - t_1) \tag{1}$$

where the constant of proportionality depends on the substance, being large for some materials and small for others. This constant for water, for example, is about 33 times as great as for lead; water is said therefore to have a greater *heat capacity* than lead. Notice that the constant in Equation (1) actually gives the numerical value of Q which is required to warm a unit mass of the substance through a temperature interval of exactly 1°. This constant is accordingly called the *specific heat capacity* (usually abbreviated to *specific heat*) and is indicated by c. Since it is found that the value of the specific heat, particularly for gases, but in principle for all materials, depends on the conditions under which the heat is absorbed, this must be indicated. We thus have c_p and c_r, for example, for the two important cases of absorption at constant pressure and constant volume, respectively. Since the former characterizes the common laboratory case of working under atmospheric pressure, we accordingly rewrite Equation (1) as

$$Q_p = c_p m(t - t_1) \tag{2}$$

Q_p now measures the heat absorbed under constant pressure, and c_p is the constant pressure specific heat. Since the right side of Equation (2) contains *three* quantities, a mere choice of a mass unit and a degree unit is insufficient to establish a unit of heat. It is necessary to select some substance as a standard reference body and assign an arbitrary value of, say c_p equal to unity for it. Water is the universal choice for this standard body due not only to its cheapness and ease of purification, but also to its large heat capacity.

With the selection of water as the standard with $c_p = 1$, the left side of Equation (2) clearly becomes of unit value when m and $(t - t_1)$ are each of unit value. In the English system, we accordingly have the *British thermal-unit* (or Btu) as the heat required to warm 1 pound of pure water through an interval of 1°F. In the metric system, the corresponding unit is the *calorie*, the heat required to warm 1 gram of water 1°C. A large unit or *kilocalorie* corresponding to 1,000 ordinary calories is also frequently used in scientific work.

Specific Heats. Use of Equation (2) reveals that the values of c_p obtained experimentally depend on the temperature interval used, indicating a dependence of c_p on temperature. Thus, if c_p for water were actually uniform throughout the 0 to 100°C range, a mass of water at 100°C mixed with an equal mass at 10°C would give a final mixture at exactly 50°C. The actual value is near 50.05°; this difference although small, indicates the need to specify the calorie at some particular temperature. For this purpose, we suppose a system of mass m is warmed from t to $t + \Delta t$ by the addition at constant pressure of an increment of heat ΔQ_p. Then Equation (2) becomes

$$\Delta Q_p = m\bar{c}_p \Delta t \tag{3}$$

where now $\bar{c}_p$ is an average value of c_p over this interval. Then we define the *instantaneous* heat capacity, c_p at t by the following relation

$$c_p \frac{1}{m} \lim_{\Delta t \to 0} \frac{\Delta Q_p}{\Delta t} = \frac{1}{m} \frac{dQ_P}{dt}$$

i.e., the heat absorbed per unit mass per degree as the interval becomes smaller and smaller without limit. This leads to the differential form of Equation (3)

$$dQ_p = mc_p \, dt \tag{4}$$

where dQ_p is the differential heat absorption which produces a differential temperature rise dt in a body of mass m and specific heat c_p.

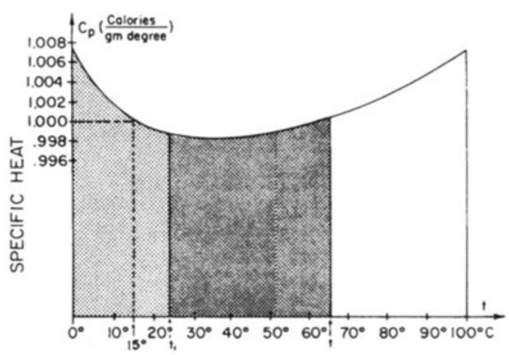
The standard or 15° calorie is now defined as the rate of absorption of heat per gram per degree at 15°C and in practice is essentially the same as the average calorie over the 1° interval from 14.5 to 15.5°C.

If a mass m of water is warmed from t_1 to t, the integral of Equation (4) gives for the total heat absorbed in 15° calories

$$Q_p = \int_{t_1}^{t} dQ_p = m \int_{t_1}^{t} c_p\, dt = m\left[\int_{0}^{t} c_p\, dt - \int_{0}^{t_1} c_p\, dt\right] \quad (5)$$

where the integral of c_p over the range t_1 to t has been written as the difference of two integrals from a common lower limit of 0°C. If, therefore, we evaluate an integral of the type $\int_0^t c_p\, dt$ with t varying in 1° steps and arrange these in a table, the right side of Equation (5) may be evaluated by merely subtracting appropriate entries.

In the accompanying figure, the value of c_p in 15° calories per gram per degree is plotted graphically from 0 to 100°C, and the integrals on the right of Equation (5) are represented by appropriate areas under the c_p curve. Thus the integral from 0° to t is hatched with lines sloping up to the right, while that from 0° to t_1 has the lines sloping up to the left. The value of Q_p is then the singly hatched area.



Specific heat of water versus temperature.

With heat quantities measured in 15° calories, from the observed rise or fall of temperature in known masses of water, the specific heats of various substances, the heats absorbed on melting solids to liquids (heats of fusion), the heats absorbed on passage from the liquid to the vapor state (heats of vaporization), the heats evolved on combination of various substances, and the heats absorbed or evolved in chemical changes are at once determinable (see **Calorimetry**). For the present purpose, the accompanying table gives the values of the constant pressure heat capacities of a few typical substances, variations with temperature being disregarded. Notice that c_p, although expressed in terms of calories per gram per degree, is in fact independent of the system of units since water is the reference body in all systems. Thus the specific heat of water in the English system would be 1 Btu per pound per degree Fahrenheit.

APPROXIMATE CONSTANT-PRESSURE SPECIFIC HEAT OF SELECTED MATERIALS

Substance	State	c_p(cal/g deg)
Water	Vapor	0.48
Water	Liquid	1.00
Water	Solid	0.50
Ethyl alcohol	Liquid	.54
Hydrogen	Gas	3.44
Air	Gas	.24
Aluminum	Solid	.22
Iron	Solid	.11
Lead	Solid	.03

The Mechanical Nature of Heat. The conservation of heat *per se* is observed only for systems involving the performance of no mechanical or electrical work. (Count Rumford (ca. 1800) was the first to establish this fact in his famous cannon-boring experiments carried out in the arsenal of the Dutchy of Bavaria in Munich. He observed that when his drills became dull, heat was produced in great quantities limited only by the amount of work done against friction. He concluded that the large scale mechanical energy used in overcoming friction could only be converted into the motions of the ultimate particles of matter, a motion not directly observable but detected by our senses as heat. His results were confirmed and extended by the later work of Joule and Helmholtz, in particular, and also provided a more reliable value for the so-called *mechanical equivalent of heat*. This is taken as the amount of mechanical (or electrical) energy which when converted into heat is equivalent to exactly 1 calorie. The presently accepted value for this important constant is 4.185 joules per 15° calorie. Here the joule is the work performed when power is expended at the rate of 1 watt for 1 second. Thus an ordinary 100-watt lamp bulb converts 100 joules of electrical energy to thermal each second; this amounts to 100/4.185 or about 24 calories.

As a result of experiments such as these and a host of others, we are forced to recognize that heat is merely another form of the universal quantity *energy*. Its transformation always occurs at the rate of 4.185 joules per calorie whether heat goes into external work or work is dissipated through friction into heat.

See also entries which follow; and **Thermodynamics.**

Additional Reading

Butterworth, D., and C. F. Mascone: "Heat Transfer Heads Into 21st Century," *Chem. Eng. Progress*, 30 (September 1991).

Gubbins, K. E.: "The Future of Thermodynamics," *Chem. Eng. Progress*, 38 (February 1989).

Jolls, K. R.: "Understanding Thermodynamics Through Interactive Computer Graphics," *Chem. Eng. Progress*, 64 (February 1989).

Kenney, W. F.: "Current Practical Applications of the Second Law of Thermodynamics," *Chem. Eng. Progress*, 57 (February 1989).

Sandler, S. I.: "Chemical and Engineering Thermodynamics," Wiley, New York, 1989.

Seely, J. H.: "Elements of Thermal Technology," McGraw-Hill, New York, 1981.

HEAT (Atomic). See **Atomic Heat.**

HEAT BALANCE (Distillation). See **Distillation**

HEAT BALANCE (Planet). The equilibrium which exists on the average between the radiation received by a planet and its atmosphere from the sun and that emitted by the planet and atmosphere. That the equilibrium does exist in the mean is demonstrated by the observed long-term constancy of the earth's surface temperature. On the average, regions of the earth nearer the equator than about 35° latitude receive more energy from the sun than they are able to radiate, whereas latitudes higher than 35° received less. The excess of heat is carried from low latitudes to higher latitudes by atmospheric and oceanic circulations and is reradiated there.

HEAT BALANCE (Process). A heat balance is a method of accounting for all heat units in a process or change during which heat is transferred. Examples of cases where heat balances might be undertaken are: (1) Determining the nature and the magnitude of the various losses which occur when fuel is burned in a steam boiler furnace. (2) Accounting for all heat units during the operation of a prime mover, such as a Diesel engine or a steam turbine. (3) Determining the distribution of heat in a static heating device, such as a water heater supplied with steam.

Heat balance work is based upon the first law of thermodynamics, a statement of which is: Energy may not be created or destroyed, but may be converted from one form to another. The significance of this law applied to the heat balance is that the total energy may be accounted for by straight addition, hence striking a heat balance resembles bookkeeping, with heat supplied on the credit side of the ledger, and various heats usefully employed on the debit side. One way of showing a heat balance is a tabular form; another shows the heat as a stream, properly branched and subdivided to indicate the distribution of heat. Briefly, a heat balance might be said to be the bookkeeping by which heat supplied is shown to be equal to the sum of heat utilized and lost.

(It should be added that the above statement of the First Law, while adequate for many engineering calculations, is subject to modification in accordance with the principle of mass-energy equivalence.)

HEAT CAPACITY. The amount of heat necessary to raise the temperature of a system, entity, or substance by one degree of temperature. It is most frequently expressed in calories per degree centigrade or Btu per degree Fahrenheit. If the mass of a substance is specified, then certain derived values of the heat capacity can be obtained, such as the atomic heat, molar heat, or specific heat.

HEAT CAPACITY EQUATION (Einstein). A quantum relationship for the heat capacity at constant volume of an element of the form:

$$C_v = 3R\left(\frac{h\nu}{kT}\right)^2\left(\frac{e^{h\nu/kT}}{(e^{h\nu/kT}-1)^2}\right)$$

in which C_v is the heat capacity at constant volume for one gram-atom of an element, R is the gas constant, h is Planck's constant, k is the Boltzmann constant, ν is the characteristic frequency of oscillation of the atoms of the element, T is the absolute temperature, and e is the natural logarithmic base.

The Einstein equation was the first approximation to a quantum theoretical explanation of the variation of specific heat with temperature. It was later replaced by the Debye theory of specific heat and its modifications.

HEAT CONSERVATION. See **(Insulation (Thermal).**

HEAT CONTENT. See **Enthalpy.**

HEAT ENGINE. As used in thermodynamics the term denotes a thermodynamic system, e.g., a sample of gas, carried through a cyclic process in such a way that a closed path is traced out on a pressure-volume (P-V) diagram, and positive work is done by the system. If Q_1 is the positive amount of heat energy absorbed by the system, Q_2 the positive amount of heat energy rejected by the system and W the net amount of work done by the system, then the first law of thermodynamics (conservation of energy) gives $W = Q_1 - Q_2$. The efficiency of the engine is defined as

$$\eta = \frac{W}{Q_1} 1 - \frac{Q_2}{Q_1}$$

For an engine following a reversible Carnot cycle (Carnot engine), the efficiency is given by $\eta = (T_1 - T_2)/T_1$, where T_1 is the Kelvin temperature of the reservoir at which Q_1 is absorbed, and T_2 is the Kelvin temperature at which Q_2 is rejected. The second law of thermodynamics states that no engine working between these same two temperatures can have a greater efficiency than that of the Carnot engine.

A thermodynamic engine run backwards becomes a *refrigerator*. Thus a positive amount of heat Q_2 is absorbed at a low temperature, work W is done, and positive heat Q_1 is rejected at a higher temperature. The first law now gives $Q_1 = W + Q_2$. The ratio Q_2/W is known as the *coefficient of performance* of the refrigerator. See also **Solar Energy.**

HEATER (Hysteresis). See **Hysteresis Heater.**

HEAT EXCHANGERS. See **Heat Transfer.**

HEATHER SHRUBS AND TREES. The heather or heath family (*Ericaceae*) is comprised of a number of genera, many species, hybrids, clones, and cultivars. Three of the main genera are: *Arbutus*, small to large evergreen trees, of which the strawberry tree (not to be confused with the fruit-bearing plant of the rose family) and the madrona tree are examples; *Clethra*, a small genus of deciduous or evergreen trees or shrubs, of which the Lily-of-the-valley clethra is representative; and *Rhododendron*, evergreen or deciduous shrubs (usually) or trees, of which azaleas and rhododenrons are members.

The strawberry tree (*Arbutus unedo*) is characterized by small whitish flowers occurring in clusters, a small fruit, about $\frac{1}{2}$ inch (1.2 centimeters) in diameter, which appears something like a strawberry, and narrow, oval leaves of medium length. This tree, which can reach a height of 40 feet (12 meters), does well in southwestern Ireland and the Mediterranean region. However, the tree can withstand somewhat colder climes and is found in parts of Britain and North America. Some authorities describe the fruit more as a roughened cherry than as a strawberry. Flowering occurs in late autumn, a definite attraction to the gardener. The *A. andrachne* is the strawberry tree of the eastern Mediterranean region. It is a slightly smaller tree, generally attaining a height of about 30 to 35 feet (9 to 10.5 meters). The flowers are an off-white and occur in broad clusters. Flowering occurs in the spring. The fruit is similar to the *A. unedo*.

The strawberry tree of western and southwestern North America is the *A. menziestii*, or, as commonly termed, the *madrona* or *madrone* tree. As shown by the accompanying table, there are closely related, localized species, such as the *A. arizonica* and *A. texana*. As noted, these trees are capable of achieving excellent heights under favorable conditions. The leaves are of medium length, oval, with dark green coloration above and a bluish-white color underneath. The tree usually flowers late in the spring. The fruit is a pea-sized berry. The tree is also found in Europe, where it may achieve a height of 50 to 55 feet (15 to 16.5 meters).

Not previously mentioned, the genus *Oxydendrum* claims the sorrel tree (*O. arboreum*), which occurs in the eastern United States and is related to the strawberry tree. The sorrel is a relatively small tree, ranging from 20 to 55 feet (6 to 16.5 meters) in height, with a trunk diameter up to about 20 inches (50.8 centimeters). The bark is gray-brown, somewhat furrowed. The branches are pendulous. Leaves are elliptically shaped, pointed, dark green, and finely-toothed. The flowers are white and occur in drooping clusters. The tree occurs from Pennsylvania westward to Indiana and southward along the Alleghany Mountains into Louisiana and western Florida. Some of the tallest specimens are found on the eastern slopes of the Blue Ridge Mountains.

Of the genus *Clethra*, the Lily-of-the-valley clethra attains a height of about 30 feet (9 meters) and can be classified as a small tree. The tree is found on the Island of Madeira, but has been introduced elsewhere. The leaves are dark green and alternate; the flowers are white and fragrant. The tree is sensitive to climate and soil, requiring a rich and acid mix.

It is interesting to note that prior to 1820, rhododendrons other than the European and American varieties were unknown. In particular, the *R. maximum* of the eastern United States was best known then. In that year, the *Rhododendron arboreum* was brought out of the Himalayas. During the intervening years, numerous species from Asia have been found and propagated.

The *R. maximum*, also sometimes referred to as the Great Laurel or rose bay, is a shrub/tree that can attain a height of 40 feet (12 meters) under favorable conditions, with a trunk diameter approaching a foot (0.3 meter). The bark is gray-brown, smooth, with minor scaling. The leaves are evergreen and lustrous, quite large—from 4 to 8 inches (10.1 to 20.3 centimeters) in length. The flowers occur in large clusters and may be described as pale pink with spots of coloration in the upper part of the throat. In nature, the plant is found, usually in damp woodsy areas or along streams, from Nova Scotia westward through Quebec and Ontario to Ohio and Lake Erie, and as they range southward, they become more numerous, notably through the Alleghany Mountains and in the southeastern states as far as Georgia.

Other species of rhododendrons occurring in the mountains and woods of the United States, notably east of the Rocky Mountains include: *R. viscosum*, also known as clammy azalea or white swamp hon-

RECORD MADRONE AND RELATED TREES IN THE UNITED STATES[1] (Heather or Heath Family)

Specimen	Circumference[2]		Height		Spread		Location
	(Inches)	(Centimeters)	(Feet)	(Meters)	(Feet)	(Meters)	
Clethra or Cinnamon Tree (*Clethra acuminata*)[3] (1981)	11	28	27	8.2	12	3.7	South Carolina
MADRONES							
Arizona madrone (1970) (*Arbutus arizonica*)	143	363	53	16.2	52	15.8	Arizona
Pacific madrone (1955) (*Arbutus menziesii*)	408	1036	96	29.3	113	34.4	California
Texas madrone (1982) (*Arbutus texana*)	112	284	32	9.8	42	12.8	Texas
RHODODENDRONS							
Catawba (1985)[4] (*Rhododendron catawbiense*)	16	41	14	4.3	14	4.3	Virginia
Catawba (1991)[4] (*Rhododendron catawbiense*)	10	25	26	7.9	11	3.4	North Carolina
Pacific (1976) (*Rhododendron macrophyllum*)	20	51	33	10.1	20	6.1	California
Rosebay (1981) (*Rhododendrom maximum*)	25	64	40	12.2	22	6.7	South Carolina

[1]From the "National Register of Big Trees," The American Forestry Association (by permission).
[2]At 4.5 feet (1.4 meters).
[3]Not to be confused with the cinnamon (spice) tree (*Cinnamomum zeylanicum*) of the laurel family.
[4]Cochampion.

eysuckle; *R. nudiflorum*, also known as Pinxter flower; *R. arborescens* or the smooth azalea; *R. canescens* or mountain azalea; *R. calendulaceum* or flame azalea; *R. canadense* or rhodora; *R. catawbiense* or rose bay; *R. lapponicum* or Lapland rose bay; and *R. hispida* or rose acacia of the southeastern United States. Rhododendrons from Asia include the previously mentioned *R. arboreum*. This plant is found in the Himalayas in the Khasia Hills, Sri Lanka and from Kashmir to Bhutan. It displays bell-shaped flowers of dark red color. The height ranges from 30 to 40 feet (9 to 12 meters). The *R. barbatum* is found in Bhutan, Nepal, and Sikkim and is capable of growing to a height of 40 feet. The flowering is similar to that of *R. arboreum*. The *R. calophytum*, with similar flowers but ranging from white to rose pink and with a characteristic maroon blotch, is found in western China. It can attain a height of about 35 feet (10.5 meters). Also from the Bhutan, Nepal, and Sikkim regions is the *R. falconeri*, with purple-blotched white flowers, and reaching a height of about 30 feet (9 meters). The *R. giganteum* of Yunnan is well named because it ranges in height between 40 and 80 feet (12 and 24 meters). The flowers are of a deep rose-crimson color. Also of Yunnan and of southeastern Tibet and upper Burma is the *R. sinogrande* which can rise to a height of about 45 feet (13.5 meters), displaying white/yellow flowers.

In a brief description of the heather family, certainly the common heather shrub (*Culluna vulgaris*) should not be omitted. This is a small, straggly shrub ranging from about 6 to 16 inches (15.2 to 40.6 centimeters) in height. It was introduced into America from Europe. The leaves are small, gray-green, perhaps $\frac{1}{8}$ inch (0.3 centimeters) in length and overlap the branches. Tiny bell-shaped flowers, white or of a deep pink color, occur as spikes. The Scottish heather counterpart is the *Erica cineria*. Another heather, the *Erica Tetralix*, was introduced along with the aforementioned two species into Nantucket Island, at the same time the Scots pine was introduced, during the years 1875–1877.

HEATING DEGREE DAY. See **Climate.**

HEATING (Geothermal). See **Geothermal Energy.**

HEATING OILS. See **Petroleum.**

HEATING (Solar). See **Solar Energy.**

HEATING VALUE (Coal). See **Coal.**

HEATING VALUE (Natural Gas). See **Natural Gas.**

HEAT (Molecular). See **Molar Heat.**

HEAT OF COMBUSTION. See **Calorimetry; Combustion.**

HEAT STORAGE (Solar). See **Solar Energy.**

HEAT PUMP. A system involving a compressor, heat exchangers, a refrigerant, and a flow restriction that can be used to supply or remove heat. In its cooling cycle, the traditional heat pump operates very much like a conventional air conditioner. Although the principle of the heat pump has been known for decades, it has received renewed interest in recent years in connection with the search for more energy efficient systems. The fundamentals of the heat pump for both its heating and cooling cycles are described briefly in the caption for Fig. 1.

The heat pump has been well established as an efficient user of electrical power. However, heat pumps are not universally applicable to year-round heating/cooling applications if they are to operate at maximum efficiency and thus outperform separate heating and cooling equipment. Comparisons can be made by determining the coefficient of performance (COP) of a given heat pump for a given application. The COP is a ratio, namely, of the amount of heat required by the condenser (exchanger A in diagram b), or Q_c, divided by the quantity of electrical energy consumed to power the pump, or W. That is, COP = Q_c/E.

Assuming that the average efficiency of an electric generating and distributing utility is 30%, the use of a heat pump can bring the overall heating efficiency up to 75%. (See "Design Improvements" later in this article.) It is interesting to note that a heat pump can yield more thermal

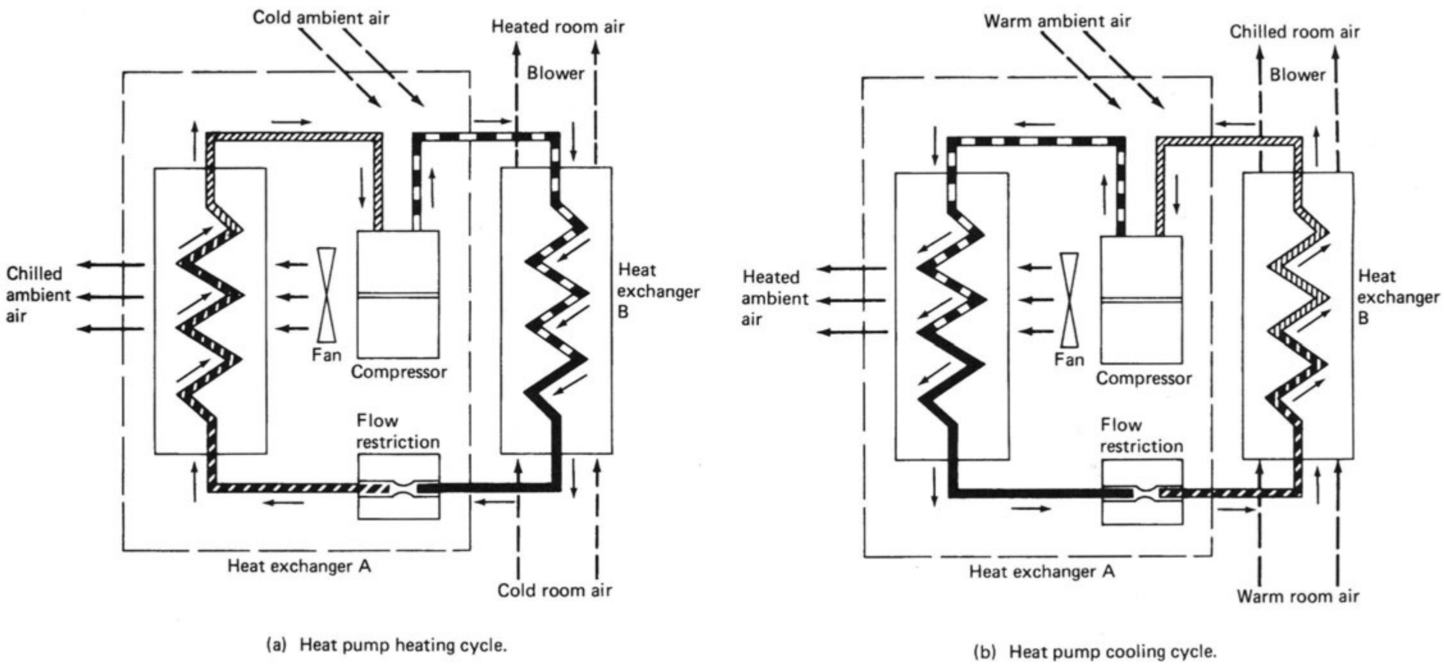


Fig. 1 Operating cycles of traditional heat pump. Although the heat pump, which can be used for both heating and cooling, is a relatively simple concept, its operation is sometimes not fully understood from the first reading of a description. Rather than begin with the usual comparison of a heat pump with an air conditioning unit and making reference to the terms condenser and evaporator, the present description commences with the heating cycle and refers to heat exchangers.

(a) There are four principal elements of equipment—a compressor, a flow restriction, and two heat exchangers, *A* and *B*. These are represented very schematically in the diagram. The heat-exchange medium (a refrigerant liquid, such as Freon*O*) exists the compressor in the vapor phase at a high temperature and pressure. It passes to heat exchanger *B*, where it is gradually cooled by the cold room air, which it in turn warms. The medium exits heat exchanger *B* in the liquid phase at moderate temperature, but still under high pressure, then proceeds to a flow restriction which effects a pressure drop. The medium exits the restriction as mixed liquid and vapor phase at a much lower temperature and pressure. (The lower temperature is the result of cooling caused by expansion.) This liquid/vapor mixture enters heat exchanger *A*, where it absorbs some heat from cold ambient air, making the ambient air just a bit colder in the vicinity of the unit. The medium exits the exchanger *A* as a vapor at low temperature and pressure, and returns to the compressor, where the cycle begins anew.

(b) By using a simple arrangement of valves, the flows can be reversed to make the heat pump a means for chilling room air rather than warming it. The heat-exchange medium exits the compressor in the vapor phase at high temperature and pressure. It passes to heat exchanger *A*, from which it exits in the liquid phase at moderate temperature and high pressure, then passes through the flow restriction, from which it exits as a mixed liquid and vapor at a low temperature and pressure. Again, the cooling is the result of expansion of the vapor. This mixed liquid and vapor phase enters heat exchanger *B*, where it is gradually warmed by warm room air, which it in turn cools. The medium exits heat exchanger *B* in the vapor phase at low temperature and pressure, and thence returns to the compressor, where the cooling cycle begins anew.

It will be noted that during the heating cycle the medium actually extracts heat from already cold ambient air as it passes through heat exchanger *A*. In contrast, during the cooling cycle, the medium actually adds heat to already warm ambient air as it passes through heat exchanger *A*.

It will be evident that, when operating in the cooling cycle, the heat pump operates like a conventional air conditioner. Heat exchanger *A* is the evaporator and heat exchanger *B* is the condenser. These units reverse their roles for the heating cycle.

energy than the electrical energy which it consumes when operating under certain conditions. This in no way defies the law of energy conservation because the heat pump picks up increments of thermal energy from the evaporator when used in the heating cycle. As pointed out in diagrams *a* and *b*, when on the heating cycle, a heat pump system, by virtue of absorbing heat from already cold ambient air, dumps air that is below ambient temperature to the atmosphere. And, when on the cooling cycle, a heat pump system, by virtue of absorbing heat from already warm room air, dumps air that is above ambient temperature to the atmosphere.

For a given capacity heat pump, the volume flow rate of refrigerant vapor through the compressor is approximately constant. It will be noted from diagrams *a* and *b* that in either the heating cycle or cooling cycle the medium taken into the compressor is in the vapor phase at a comparatively low temperature and pressure; and that the medium exiting the compressor is in the vapor phase at a comparatively high temperature and pressure. It is evident from diagram *a* that the temperature of the cold ambient air in its effect on heat exchanger *A* determines the temperature of the medium exiting the exchanger and thus entering the compressor. Thus, the colder the ambient air, the lower will be the vapor pressure and the density of the medium. These conditions reduce the effective mass of the medium moving through the compressor. This decreased mass flow rate lowers the thermal capacity of the medium, thus reducing the quantity of heat energy which it can transfer from the compressor to heat exchanger *B* and thence to heat the conditioned space. Depending upon the base capacity of the unit, the point may be reached where the ambient temperature is just too low to permit the heat pump to heat the conditioned space adequately. In the present state of the art, this leaves the designer two options—either use a unit with greater capacity (larger initial investment, etc.) or arrange to furnish for auxiliary heating during abnormally cold periods when such conditions may persist. Depending upon the form of the auxiliary energy and the efficiency of the auxiliary system, some or all of the advantages of the heat pump (cost, normal operating efficiency, etc.) may be negated. In general, it has been a practice to size the heat pump to the summer's cooling requirements and permit the winter heating to fall where it may, with dependence upon auxiliary heating. In southern climes, this may be acceptable because relatively little if any heating may be required of the unit during the winter season. In northern climes, designing to the summer cooling load will normally lead to the requirement for auxiliary heating.

Design Improvements. As has been pointed out over the years since the introduction of the heat pump, there has been a wide disparity of opinion among experts between the ideal heat pump performance and the actual performance of the equipment. Some of the shortfall of performance, particularly in older installed equipment, arises from an emphasis on minimizing manufacturing costs (thus initial costs) and consumer prices. For example, heat exchangers, condensers, and evaporators can be made of a relatively small size to keep initial costs down. But small components have limited capacity to transfer heat be-

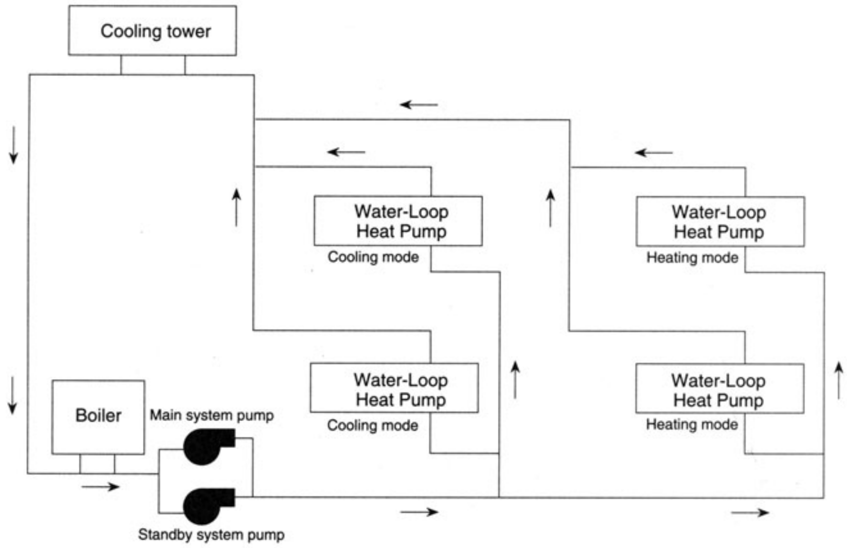


Fig. 2. A typical water-loop heat pump system, consisting of a water circulation loop (a two-pipe supply and return system), a series of heat pumps that use the water as a heat source or sink to perform heating or cooling, and a boiler and a cooling tower that operate as required to keep the temperature of the circulating water within an optimum temperature range. Heat pumps operating in the cooling mode add heat to the loop; those operating in the heating mode extract heat from it. (*Electric Power Research Institute.*)

tween refrigerant and air. To achieve a high rate of heat transfer with a small heat exchanger requires a rather large temperature difference between the refrigerant and the air. Thus, in the heating mode, the refrigerant in the condenser must be much warmer than the indoor air, and the refrigerant in the evaporator must be much colder than the outside (ambient) air. In some designs, to maintain these exaggerated temperature differences, the compressor literally must work overtime. The expected result is that the cost of performance is lower than it would be with larger, more expensive heat exchangers.

To improve heat performance, several concepts have been proposed in recent years. These include: (1) Use of a volume of water to store and provide low-temperature heat—with some ice forming in the evaporator and deliberately used. The concept has been called the "annual cycle energy system." (2) Use of a heat pump to supplement a solar collector system. (3) Combined use of a thermal storage system and heat pump to partially solve the problems of oversize heat pumps for cooling, particularly in northern climes. (4) A system for varying the capacity of the system by throttling down large heat pumps when their full capacity is too great for either heating or cooling requirements. (5) Use of high-efficiency natural gas-fired heat pumps as the energy source.

Enhanced Water-loop Heat Pump. Research by EPRI (Electric Power Research Institute) and others in a cooperative effort have developed an enhanced water-loop heat pump for heating and cooling large and medium-size commercial buildings. Claims for the new system include improved energy efficiency through inherent heat recovery, low first cost, zoning flexibility, simple control, and reduced space requirements.

The typical water-loop heat pump system (WHLP) is simple in concept, consisting of a pipe loop for circulating water and a *series* of heat pumps (one in each thermal zone) that use the piped water as a heat source or sink. The system also requires a means of removing heat from the pipe loop (typically a cooling tower) and a means of adding heat (typically a boiler).

The cooling tower and the boiler operate as necessary to keep the temperature of the water in the loop within a 60–90°F (15.6–32.2°C) range, allowing use of uninsulated piping, which significantly reduces installed costs. Because each heat pump can perform both heating and cooling, it is possible to use a two-pipe system rather than the usual four-pipe system, further cutting distribution system costs.

The WLHP efficiency is evident when a building requires simultaneous heating and cooling needs. See Fig. 2. The system has been tested in an office building in the northeastern part of the United States for over 3 years. As a test installation, the system has been subjected to severe research scrutiny. A number of options also have been designed into the basic system. (See EPRI reference listed.)

Additional Reading

Bevington, R., and A. H. Rosenfield: "Energy for Buildings and Homes," *Sci. Amer.*, 76 (September 1990).
Culp, A. W.: "Principles of Energy Conversion," 2nd Edition, McGraw-Hill, New York, 1991.
EPRI: "WLHP Information Brochure," Electric Power Research Institute, Palo Alto, California 1992.
Lucas, C. E.: "Refrigeration System Stability Linked to Compressor and Process Characteristics," *Chem. Eng. Progress*, 37 (November 1989).
Maili, A.: "Heat Pump for Distillation Columns," *Chem. Eng. Progress*, 60 (June 1990).
Staff: "Handbook of Fundamentals," Amer. Soc. Heating, Refrigeration and Air Conditioning Engineers, New York. (Revised periodically.)

HEAT STRESS, EXHAUSTION, AND STROKE. With ever-increasing emphasis on exercise and sports activities, there is a growing awareness of the effects of heat and exercise stress. The effects of heat stress can be serious and sometimes life threatening.

Weakness, mental fogginess, incapacity for work, and irritability are characteristics of heat exhaustion. These symptoms also appear in dehydration, alcoholism, and periods of insufficient rest and sleep. Heat exhaustion is produced in some persons when they are confined to an uncomfortably warm environment for an extensive period, during which time body fluids and salt may be depleted. The usual immediate symptoms are subnormal body temperature, clammy skin, gastic muscular spasms, and, less frequently, vomiting and diarrhea. Some authorities suggest that this syndrome may be the result of a sharp curtailment of heat production within the body, with a corresponding suppression of other functions of the adrenal cortex. Incidences of heat exhaustion can be prevented in many instances by curtailing vigorous physical exercise (as in the case of military training exercises in hot, dry or hot, humid areas) when the temperature exceeds 100°F (38°C). Another preventive measure is the scheduling of frequent rest periods in cooler locations where this is practical. See also accompanying table.

Heatstroke is a much more serious manifestation of similar factors and usually occurs when the body is subjected to very high temperatures and high humidities over relatively long periods. Epidemics of heatstroke occur in metropolitan areas during a heat wave, causing the

HEAT STRESS CHECKLIST

Recognizing Symptoms of Heat Stress

Heat Exhaustion

Dizziness, lightheadedness, fainting
Fatigue
Headache
General weakness
Gastrointestinal discomfort, nausea, and vomiting
Pale, moist skin
Heat cramps

Commonly encountered in sports activities, such as football. Result from muscular tightening and spasm occurring during or after prolonged exercise in a hot environment. Exquistely painful, usually involving larger muscles of the calf and thigh. Abdominal or stomach muscles also can be affected.

Heatstroke

Malfunction of the central nervous system
Aggressiveness or irritability
Restlessness or delirium
Confusion or disorientation
Muscular incoordination
Incoherent speech
Seizures
Hot, flushed, dry skin

Heat stress occurs when the heat produced by the body and heat transmitted to the body from a hot environment exceeds the body's ability to dispose of the heat. The human body is designed to function within a narrow internal temperature range. To maintain its operating temperature, the body must rid itself of the large amount of heat that working muscles produce. Heavy sports equipment, as encountered in football and other sports, increase muscular work and hence heat production. Such excess heat can cause death in 15 to 20 minutes if allowed to accumulate. Heat from muscles is transferred to the blood and thence removed in two ways:

1. Amount of blood flow to the skin increases. The "hot" blood can passively transfer heat to the environment, or
2. It can be used to evaporate sweat.

Passive heat transfer requires that the air temperature be lower than the skin temperature. If outside temperature is higher, the body will gain rather than lose heat. In hot weather, sweat evaporation is the main method of heat loss. To be effective, this method of cooling requires that (1) there must be adequate blood flow to transport heat to the skin, (2) the skin must be exposed to the air, and (3) the air must be able to absorb additional moisture (determined by humidity). Death from heatstroke has occurred at temperatures below 75°F (23.9°C) when humidity is 95%. Sweating is ineffective as a cooling medium when the sweat cannot evaporate. Sweat production during exercise in hot conditions can exceed 1 to 2 quarts (liters) per hour. With continued exercise, blood volume (50% water) drops and thus cannot furnish critical body needs. It is the inadequate supply of blood to the brain that produces the warning signals—headache, dizziness, nausea, and eventually unconsciousness and seizures.

High-risk Factors for Heat Stress

Alcohol consumption, including recovery period from excessive drinking.
Age, particularly the preteens and over 50 years.
Excess body weight loss (over 5% of body weight during a short period of minutes to a few hours).
Prior tendency toward heat stress problems.
Inadequate sleep.
Excessive body fat and decreased aerobic conditioning.
No prior heat acclimitization.
Recent fever or gastrointestinal illness.

Preventive Measures to Avoid Heat Stress

Adequate hydration

Thirst, alone, is an insufficient warning of dehydration. Prior to increased and vigorous exercise, one should practice "forced" drinking of water. As a general rule, athletes (football players, for example) should consume 12 to 20 ounces (0.4 to 0.6 l) of cool water or other dilute fluid prior to exercise and a minimum of 8 ounces (0.3 l) for every 15 to 20 minutes of active play. Players and coaches should monitor fluid consumption to assure that it is adequate.

Adequate body cooling

Clothing should be worn that assists the body's cooling mechanism. Changing sweat-soaked clothing improves sweat evaporation. Loose-fitting jerseys and shirts permit air to reach the skin. Low-cut socks are helpful. Head gear should be removed as often as practical because the blood supply to the head is high.

Fluid electrolytes

Normally, cool water is the best fluid replacement. However, specially prepared drinks for athletes are available to assure adequate replacement of sodium and other ions needed by the body. Salt usually is not necessary. Some drinks also incorporate glucose as an energy supplement, but some authorities observe that glucose tends to slow water absorption from the gut and may inhibit the body's normal cooling mechanisms.

Acclimitization

Vigorous exercise over long periods should be approached in a step-like manner to permit the body to adjust to an exercise program, rather than expose the body suddenly to thermal shock.

Drug avoidance

Heat intolerance is increased by medications, such as antihistamines, anticholinergics, beta-blockers, diuretics, thyroid preparations, and antidepressants.

deaths of hundreds of persons, particularly the elderly. Many such deaths go unreported and are not always identified with the cause. Poorly ventilated areas, such as barracks, sauna baths, and crowded facilities, as found in old nursing homes for the aged, aggravate the underlying conditions.

Heatstroke frequently may be manifested quite precipitously, with delirium, impaired senses, seizures, and coma. Heatstroke victims may be found with body temperatures as high as 104°F (40°C). Sweating is not always present. There may be extreme tachycardia, circulation may fail, and pulmonary edema and shock may result. Dehydration is often present. Most patients have vomiting and diarrhea.

Many deaths result from heatstroke because persons are not found and advised in time to initiate treatment. The initial step in treatment is removal of the person from the causative hot, humid environment. This should be followed by immersion in ice baths, application of ice packs, or sponging with alcohol in a relatively cool environment where there is good movement of air. Preferably with a rectal temperature sensor, the body temperature should be carefully monitored so that the patient will not be overcooled or allowed to reaccumulate risky heat loads. Phenothiazines may be administered to prevent excessive shivering or seizures, both of which conditions tend to increase body temperature. Water and electrolyte deficiencies should be corrected. The patient should be checked for possible renal failure and disseminated intravascular coagulation.

Information furnished for parts of this article by the Hughston Sports Medicine Foundation, Inc., Columbus, Georgia, is gratefully acknowledged.

HEAT TRANSFER. Although there are three generally accepted methods for transferring heat from one medium to another, or from one locale to another within a given medium, it is uncommon for one method to act unilaterally. Particularly where convection may predominate, some conduction of heat will be involved. In conduction, heat must diffuse through material substances; in convection, heat is essentially carried from one locale to another by actual movement of the transport medium; in radiation, heat transfer involves radiant wave energy.

Conduction. From a microscopic standpoint, thermal conduction refers to energy being handed down from one atom or molecule to the next one. In a liquid or gas, these particles change their position continuously even without visible movement and they transport energy also in this way. From a macroscopic or continuum viewpoint, thermal conduction is quantitatively described by Fourier's equation, which states that the heat flux q per unit time and unit area through an area element arbitrarily located in the medium is proportional to the drop in temperature, -grad T, per unit length in the direction normal to the area and to a transport property k characteristic of the medium and called *thermal conductivity*:

$$q = -k \operatorname{grad} T \tag{1}$$

Predictions for the value of the thermal conductivity k can be made from considerations of the atomic structure. Accurate values, however, require experimentation in which the heat flux q and the temperature gradient, grad T, are measured and these values are inserted into Fourier's equation. Thermal conductivity values for a number of media over a large temperature range are shown in Fig. 1. Metals have the largest conductivities and, among these, pure metals have larger values than alloys. Gases, in contrast, have very low heat conductivity values. Electrically nonconducting solids and liquids are arranged in between. The low thermal conductivity of air is utilized in the development of thermally insulating materials. Such materials, like cork or glass fiber, consist of a solid substance with a very large number of small spaces filled by air. The thermal transport occurs then essentially through the air spaces, and the solid structure only supplies the framework which prevents convective currents. It will be noted that the thermal conductivities indicated in Fig. 1 (at ambient temperature) extend through five powers of 10. This range is still small when compared with the range for the electric conductivity of various substances, where electric conductors have values which are larger by 25 powers of 10 than electric insulators. As a consequence, it is much easier to channel electricity along a desired path than to do so with heat, a fact which accounts for the difficulty in accurate experimentation in the field of heat transfer.

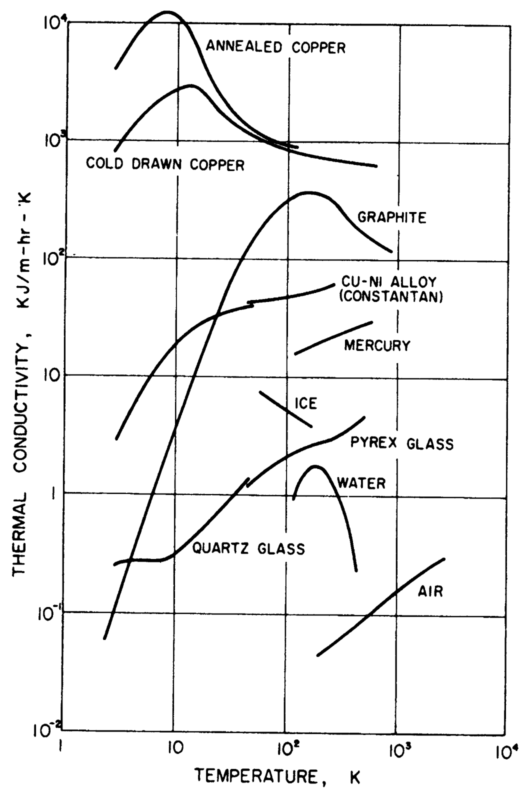


Fig. 1. Thermal conductivity values for a wide range of substances and over a temperature range of 1 to 10^4 K.

Fourier's equation can be used together with a statement on energy conservation to derive a differential equation describing the temperature field in a medium. Fourier was the first person to develop this equation and to devise means for its solution. In vector notation, this equation is:

$$\rho c = \frac{\partial \mathbf{T}}{\partial t} = \nabla (k \nabla \mathbf{T}) \tag{2}$$

where ρ is the density, c is the specific heat, t is time ,and ∇ is the Nabla (vector differential) operator. The temperature field in a substance can either change in time (unsteady state), or it can be independent of time (steady state, $\partial\mathbf{T}/\partial \mathbf{t} = 0$). For a steady-state situation, the temperature field depends primarily on the geometry of the body involved and on the boundary conditions. The simplest case of a steady-state temperature field is a plane wall with temperatures which are uniform on each surface, but different at the two surfaces. The temperature in the wall then changes linearly in the direction of the surface normal as long as the variation of the thermal conductivity in the temperature range involved can be neglected. For an unsteady process, the capacity of the medium to store energy enters the energy conservation equation; correspondingly, the specific heat of the material and its density become factors for the conduction process, as well as the thermal conductivity. A combination of these properties, defined as the ratio of the thermal conductivity to the product of specific heat and density, called *thermal diffusivity* ($k/\rho c$), then determines how fast existing temperature differences in a medium equalizes in time. It is found that metals and gases have thermal diffusivity values which are approximately equal in magnitude and are considerably higher than thermal diffusivities of liquid and solid nonconductors. This means that temperature differences equalize much faster in metals and gases than in other substances.

Various other physical processes lead in their mathematical description to equations of the same form as Eq. (2), especially in its steady-state form. Such processes include the conduction of electricity in a conductor, or the shape of a thin membrane stretched over a curved boundary. This situation has led to the development of analogies (electric analogy, soap film analogy) to heat conduction processes which are useful because they often offer the advantages of simpler experimentation.

Convection. When energy is transported by convection in fluids, conduction usually takes care of the transport of heat from one stream tube to another and is the dominating mode of transfer near solid walls. Convection transports heat along the stream lines and is dominating in the main body of the fluid where the velocities are large. In many situations, the flow is turbulent; this means that unsteady mixing motions are superimposed on the mean flow. These mixing motions contribute also to a transport of heat between stream tubes, a process which can be described by an "effective" conductivity which often has values by several powers of ten larger than the actual conductivity of the fluid.

Movement of the fluid may be generated by means external to the heat transfer process, as by fans, blowers, or pumps. It may also be created by density differences connected with the heat transfer process itself. The first mode is called *forced convection*; the second one *natural* or *free convection*. Convection heat transfer may also be classified as heat transfer in *duct flow*, or in *internal flow* (over cylinders, spheres, air foils, and similar objects). In the case of external flow, the heat transfer process is essentially concentrated in a thin fluid layer surrounding the object (boundary layer).

Of special interest in such heat transfer processes is the knowledge of the heat flux from the surface of a solid object exposed to the flow. This heat flux q_w per unit area and time is conventionally described by Newton's equation:

$$q_w = h(T_w - T_f) \tag{3}$$

where T_w is the surface temperature and Tf is a characteristic temperature in the fluid. This equation defining the heat transfer coefficient h is convenient because in many situations the heat flux is at least approximately proportional to the temperature difference $T_w - T_f$. Information on the heat transfer coefficients can be obtained by a solution of the Navier-Stokes equation describing the flow of a viscous fluid and the related energy equation, or they are found by experimentation. Computers enhance the ability to study heat transfer analytically at least for laminar flow, whereas in turbulent flow the bulk of the information is determined experimentally.

Experimentation is difficult because of the large number of parameters involved. Dimensional analysis has been applied to reduce the number of influencing parameters, and relations for convective heat transfer are correspondingly presented in many handbooks as relations between dimensionless parameters. Such an analysis demonstrates that heat transfer in forced flow can be described by a relation of the form.

$$Nu = f(Re, Pr) \tag{4}$$

in which the Nusselt number Nu is a dimensionless parameter hL/k, containing the heat transfer coefficient h, the Reynolds number $Re = \rho(VL/\mu)$ describes essentially the nature of the flow, and the Prandtl number $Pr = c_p\mu/k$ can be considered a dimensionless transport property characterizing the fluid involved. L and V are an arbitrarily selected characteristic length and velocity, respectively; ρ denotes the density, μ the viscosity, and C_p the specific heat of the fluid at constant pressure. See also **Reynolds Number.**

Convection is frequently thought of in terms of space heating and industrial heat-exchange processes. It should be pointed out that convection plays a cosmic role (in the sun's photosphere, for example), and a very large role in connection with the atmosphere of the earth and some other planetary bodies. For example, when normal convective transport is inadequate, temperature inversions occur and create smog hazards over large cities. See **Atmosphere (Earth).**

Attempts to develop a theory for convection date back at least to the 1790s when Thompson (Count Rumford) introduced the concept of heat convection. Very little theoretical work was undertaken, however, until the early 1900s, when Bénard (France) undertook experimental investigations. Modern convection physics stems from the work of Lord Rayleigh, who first published on the subject in 1916. In current times, advanced convection research studies have been undertaken by Velarde and Normand (1980), among others. See reference listed. See also **Boiler;** and **Heat.**

Radiation. In the transfer of energy from one location to another in the form of photons (electromagnetic waves), usually a multiplicity of wavelengths is involved. In vacuum, all waves regardless of their wavelength move with the same speed (2.9977×10^8 meters per second). In various substances, the wave velocity c changes somewhat with wavelength, and the ratio of the wave velocity in vacuum to the velocity in a substance is equal to the optical refraction index. Air and generally all gases have refractive indices which differ from one only in the fourth decimal. Their wave velocity is therefore practically equal to that in vacuum. See also **Waves and Wave Mechanics**.

Prévost's principle states that the amount of energy emitted by a volume element within a radiating substance is completely independent of its surroundings. Whether the volume element increases or decreases its temperature by the process of radiation depends upon whether it absorbs more foreign radiation than it emits or vice versa. One refers to thermal radiation when the emission of photons is thermally excited. i.e., when the substance within the volume element is nearly in thermodynamic equilibrium. For such radiation, Kirchhoff was able to derive a number of relations by consideration of a system of media in thermodynamic equilibrium. If jv indicates the coefficient of emission, i.e., the radiative flux at the frequency v^* emitted per unit volume into a unit solid angle, and η is the coefficient of absorption at the same frequency, i.e., the fraction of the intensity of a radiant beam which is absorbed per unit path length, then one of these relations states:

$$c^2 \frac{jv}{\eta_v} = f(T, v) \tag{5}$$

with c denoting the wave velocity. According to this relation, the combination of parameters on the left-hand side of Eq. (5) is a function of temperature T and frequency v of the radiation only, but does not depend upon the substance under consideration. Kirchhoff's law can also be expressed in parameters which refer to the interface of two media (1 and 2). It then takes the form:

$$c^2 \frac{i_v}{\alpha_v} = f(T, v) \tag{6}$$

in which i_v is the monochromatic intensity of the radiative flux at frequency v originating in medium 2 and traveling through the interface into medium 1 per unit solid angle and area normal to the direction of the radiant beam. α_v is the monochromatic absorptance or absorptivity, i.e., that fraction of a radiant beam approaching the interface in the medium 1 in the opposite direction that is absorbed in medium 2. The wave velocity in medium 1 is c. Kirchhoff's law states that the combination of the parameters on the left-hand side of Eq. (6) is again a function of temperature and frequency only, but does not depend upon the nature of the medium. A medium which absorbs all the radiation traveling into it through an interface ($\alpha_v = 1$) is called a *blackbody*. The intensity of radiation emitted by an arbitrary medium is, according to Eq. (6), in the following way related to the intensity of radiation i_{bv} emitted by a black body at the same temperature and frequency:

$$\frac{i_v}{v}\, i_{bv} \tag{7}$$

See also **Planck Radiation Formula.**

The amount of heat transferred by radiation can be determined by use of the *Stefan-Boltzmann law*:

$$Q = bA(T_1^4 - T_2^4) \tag{8}$$

where Q is the amount of heat transferred per unit time, b is a constant, A is the area of the radiating surface, T_1 is the absolute temperature of the radiating body and T_2 is the absolute temperature of the receiving body. Various correction factors are introduced into the formula to account for the shape of the bodies, their thermal radiation characteristics and the properties of the media through which the radiant rays must pass while traveling from radiator to absorber. The thermal radiation characteristics are its emissivity, a measure of its ability to radiate at a given temperature, its absorptivity, a measure of its ability to absorb heat and its reflectivity, which measures its ability to reflect without absorbing.

Radiant energy travels in a straight line. Therefore to transmit it to an object out of sight of the radiator requires a reflector, such as a furnace wall, to deflect the rays to their objective.

It is possible to set up controlled laboratory radiation between simple plane surfaces and determine therefrom accurate coefficients to incorporate into radiation equations. However, the radiation of heat from furnace gases, consisting of non-luminous gases, luminous carbon particles in flame, ash globules, etc., to the walls and tubes of a steam generator in commercial operation at variable load, is another matter. Here, empirical data which are gathered and interpreted from field tests on similar equipment, must still be resorted to however great the designer's urge to go back to basic laws of heat transfer.

Radiant heat transfer in furnaces is roughly proportioned to the difference in the fourth power of the absolute temperatures of the radiating and receiving surfaces. The water wall surface is approximately at boiler saturation temperature, while the superheater surface varies from this to somewhat above the temperature of the steam at the superheater outlet. However, the mean radiating temperature of the furnace gases is usually over 1204°C. The fourth power of the receiving surface temperature is thus seen to be small compared to the fourth power of the transmitting surface temperature; consequently the latter controls the transmittance, and boiler tube temperature does not need to be considered a variable to be accounted for.

Figure 2 shows some of the arrangements in which radiant heat-absorbing surface is disposed. It may be used to illustrate another of the difficulties which beset the designer in following a rational or semi-rational form of radiation analysis. Projected radiant surface is one thing; actual radiant energy receiving surface may be quite a different area. For example, suppose the tubes of case (a) to be separated and spaced l_1 inches on centers. The *projected* areas of cases (a) and (c) would then be the same, but it seem obvious that re-radiation from the wall causes more of a (c) tube to receive radiant energy than is the case with an (a) tube. Also, if δ is a factor correcting projected area to *equivalent* absorbing surface, what value should be assigned to it in the case of a bank of tubes which may receive by re-radiation some radiant energy deep in the tube bank? Here δ has a minimum value of 1, but some investigators have derived expressions which indicate that δ may have a magnitude of 3 or more.

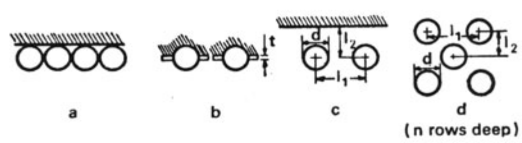


Fig. 2. Arrangements of radiant heat-absorbing surface.

Industrial Heat Transfer Equipment

Some of the more common cases of industrial heat transfer are:

1. Radiation from fuel beds and luminous gases to absorptive surfaces such as boilers, cylinder walls, etc.
2. Radiation from heat generators such as drying lamps.
3. Convection of heat out of combustion regions.

4. Convection of heat from hot surfaces under either free or forced convection.
5. Conduction of heat through the tubes of boilers, heaters, heat exchangers, condensers, etc.
6. Conduction in walls, pipe covering, and other so-called "heat insulators."
7. Conduction of heat through the plates of plate-type heat exchangers and regenerators.

Heat exchangers perform many functions within a manufacturing facility. Often they are given special names, even though they remain fundamentally heat exchangers. These include:

Chiller—a device which cools fluids to temperature below those obtainable with ordinary cooling water by using the vaporization of a refrigerant. The fluid to be cooled is routed through the tubes while the low-boiling refrigerant vaporizes from a pool of liquid in the shell.

Partial Condensers—Many overhead vapors from distillation columns in petroleum-refinery services are a mixture of light and heavy hydrocarbons and noncondensable gases, i.e., gases that are not condensed at the outlet temperature and pressure of the condenser (air, hydrogen sulfide, methane, and other light ends). These vapors are routed through the shell side while water is used as the cooling medium on the tube side of the unit. Condensation on the shell side begins at the saturation temperature of the heavy components and continues over a decreasing temperature range until part of the lighter components are condensed. Part of the existing liquid is sent back to the tower as reflux, while the remainder is further refined or passes to the trim cooler and storage.

Trim Cooler—This unit condenses the last remaining light-end vapors and cools the liquid to the ultimate storage temperature (often about 100°F: 38°C) by using cooling water. This cooling usually is not conducted in the main condenser because it would reduce column pressure.

Thermosiphon Reboiler—Flow of the vaporizing fluid depends upon the difference in static head between the column of liquid flowing from the tower to the reboiler and the partially vaporized column of liquid returning from the exchanger to the tower.

Reboilers—These exchangers operate in conjunction with a distillation tower to vaporize enough liquid to assure vaporization of the overhead product. A hot process stream of steam may be used as the heating medium. Most reboilers are shell-and-tube exchangers located at the base of the tower. The vaporizing fluid is routed through the shell side of the exchanger.

Forced-circulation Reboiler—A pump is used to provide more positive circulation than available with the thermosiphon effect, e.g., in the vaporization of viscous fluids.

Vapor Heat Exchanger—Units of this type preheat a cool stream of process fluid by using heat from partially condensing vapor. The objective is to conserve heat and eliminate the requirement for a separate preheater.

Air-cooled Exchanger—As used in the petroleum industry, air-cooled exchangers normally comprise two headers joined by a horizontal bank of finned tubes. Usually two motor-driven fans located above (induced draft) or below (forced draft) the tubes are used to circulate the air over the finned surface.

Superheater—A unit of this type heats vapor above the saturation temperature.

Waste-heat Boiler—A unit of this type generates steam and is similar to a regular steam generator except that hot gas or liquid produced by a chemical reaction (often combustion) is the heating medium.

Types of Heat Exchangers. In terms of heat exchange for recovering and recycling thermal energy, the shell-and-tube heat exchanger of the type shown in Fig. 3 for many decades has been the most common type. It can be used with liquid on both sides, gas on both sides, or liquid on one side and gas on the other side. The most common requirement is for liquid-liquid exchangers. Heat exchanges may be used strictly for processing purposes—that is, materials need to be heated (or cooled) prior to entering some processing application, such as reacting, distilling, vaporizing, and the like. Or heat exchangers may be used simply for recovering the energy from hot fluids for use elsewhere.

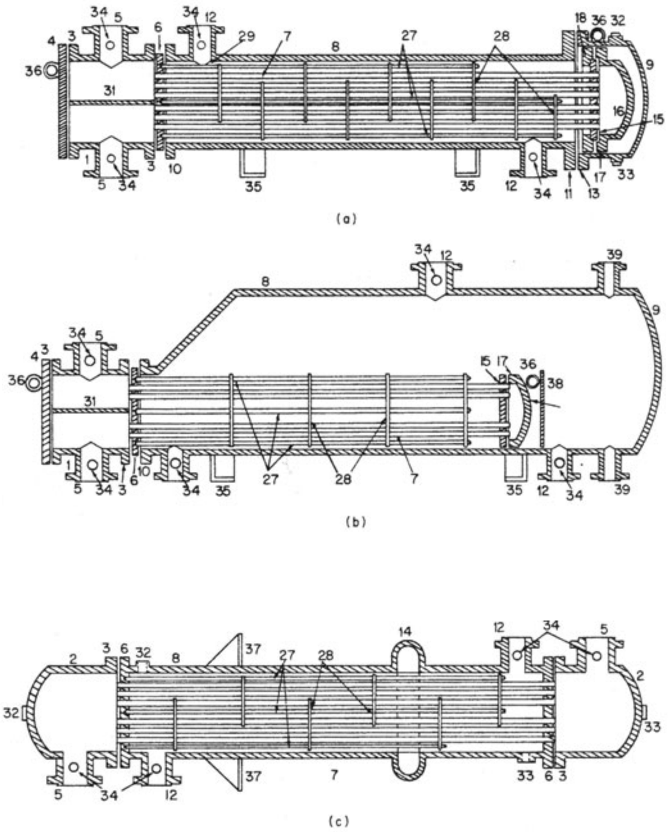


Fig. 3. Three common types of shell-and-tube heat exchangers: (a) Type AES, internal-floating-head exchanger (with floating-head backing device); (b) Type AKT, kettle-type floating-head reboiler; and (c) Type BEM, fixed-tube sheet exchanger. (*Sketches adapted from specification diagrams from the Standards of Tubular Exchange Manufacturers Association.*) Legend: (1) Stationary head, channel, (2) stationary head bonnet, (3) stationary-head flange, channel, or bonnet, (4) channel cover, (5) stationary-head nozzle, (6) stationary tube sheet, (7) tubes, (8) shell, (9) shell cover, (10) shell flange, stationary-head end, (11) shell flange, rear-head end, (12) shell nozzle, (13) shell cover flange, (14) expansion joint, (15) floating tube sheet, (16) floating-head cover, (17) floating-head flange, (18) floating-head backing device, (19) split shear ring, (20) slip-on backing flange, (21) floating-head cover, external, (22) floating tube sheet skirt, (23) packing-box flange, (24) packing, (25) packing follower ring, (26) lantern ring, (27) tie rods and spacers, (28) transverse baffle, (29) impingement baffle, (30) longitudinal baffle, (31) pass partition, (32) vent connection, (33) drain connection, (34) instrument connection, (35) support saddle, (36) lifting lug, (37) support bracket. (38) weir, and (39) liquid-level connection.

Although immensely improved over the years from the standpoint of design efficiency, resistance to corrosion, and ease of maintenance, among other objectives, the fundamental design has remained unchanged. However, within the past few years, the plate-and-frame heat exchanger has been introduced. See Fig. 4. This type of exchanger consists of a frame that carries a series of closely spaced metal plates that have been pressed, with a corrugated trough pattern. The plates, which are clamped between a fixed head and movable follower, have corner ports to permit the passage of process and service liquids. There are elastomeric gaskets around the ports and plate edges to avoid leakage. The plates are grouped into passes witin the heat exchanger. The product and service fluids flow countercurrent to each other between the parallel passages in each pass. Initially, plate-and-frame exchangers were used for liquid-liquid thermal exchange purposes. Increasingly, they are finding use in condensing and boiling applications, where their compact size and thinner material requirements for wetted parts offer advantages over other types of heat-exchange designs.

Less commonly used are the heat-transfer configurations, as shown in Figures 5 through 9.

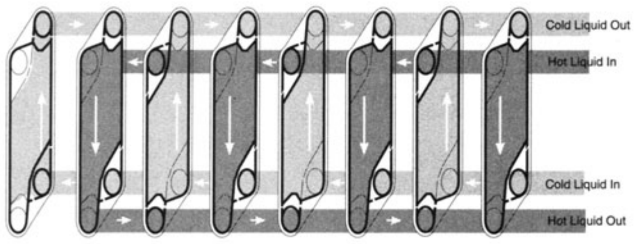


Fig. 4. Principle of plate-and-frame heat exchanger. (*After Carlson.*)

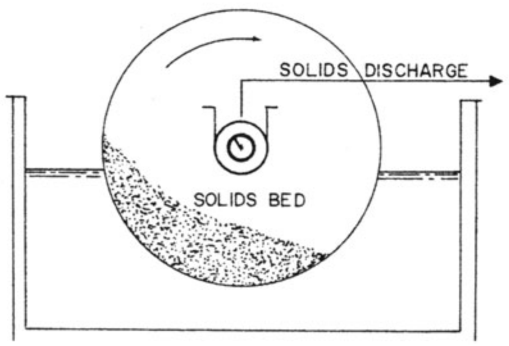


Fig. 5. Plain rotating shell used for both heating and cooling. For high-range heating, tempered combustion gases may be used instead of water.

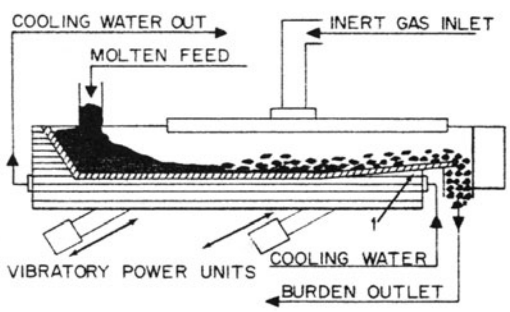


Fig. 6. Vibrating-type heat-transfer equipment for batch solidification. Sometimes referred to as a *caster,* the machine is used widely in a number of industries. After cooling and solidification, intense vibratory action shatters cake into lumps.

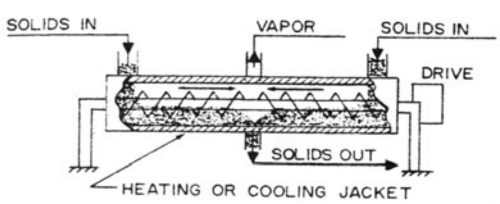


Fig. 7. Tank equipped with mixing ribbon spirals provides considerable agitation and is useful for melting or cooking dry powdered solids. Heat-transfer efficiency is only moderate because of the relatively deep beds of solid particles.

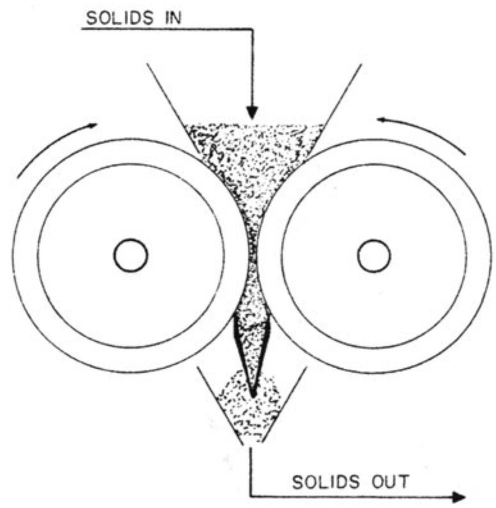


Fig. 8. Double drum. Scraping knives may be engaged continuously or intermittently, depending upon the nature of the heated product.

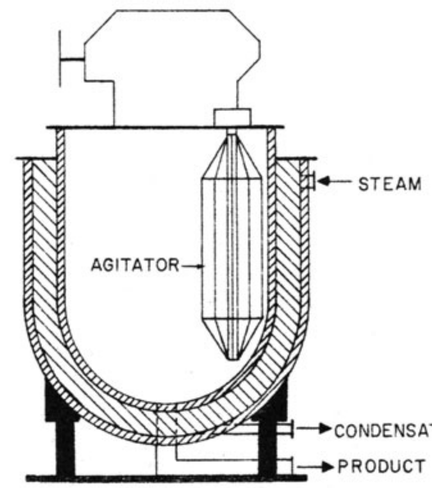


Fig. 9. Vertical agitated kettle. Although heat transfer through the jacket normally is quite poor, a kettle offers convenience in handling and cleaning and is particularly useful where batches of different materials must be processed frequently.

Heat Storage

It is often necessary to store heat in rather large quantities in specially designed apparatus. Hot water, of course, is one of the easiest forms in which to store thermal energy that is immediately available. As contrasted with hot water, electric energy and steam have to be generated on an as-needed basis. The blast furnace poses a difficult heat storage problem which obviously cannot be handled by storing heat in water. Great amounts of hot gas are required on a cyclic basis. To heat such quantities of air on a continuous, as-required, basis would be quite impractical with the present stage of the art. The solution used involves several stoves which are quite large, often over 100 feet (30 meters) in height and about 25 feet (7.5 meters) in diameter. The blast temperature of approximately 1,000°F (538°C) is accomplished by preheating the stove checkerwork to a much higher temperature. Checkerwork is comprised of refractory material forms constructed in high walls in checkerboard fashion to permit free passage of air through the interstices when under pressure. The gas passing through the stove exhausts initially at 2000°F (1093°C). Mixing this with unheated air produces the required blast temperature for the blast furnace. The stoves usually are heated for a period of three hours and exhaust (termed "on wind") for a period of about one hour. See Fig. 10. A similar system of checkerwork regenerators is used in connection with glass-tank heat-storage systems.

Flowing streams of pebbles also have been used in the chemical industry for removing heat from gases. Pebbles and stones are also used in some solar energy storage systems. See Fig. 11. See also **Solar Energy**.

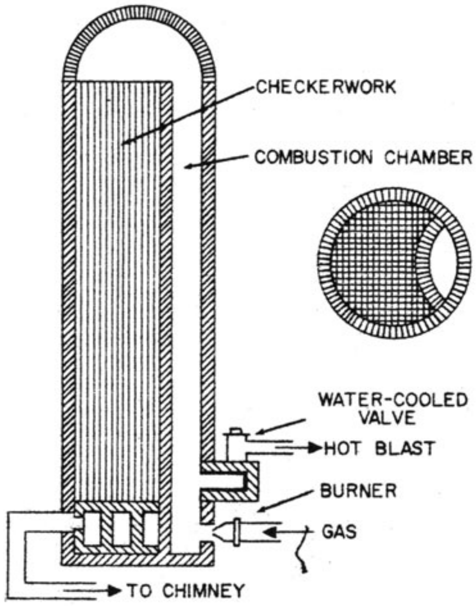


Fig. 10. Blast furnace stove for preheating large quantities of air.

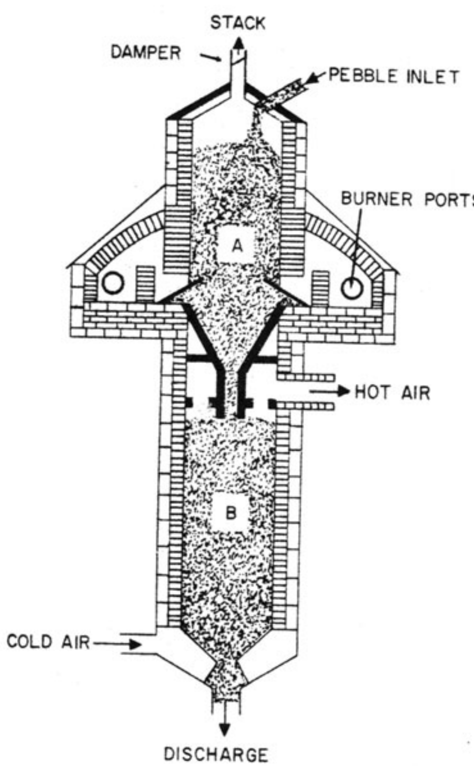


Fig. 11. Pebble heater for heating steam to temperatures impractical in metallic units. Also used for heating air, hydrogen, methane, and other gases for processing purposes. In reverse, a pebble heater may be used to recover heat from hot gases. The pebbles are heated in top chamber *A* by direct contact with combustion gases and passed through a throat to lower chamber *B,* where heat is transferred to cool gases. The two chambers are maintained at the same temperature so that there will be no gas flow between them. An average cycle on the pebbles is 30–50 minutes.

Additional Reading

Butterworth, D., and C. F. Mascone: "Heat Transfer Heads Into the 21st Century," *Chem. Eng. Progress*, 30 (September 1991).
Carlson, J. A.: "Understand the Capabilities of Plate-and-Frame Heat Exchangers," *Chem. Eng. Progress*, 26 (July 1992).
Corsi, R.: "Specify Bayonet Heat Exchangers Properly," *Chem. Eng. Progress*, 32 (July 1992).
Ganapathy, V.: "Heat-Recovery Boilers: The Options," *Chem. Eng. Progress*, 59 (February 1992).
McKetta, J. J., Editor: "Heat Transfer Design Methods," Marcel Dekker, Inc., New York, 1992.
Mukherjee, R.: "Use Double Segmental Baffles in Shell-and-Tube Heat Exchangers," *Chem. Eng. Progress*, 47 (November 1992).
Perry, R. H., and D. W. Green: "Perry's Chemical Engineers' Handbook," 6th Edition, McGraw-Hill, New York, 1984.
Someah, K.: "On-Line Tube Cleaning: The Basics," *Chem. Eng. Progress*, 39 (July 1992).
Wood, R. M., et al.: "A New Option for Heat Exchanger Network Design," *Chem. Eng. Progress*, 38 (September 1991).
Yokell, S. A.: "A Working Guide to Shell-and-Tube Heat Exchangers," McGraw-Hill, New York, 1990.

HEAT TRANSFER (Nusselt Number). See **Nusselt Number.**

HEAT TREATING. Heating and cooling of metals to effect changes in properties. Annealing and normalizing are generally for the purpose of softening or improving the grain structure. Patenting is also a softening process in which cold drawn carbon-steel wire is heated above its critical temperature range followed by cooling to below this range in a molten lead or molten salt bath, with subsequent cooling to room temperature.

While heat treating includes the softening treatments, it most often implies hardening and strengthening. In the case of steels this requires heating to above the critical temperature range followed by rapid cooling (quenching) in oil, water, or brine, except in the case of special grades which harden on cooling in air. This is followed by tempering, a low-temperature reheating treatment which reduces the internal stresses caused by the hardening treatment. Tempering may be carried to a high enough temperature to reduce somewhat the extreme hardness of the as-quenched steel and increase the toughness and ductility, depending on the requirements of the part. See **Iron Metals**, **Alloys**, and **Steels**.

Another important form of heat treatment for hardening is precipitation hardening. See also **Annealing**; **Carbonitriding**; **Carburizing**; **Case Hardening**; and **Nitriding**. See Table 1.

Thermal or heat treating has been an inherent part of metalworking and fabricating for well over a century, and, in fact, some aspects of the topic date back to ancient times. Within the last few decades, heat treating has become quite sophisticated through the incorporation of modern computing and modeling techniques. Process modeling can be used as a scheduling tool to optimize throughput of a continuous furnace, for example. Much more instrumentation has been added to heat-treating processes, allowing better control over a larger number of variables that affect final product quality.

As of the early 1990s, one of the most interesting new processes is the use of solar energy as a direct heat source for surface hardening, cladding, and other surface modifications. An impressive demonstration project has been established at the Solar Energy Research Institute (SERI) in Golden, Colorado. Its heliostat has an area of 31.8 square meters (342 sq ft) and features an ultraviolet (UV)–enhanced aluminum coating on its front surface. The primary concentrator consists of 23 hexagonal facets, each of which is a spherical mirror ground to a 14.6-meter (48-ft) radius of curvature and is aluminum coated. At the target, 94% of the energy falls inside a 100 millimeter (4-in–diameter) circle. The beam has a Gaussian shape, with a peak flux of 2.5 MW/m^2, without a secondary concentrator.

A solar facility requires a major resource of direct normal radiation. Thus, in the United States, a facility of this type is limited to locations in Arizona, Colorado, Nevada, New Mexico, and Utah, or in a nearby region of one of these bordering states. Obviously, the facility cannot operate at night or during periods of dense cloud cover. Even with these limitations, design calculations show that a solar furnace can compete economically with laser and arc-lamp sources. The case for solar-furnace technology becomes even more attractive for materials-processing applications in space.

Currently, high-flux solar facilities are comparatively few, as shown in Table 2.

TABLE 1. PRINCIPAL HEAT-TREATING PROCESSES

Carburizirg

Pack and gas carburizing create a diffused carbon case. Base metals are low-carbon steels and low-carbon alloy steels. Process temperature range is 815–980°C (1500–2000°F).

Liquid carburizing creates a diffused carbon (possibly nitrogen) case. Base metals are low-carbon steels and low-carbon alloy steels. Process temperature range is 815–980°C (1500–1800°F).

Vacuum carburizing creates a diffused carbon case. Base metals are low-carbon steels and low-carbon alloy steels. Process temperature range is 815–1090°C (1500–2000°F).

Nitriding

Gas nitriding creates a diffused nitrogen (nitrogen compounds) case. Base metals are alloy steels, nitriding steels, and stainless steels. Process temperature range is 480–590°C (900–1100°F).

Salt nitriding creates a diffused nitrogen (nitrogen compounds) case. Base metals are ferrous metals, including cast irons. Process temperature range is 510–565°C (950–1050°F).

Ion nitriding creates a diffused nitrogen (nitrogen compounds) case. Base metals are alloy steels, nitriding steels, and stainless steels. Process temperature range is 340–565°C (650–1050°F).

Carbonitriding

Gas carbonitriding creates a diffused carbon and nitrogen case. Base metals are low-carbon steels, low-carbon alloy steels, and stainless steels. Process temperature range is 760–870°C (1400–1600°F).

Liquid (cyaniding) creates a diffused carbon and nitrogen case. Base metals are low-carbon steels. Process temperature range is 760–870°C (1400–1600°F).

Ferric nitrocarburizing creates a diffused carbon and nitrogen case. Base metals are low-carbon steels. Process temperature range is 565–675°C (1050–1250°F).

Aluminizing

Creates a diffused aluminum case. Base metals are low-carbon steels. Process temperature range is 870–980°C (1600–1800°F).

Siliconizing (chemical vapor deposition).

Creates a diffused silicon case. Base metals are low-carbon steels. Process temperature range is 925–1040°C (1700–1900°F).

Chromizing (chemical vapor deposition).

Creates a diffused chromium case. Base metals are low- and high-carbon steels. Process temperature range is 980–1090°C (1800–2000°F).

Titanium carbide

Creates a diffused carbon, titanium, and TIC case. Base metals are alloy and tool steels. Process temperature range is 900–1010°C (1650–1850°F).

Boriding

Creates a diffused born (boron compounds) case. Base metals are alloy and tool steels; cobalt- and nickel-base alloys. Process temperature range is 400–1150°C (750–2100°F).

TABLE 2. HIGH-FLUX SOLAR FACILITIES WORLDWIDE

Location	Total Power kW	Peak Flux MW/m^2
Alberquerque, New Mexico		
(Central receiver test facility)	5000	2.4
Furnace	22	3.0
Atlanta, Georgia		
Furnace	1.3	9.5
Golden, Colorado	10	2.5
White Stands, New Mexico	30	3.6
Odello, France		
Horizontal furnace	1000	16.0
Vertical furnace	6.5	15.0
Rehovot, Israel		
(Central receiver test facility)	2900	—
Furnace	16	11.0
Uzbek, Russia	1000	17.0

Source: Solar Energy Research Institute, Golden, Colorado.

Additional Reading

Coffey, J. A.: "Supercell Carburizing," *Adv. Mat. & Proc.*, 81 (September 1989).
Conybear, J. G.: "Advanced Controls Offer Heat-Treating Flexibility," *Adv. Mat. & Proc.*, 38 (October 1989).
Conybear, J. G.: "Gas-Fired Vacuum Furnace Speeds Ion Nitriding," *Adv. Mat. & Proc.*, 87 (September 1991).
Dekumbis, R.: "Surface Treatment of Materials by Lasers," *Chem. Eng. Progress*, 23 (December 1987).
Doak, K. W.: "Furnaces Focus on New Processes, New Materials," *Adv. Mat & Proc.*, 84 (September 1989).
Holm, T.: "Synthetic Heat-Treating Atmospheres," *Adv. Mat. & Proc.*, 45 (October 1989).
Jones, L. E., and E. D. Jamieson: "Computers Tackle Heat-Treating Problems," *Adv. Mat. & Proc.*, 33 (March 1990).
Krauss, G.: "Thermal Processing of Steel," *Adv. Mat & Proc.*, 57 (January 1990).
Moerdijk, I. W.: "Polymer Quenchants," *Adv. Mat. & Proc.*, 19 (March 1990).
Persampieri, D., San Roman, A, and P. D. Hilton: "Process Modeling for Improved Heat Treating," *Adv. Mat. & Proc.*, 19 (March 1991).
Smidt, F. A.: "Surface Modification," *Adv. Mat. & Proc.*, 61 (January 1990).
Stanley, J. T., Fields, C. L., and J. R. Pitts: "Surface Treating (Solar)," *Adv. Mat. & Proc.*, 16 (December 1990).
Totten, G. E.: "Polymer Quenchants: The Basics," *Adv. Mat. & Proc.*, 51 (March 1990).

HEAT UNITS. See **Units and Standards.**

HEAVISIDE LAYER. See **Ionosphere.**

HEAVY HYDROGEN. See **Deuteron.**

HEAVY WATER REACTOR. See **Nuclear Power Technology.**

HEEL. The prominence at the posterior end of the foot. It is based on the projection of one bone, the calcaneum, behind the articulation of the bones of the lower leg. In the long-footed mammals, both the hoofed species and the clawed forms which walk on the toes, the heel is well above the ground at the apex of the angular joint known as the hock or hough. In plentigrade species it rests on the ground.

HEISENBERG FORCE. The Heisenberg force is a phenomenologically postulated force between two nucleons derivable from a potential in which there appears an operator which exchanges the spins and positions of the two particles.

HEISENBERG REPRESENTATION. Representation of the equations of motion in quantum mechanics and in quantized field theory where the vector describing the state is treated as constant and the time dependence is transferred to the operators which operate on this state vector. This may be represented in Hilbert space by keeping the state vector constant and allowing the axes to rotate with time as the motion of the system develops. Matrices representing operators referred to these axes are thus time dependent and obey the Heisenberg equation of motion. The theory developed in this representation is therefore called matrix mechanics.

HELIARC WELDING. See **Welding.**

HELICAL GEARING. For high pitch-line velocities and heavy loads, some form of "twisted tooth" gear is generally used. Two important types are helical gears and double-helical or herringbone gears. Both helical and herringbone gears are essentially spur gears with teeth twisted across the face in the form of a helix about the axis of rotation.

When spur gear teeth engage, the contact extends across the entire tooth on a line parallel to the axis of rotation, and may result in noise and shock at high speeds. In helical gear engagement, contact begins at one end of the entering tooth and gradually extends along a diagonal line across the tooth face as the gears rotate. The nature of the contact is such that with sufficient face width, two or more teeth are in contact and are carrying the load at all times. Helical gears are therefore used for transmission ratios as high as 10:1, and at pitchline velocities up to 2,000 feet (610 meters) per minute for commercially-cut units. Herringbone gear sets of special design have been successfully operated at pitch-line speeds of 12,000 feet (3,660 meters) per minute or above.

Tooth elements of helical gears are similar to those of spur gears. (See accompanying figure). The *helix angle H* of the tooth is measured between the line tangent to the tooth helix at the pitch circle and the shaft axis. In any pair, the gears have teeth with mating right-hand and left-hand helices. The usual method of tooth measurement is by diametral pitch P_d, which corresponds to circular pitch P_c in the diametral plane, perpendicular to the axis of rotation. By using standard pitches, the pitch diameters (and therefore the center distance of helical gear sets) can be given in commonly used fractions or integers; consequently, a spur gear set of a certain size can be replaced directly by a similar helical gear set. Actual tooth thickness depends upon the pitch and the size of the helix angle; if the circular pitch P_c be held constant, the actual tooth thickness measured perpendicular to its elements will decrease as the helix angle H is increased. A different cutter is required for every change in helix angle, although the pitch may remain constant. To eliminate an extensive variety of cutters, commercially available helical gears are made in several standard helix angles, among which are 7° 30′, 15°, and 23°.

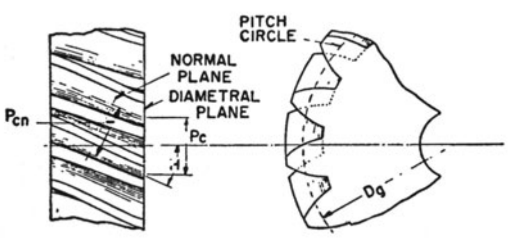
By using a standard pitch in a plane normal to the tooth helix, the pitch diameter of a helical gear can be varied to suit a particular center distance by changing the helix angle. In this method of tooth measurement, the normal diameter pitch P_n corresponds to a normal circular pitch P_{cn} in the normal plane. Helical gear teeth designed with normal diametral pitches may be cut with standard spur gear cutters or hobs.

The pitch diameter D_g of a helical gear, based upon normal pitch P_n, is given by

$$D_g = N_g/P_n \cos H$$

where N_g is the number of teeth in the gear. The power transmitting capacity of helical and herringbone gears may be found by methods analogous to those used for spur gearing.

End thrust inherent in single helical gears can be eliminated by the use of herringbone gears that consist virtually of two integral single helical gears of opposite hand, which absorb the axial thrust within the gear. Herringbone gears are used for hoisting and mining machinery, rolling mills, sugar mill and lumber machinery, turbine and compressor drives.



Tooth elements of helical gear.

HELICOPTERS AND V/STOL CRAFT. Although both fixed- and rotary-winged aircraft use airfoils to produce lift, in fixed-wing craft the wings can move no faster than the fuselage to produce lift and in order to fly; the whole aircraft must maintain considerable forward speed at all times. In the helicopter the wings (called rotor blades) are rotated at high speed, and there is no relationship between blade speed and fuselage speed. The helicopter, employing one or more horizontal rotors to give both lift and translation, can rise and descend vertically from the ground, hover over a spot on the ground, and fly backward and sideward as well as forward. With these flight characteristics, the helicopter does not require a prepared runway or landing area. A clearing about the size of a tennis court is adequate for landing, even though the surface be rough or uneven.

V/STOL aircraft are of more conventional lines. V/STOL is an abbreviation for a vertical or short take-off and landing. VTOL signifies vertical takeoff and landing.

Rotary-Wing Aerodynamics

The aerodynamics of rotary-wing and fixed-wing aircraft are basically the same. See also **Aerodynamics.** Both types of aircraft employ airfoils to produce lift; and both are subjected to identical fundamental forces of lift, drag, thrust, and gravity. It is true, however, that the flight characteristics of the helicopter differ widely from those of the fixed-wing craft.

Lift. Weight and lift are closely associated inasmuch as weight tends to pull the helicopter down and lift acts to hold it up. The similarity between the fixed-wing airplane and the helicopter is apparent; both are heavier than air, and both are sustained in flight by reaction of airflow over airfoils. The helicopter's airfoils are rotor blades which are turned at high speed. In a fixed-wing aircraft, if the angle of attack is increased, lift is increased until the stalling angle is reached; and for a given angle of attack, the greater the speed, the greater the lift. The helicopter rotates the rotor blades at high speed in rpm, to establish high-speed airflow over the airfoils. It is normal for the tip speed of the rotor blades to be as much as 350 miles (563 kilometers) per hour when the speed of the fuselage is zero. This explains why the helicopter does not require forward speed to produce lift, and why it can hover or fly backward, sideward, or forward.

Airflow. During normal operation conditions, the direction of airflow is from the top down through the main rotor system. As the blades are rotated with a positive angle of attack, they, in effect, screw upward into the air; thus a downwash of air (Fig. 1) is established through the rotor system. Notice that the leading edge of each blade bites into air throughout the complete cycle of rotation, forcing the air downward.

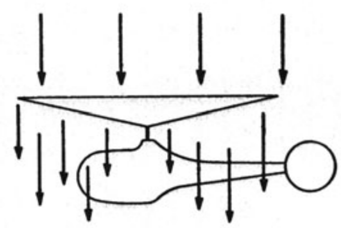


Fig. 1. Downwash through the rotor system.

At the root of the blade, airflow is slightly more than zero, but the velocity progressively increases throughout the length of the blade and at the tip may be 350 miles (563 kilometers) per hour or higher. It is the blade velocity that determines the resultant strength and direction of the relative wind at a positive angle of attack. The helicopter changes the angle of attack by varying the pitch of the main rotor blades. In a helicopter, the relative wind is developed throughout the complete cycle of 360 degrees by rotation of the rotor system, and it usually varies considerably. This variation is dependent upon flight conditions. See Figs. 2 through 6.

Angle of Incidence. This is the angle formed by the chord of the airfoil and the longitudinal axis of the aircraft. The conventional airplane's angle of incidence is built into the aircraft by the designer and

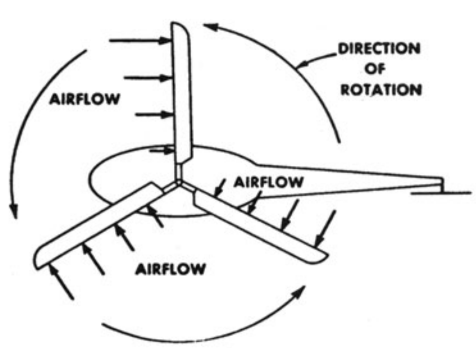


Fig. 2. Airflow in the rotor system.

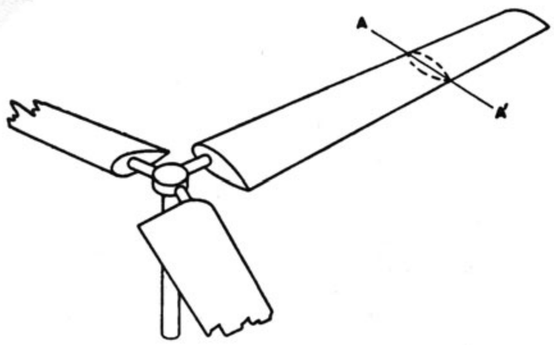


Fig. 3. High-speed blade section.

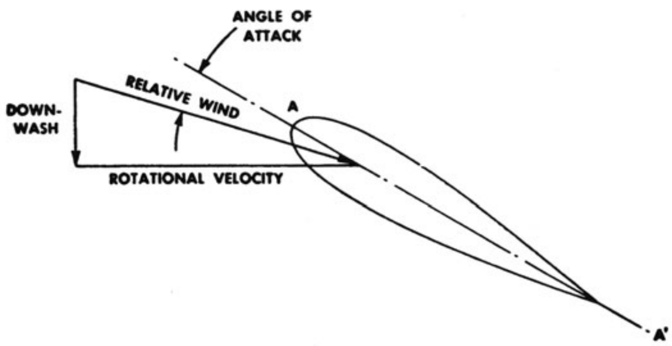


Fig. 4. Relative wind components.

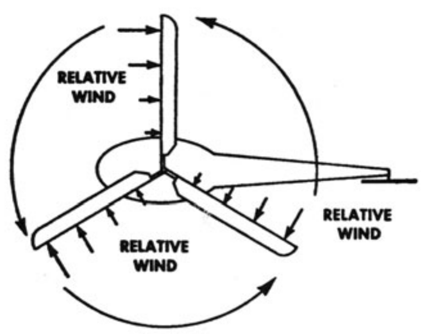


Fig. 5. Direction of relative wind.

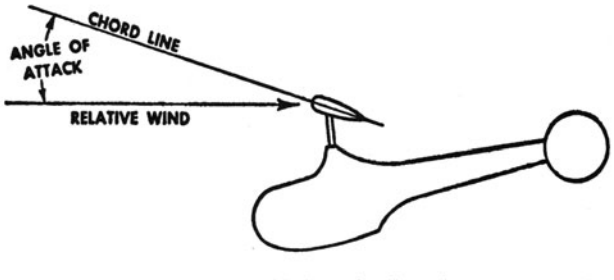


Fig. 6. Rotor-blade angle of attack.

in most aircraft cannot be changed. In the case of the helicopter, however, the pilot continually changes the angle of incidence during flight by increasing or decreasing the pitch of the main rotor blades. See Fig. 7.

Airfoil Section. The type of wing used on conventional airplanes varies considerably; the airfoils may be symmetrical or unsymmetrical, usually dependent upon some specific requirements. The unsymmetrical airfoil may be efficient for an airplane wing, but it has one disad-

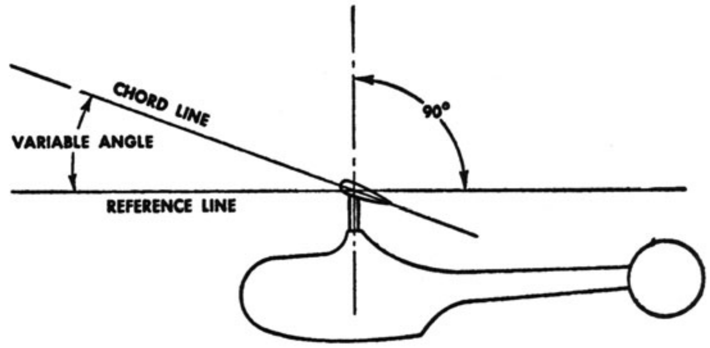


Fig. 7. Rotor-blade angle of incidence.

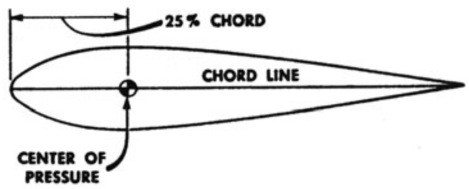


Fig. 8. Airfoils.

vantage that makes it unsatisfactory for use as a rotor blade. It is normal for the center of pressure to "walk" forward and rearward as the angle of attack is changed. The center of pressure is the imaginary point on the airfoil where all the aerodynamic forces are considered to be concentrated. See Fig. 8.

The airfoil section used for rotor blades is symmetrical, having equal camber above and below the chord line. Normally the greatest thickness of the blade is at a point about one-fourth of the way back from the leading edge. It is at this point that the center of pressure is located. There are several reasons for using the symmetrical airfoil for rotor blades: (1) There is a restricted migration of the center of pressure on a symmetrical airfoil when the angle of attack is changed; (2) the lift-drag ratio is very good even though the velocity of the blade varies from root to tip; and (3) the symmetrical airfoils permit ease of construction. If the center of pressure were permitted to travel during angle of attack variations, pitching moments would be introduced into the rotor system; this condition would set up violent vibrations. Good lift-drag ratio throughout a wide range of velocities is important because it is necessary to have the lift forces spread over a wide area in order to equalize stresses. Usually a slight twist is built into the blade to help equalize these forces.

Thrust and Drag. As weight and lift are closely associated, so are thrust and drag. Thrust moves the helicopter in a designated direction and drag tends to hold it back. The helicopter develops both lift and thrust in the main rotor system. In vertical ascent, thrust acts upward in a vertical direction; drag, the opposing force, acts vertically downward. Lift sustains the weight of the helicopter; and excess thrust is available to give translation or vertical acceleration. During vertical ascent, drag is considerably increased by the downwash of the main rotor system striking the fuselage. Thrust must be sufficient to overcome both drag and downwash. The force representing the total reaction of the airfoils with the air is divided into two components: (1) Lift, and (2) thrust. However, drag is a separate force from weight, as shown in Fig. 9.

At all times, the lift forces of the rotor system are perpendicular to the tip-path plane. The tip-path plane is the imaginary plane described

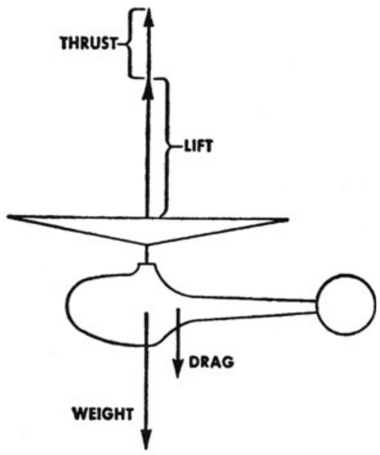


Fig. 9. Forces in vertical ascent.

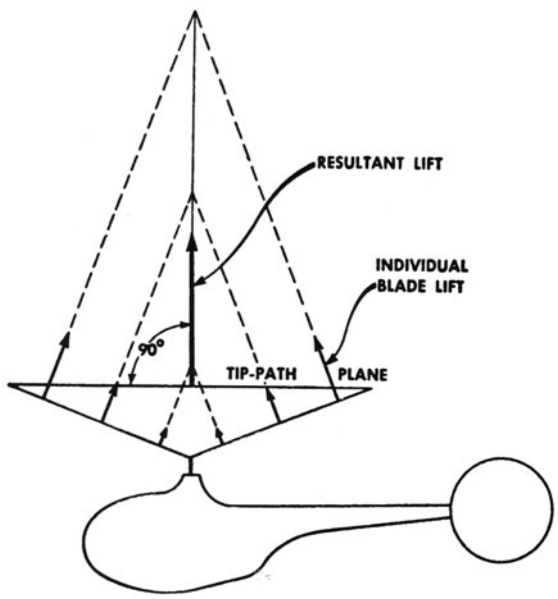


Fig. 10. Direction of resultant lift.

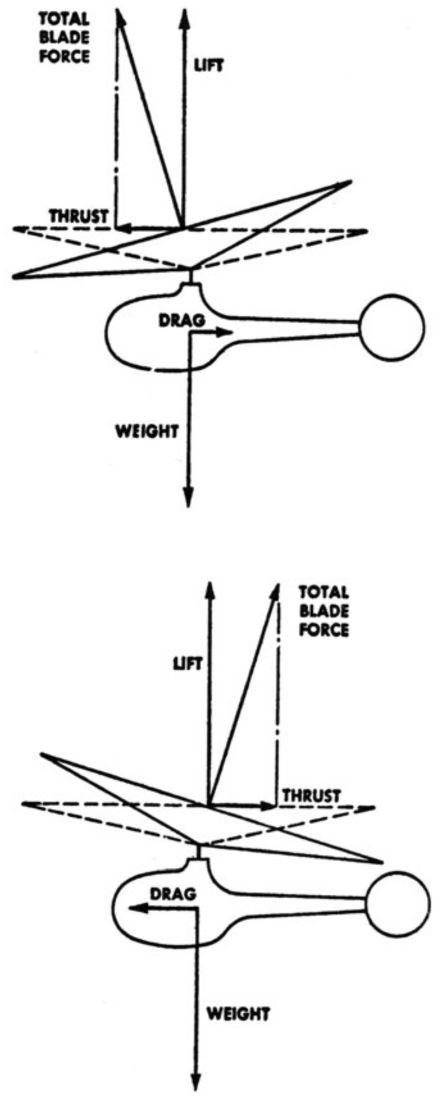


Fig. 11. Forces in forward and rear flight.

by the tips of the blades in making a cycle of rotation. The lift on the individual blade is perpendicular to the airfoil, but the resultant lift developed by the several blades is perpendicular to the tip-path plane. See Fig. 10. Lift increases in magnitude from root to tip of blade because of increase in velocity.

In vertical flight, the tip-path plane is horizontal; and in forward, backward, or sideward flight, the plane of rotation is tilted off the horizontal, thus inducing thrust in the direction of inclination. For example, to establish forward flight, resultant lift is inclined forward. See Fig. 11. Total force, being tilted off the vertical, acts both upward and forward; therefore it can be resolved into two components. One component is lift; the other is thrust. Likewise, flight may be established sideward, or in any horizontal direction, by tilting the tip-path plane in the direction of desired flight. Also, the rate of movement or speed depends upon the degree of tilt of the resultant lift force. Note the magnitude of thrust at the two speeds shown in Fig. 12.

Torque. Torque effect is displayed in a helicopter by the turning of the fuselage in the opposite direction to the rotation of the main rotor system. This reaction is in accord with Newton's third law of motion (to every action there is an equal and opposite reaction). The engine is the initiating force that drives the rotor system in a counterclockwise direction, and the reaction to this driving force would cause the fuselage of the helicopter to rotate with an equal force in a clockwise direction. See Fig. 13. Torque is of real concern to both the pilot and designer. Adequate means must be provided not only to counteract torque, but also for positive control over its effect during flight.

The designers of helicopters employ several methods of compensating for torque reaction. The dual-rotor type helicopter turns the two main rotor systems in opposite directions, thus counteracting the torque effect of one rotor by the torque effect of the other. The coaxial configuration likewise turns its rotors in opposite directions to equalize the torque effect. In the case of jet helicopters, if the engines are mounted on the tips of the rotor blades, no torque reaction is transmitted to the fuselage because the reaction is directly between the blade and the air. In the single main rotor helicopter, torque is usually counterbalanced by a vertically-mounted tail rotor which is located on the outboard end of the tail-boom extension. See Fig. 14. The tail rotor develops horizontal thrust that opposes the torque reaction. The pilot can vary the amount of horizontal thrust by activating foot pedals which are linked by cables to a pitch changing mechanism in the tail rotor system.

Ground Cushion. Also called ground effect, this is a volume of packed air built up between the rotor blades and the ground when the helicopter hovers near the ground. The downward flow of air strikes the ground and is partially trapped under the main rotor system. The air packs because it cannot escape as rapidly as the downward flow; therefore a cushion of slightly compressed air is established. The packed air is denser, thus increasing the efficiency of both the engine and the rotor system. The ground cushion is effective to a height of approximately one-half the rotor diameter; above this height, the air cannot be effectively trapped: Also, the ground-cushioning effect is lost at airspeeds in excess of ten miles per hour.

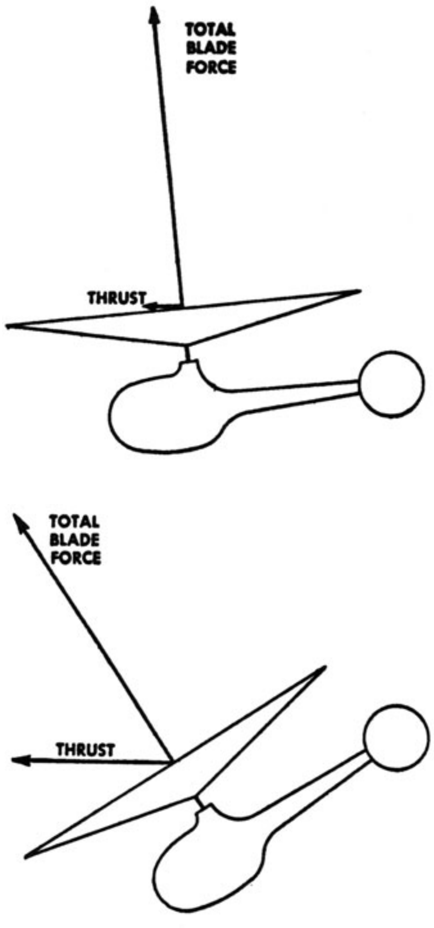


Fig. 12. Effects of slow and high speed.

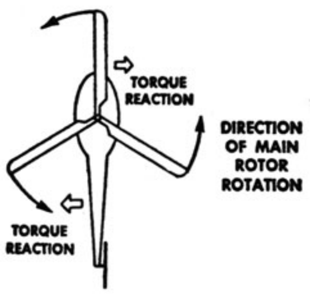


Fig. 13. Torque reaction.

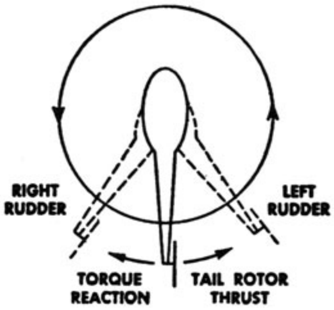
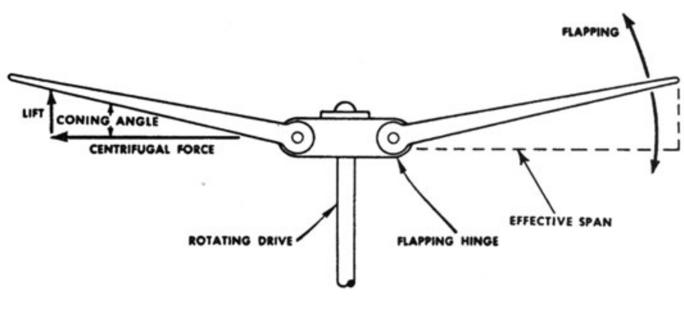


Fig. 14. Torque correction.

Translational Lift. This is the additional lift developed by a helicopter in horizontal flight. This lift becomes noticeably effective at an airspeed of 10 to 15 miles (16 to 24 kilometers) per hour and it continues to increase in magnitude as speed is increased. As horizontal flight is progressively induced, a higher inflow of air is established through the rotor disk, and greater lift is produced because of increased rotor efficiency. However, when a speed of from 45 to 50 miles (72 to 80 kilometers) per hour is reached, translational lift is canceled by fuselage drag. When hovering from 6 to 8 feet (1.8 to 2.4 meters) above the ground, the helicopter is aided by the ground-cushion effect.

Dissymmetry of Lift. This is the unequal lift that develops between the advancing half of the disk area and the retreating half of the disk area during horizontal flight. The tip-speed rotational velocity is usually constant when the helicopter is hovering in a no-wind condition. Lift is equal on the advancing and retreating halves of the disk area when the craft is hovering because the angle of attack is constant and velocity airflow over the rotor blades is the same. When the helicopter enters forward flight, however, there will be a difference in airspeed between the advancing half and the retreating half of the disk area. To the rotational velocity on the advancing side of the disk is added the forward speed; the latter is subtracted on the retreating side. Forward flight of 50 miles (80 kilometers) per hour, therefore, would establish a differential of 100 miles (161 kilometers) per hour, a condition, if uncorrected, would develop unequal lift and the helicopter would turn over.

It is normal in rotor-head design to incorporate a flapping hinge, a device which permits the rotor blade to flap upward. Under normal operational conditions, the high-speed rotation of the rotor system develops a centrifugal force of approximately 20,000 pounds (88,930 N) on each blade. Centrifugal force holds the blades in a horizontal plane, but lift will cause the blade to rise vertically. The rotor blade will take the resultant position between centrifugal force and lift. It is normal for the rotor blades to take this coned-up attitude. See Fig. 15. The centrifugal force will be constant throughout the complete cycle of 360 degrees. Lift will vary between the advancing and retreating portion of the disk area in forward flight because of difference in air flow velocity. During forward flight, the advancing blade will flap higher because it has greater lift, and the retreating blade will flap to a lower angle because it has less lift. As the rotor blade flaps up, the effective lift area is lessened; and vice versa. On the retreating half of the disk area, however, reduced airspeed developed less lift. Therefore the retreating blade will assume a more horizontal attitude. Experience has proved that blades free to flap will assume a position which develops symmetry of lift between the advancing and retreating portions of the disk area.



Gyroscopic Precession. This is the innate quality of all rotating bodies by which application of a force perpendicular to the plane of rotation will produce a maximum displacement of the plane approximately 90 degrees later in the direction of rotation. See Fig. 16. Thus, if a downward force is applied to the right side of a rotating disk, gyroscopic precession will cause the disk plane to tilt to the front, provided the disk is turning from right to left. Maximum resulting displacement occurs approximately 90 degrees further in the direction of turning, but speed of rotation, weight, and diameter of the disk, and friction are factors which determine the actual displacement of the system. The main rotor system of a helicopter displays the phenomenon of gyroscopic precession. The applied force is introduced by pitch change on the main rotor system. As the pilot moves the cyclic stick forward, it causes the control plane to tilt forward, thus introducing an equal but opposite pitch change at points 180 degrees apart in the cycle of rotation. Thus, if a linkage were not provided to take care of precession, the helicopter would fly 90 degrees out of phase. Forward stick movement

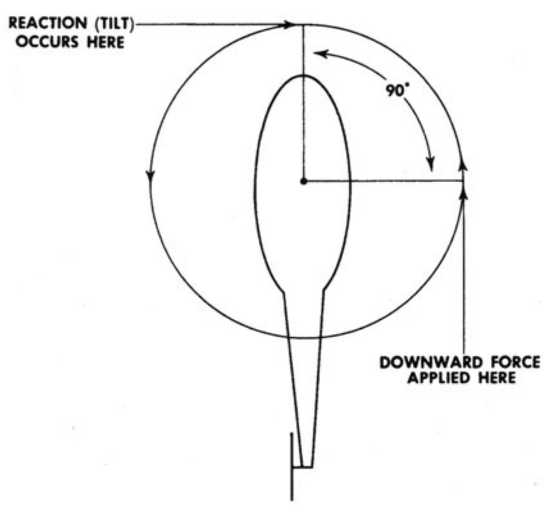


Fig. 16. Gyroscopic precession.

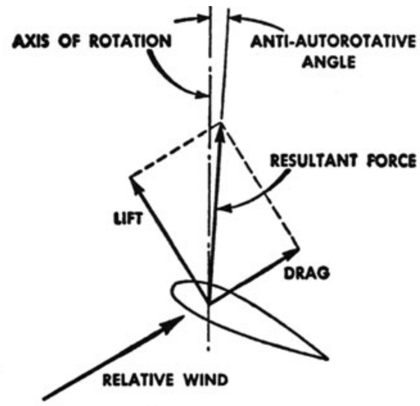


Fig. 17. Low-pitch angle.

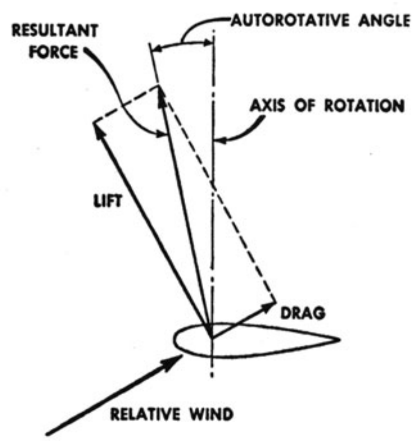


Fig. 18. High-pitch angle.

would cause the craft to fly to the left. Thus, it is common practice to set the cyclic pitch change back approximately 90 degrees in the cycle of rotation. Various types of linkages are used.

Autorotation. This is the process of producing lift with rotor blades that freely rotate because of the developed aerodynamic forces resulting from the flow of air up through the rotor system. Under power-off conditions, the helicopter will descend; thus the flow of air will be established upward through the system. The rotor is automatically disengaged from the engine by a free-wheeling device and the necessary power required to overcome parasitic and induced drag of the rotor blades is obtained from the potential energy due to the helicopter's weight and height above ground. This potential energy is converted into kinetic energy which is used to drive the rotor system during descent.

During autorotation, it is essential that the pitch angle of the rotor blades be reduced materially. The change in direction of airflow through the rotor causes a change in the direction of the relative wind which greatly increases the angle of attack at which the rotor blades are operating. If the pitch were not reduced, the blade would stall for much the same reason that a conventional airplane's wing stalls when the nose of the aircraft is pulled up too high. When the pitch angle of the blade is low and the angle of attack is large, the resultant lift force lies ahead of the axis of rotation of the blades, tending to keep the blades turning in their normal direction. See Figs. 17 and 18. If, on the other hand, the pitch angle remains high, drag is increased and the resultant lift force lies behind the axis of rotation, tending to slow and stop the rotor. Autorotation is an emergency procedure that permits the helicopter to make a safe landing in case of engine failure. It is necessary to maintain the speed of the rotor at sufficient rpm to provide not only adequate airflow over the rotor blades, but also the required centrifugal force to hold the blades in an extended attitude; otherwise the blades would fold up and the helicopter would tumble out of control.

Pendular Action. The fuselage of the helicopter is suspended from the drive shaft that mounts the main rotor head. Because the fuselage is bulky and suspended from a single point of attachment, it is free to oscillate laterally and longitudinally much in the same fashion as a freely-swinging pendulum. As the rotor system introduces horizontal translation, the fuselage is dragged in the direction of induced flight. During established forward flight, the fuselage will assume a nose-low attitude. In effect, as the tip-path plane is inclined forward, resultant lift is inclined from the vertical, thus introducing thrust. The main drive shaft of the helicopter will have a tendency to align itself with the inclined resultant lift force. See Fig. 19.

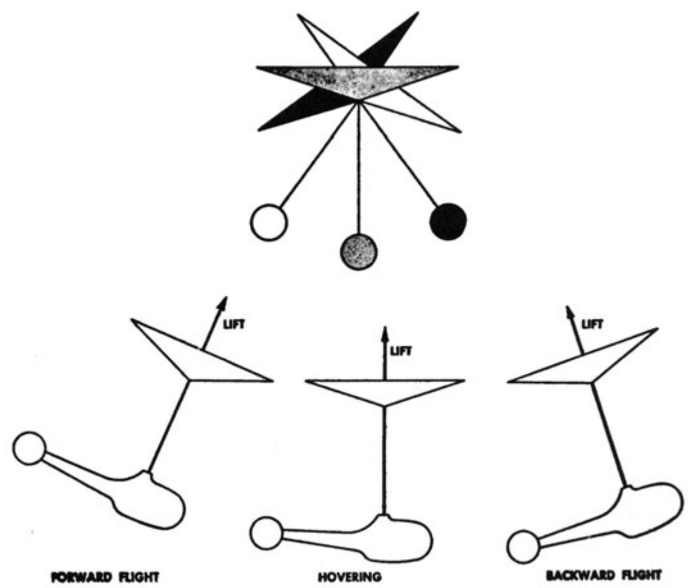


Fig. 19. Pendular action.

Other factors that are important to the design of a helicopter include (1) Resonance; and (2) weight and balance. Generally, sympathetic resonance has been well overcome by controlling design features of gear boxes and other mechanisms. Ground resonance always has been a knotty problem. This is a self-excited vibration which develops when the landing gear repeatedly strikes the ground, thus unseating the center of mass of the main rotor system. The pounding effect of the landing gear is prone to occur during take-off and landing when the helicopter is from 87 to 93% airborne. The aircraft, being light on the landing gear, bounces from one wheel to another in rapid succession, setting up a pendular oscillation of the fuselage. The succession of shocks is transmitted to the main rotor system, and the main rotor blades straddling the pounding wheel are forced to change their angular relationship. This condition unbalances the main rotor system, which in turn transmits the shock back to the landing gear. To control this potentially damaging condition, various dampening devices are used to control the unbalancing of the rotor system, and helicopter pilots are trained to avoid critical maneuvers conducive to agitating ground resonance.

Operational Helicopters

Although helicopters find numerous applications in commercial aviation for commuting, for reconnaissance by law enforcement and civilian emergency agencies (for example, in transporting the injured from the site of an accident to hospital), by the news media for coverage of special events and routinely for reporting on traffic congestion, and by the operators of large farming and ranching operations, among other important uses, helicopters find their major application by the military. As just one example of military interest in helicopters, in the late 1950s, the U.S. Navy paid special attention to improving its antisubmarine warfare capabilities. The integration of carrier-based and land-based fixed wing aircraft with helicopters and surface units proved to be a giant step in the right direction. The amphibious assault mission made great strides with the concept of vertical assault, which employed helicopters to speed materials and personnel from shipboard to points ashore. Thus, the 1960s witnessed the passing of the lighter-than-air vehicles and flying boats from the Navy's inventory. While the importance of the helicopter has not diminished, much attention in recent years has also been given to VTOL craft. The role of the helicopter in the Vietnam War hardly requires additional emphasis here. It was the principal tool used during that war for moving Army equipment and personnel.

The first American helicopter with metal rotor blades (the Sikorsky S-52) was originally flown in two-seat form on February 12, 1947. It was powered by a 178 hp Franklin engine and was claimed to have been the first helicopter to have performed a loop. In later developments for the Army and Marine Corps, the S-52 spawned several models, some of which are exhibited today at the U.S. Army Aviation Museum, Fort Rucker, Alabama. Other firms active early in the helicopter field included Bell Aircraft and Hughes Aircraft, among others.

An abridged list follows of operational helicopters, most of which are periodically modified, introduced during the last twenty years.

Sikorsky S-65 (heavy assault helicopter). The "Sky Crane." The HH-52 amphibious helicopter for search and rescue missions. The S-70 (U.S. Amry UH-60A), an 11-seat troop transport also used for medical evacuation, battlefield command and control, and as a reconnaissance aircraft. Lifting capacity of cargo hook, 8000 pounds (3600 kg). The S-76 designed for civil aviation, carrying 12 passengers for a distance of about 460 mi (740 km), with a model especially adapted for use in offshore rig support operations, having auxiliary fuel tanks which permit the craft to transport 8 passengers for more than 690 mi (1110 km).

Bell UH-1 (also called "Huey" or "Iroquois"). Introduced as a multi-purpose craft in the 1970s. Maximum weight on takeoff, 10,000 pounds (4536 kg), maximum speed of 121 mi (195 km) per hour, range of about 300 mi (483 km), ceiling of 11,500 ft (3505 m), requiring a crew of one and carrying 14 passengers. The OH-58A ("Kiowa"), introduced as a multipurpose craft in the late 1960s. Rotor blade diameter, 35.4 ft (10.8 m), maximum weight on takeoff, 3000 pounds (1361 kg), range of 350 mi (563 km), ceiling, 19,000 ft (5791 m), crew of 2, passengers 2. The AH-1 ("Cobra"), a slim-bodied gunship with a fighterlike cockpit seating a gunner in front and pilot above and behind. The craft accommodates diverse forms of armament. Rotor blade diameter, 44 ft (13.4 m), maximum weight on takeoff, 10,000 pounds (4536 kg), maximum speed 210 mi (338 km) per hour, range 360 mi (579 km), ceiling 10,550 ft (3216 m), combat crew of 2.

Dornier Do 132. Introduced in Germany in 1971. A passenger helicopter, one 720-horsepower engine; rotor blade diameter, 35.1 feet (10.7) meters); length, nearly 25 feet (7.6 meters); height, just over 9 feet (2.7 meters); empty weight, about 1500 pounds (680 kilograms); maximum weight on take-off, about 3635 pounds (1649 kilograms); maximum speed, just over 140 miles (225 kilometers) per hour; range, 275 miles (442 kilometers); crew of 1; passengers, 4.

Augusta A-106. Introduced in Italy in late 1960s as an antisubmarine helicopter; one 350-horsepower turbine engine; rotor blade diameter, just over 31 feet (9.4 meters); height, just over 8 feet (2.4 meters); length, nearly 29 feet. (8.8 meters); empty weight, about 1520 pounds (689 kilograms); maximum weight on take-off, about 3085 pounds (1399 kilograms); maximum speed, 110 miles (177 kilometers) per hour; range, 460 miles (740 kilometers); crew of 1 (total).

Aerospatiale-Westland SA-300. Introduced in France, operational in 1970; transport helicopter; two 1320-horsepower turbine engines; rotor blade diameter, just over 49 feet (14.9 meters); length, just over 46 feet (14 meters); height, nearly 14 feet (4.3 meters); empty weight, about 7560 pounds (3429 kilograms); maximum weight on take-off, 14,110 pounds (6400 kilograms); maximum speed, about 175 miles (282 kilometers) per hour; range, nearly 400 miles (644 kilometers); ceiling, 15,750 feet (4801 meters); crew of two; passengers, 16.

Westland WG 13N Lynx. Introduced in Great Britain in the 1970s; passenger helicopter; two 900-horsepower turbine engines; rotor blade diameter, 42 feet (12.8 meters); length, just over 38 feet (11.6 meters); height, just over 11 feet (3.4 meters); empty weight, nearly 7500 pounds (3402 kilograms); maximum weight on take-off, about 8780 pounds (3983 kilograms); maximum speed, nearly 185 miles (298 kilometers) per hour; range, 150 miles (241 kilometers); crew of two; plus passengers.

EH101. An Anglo-Italian three-engined helicopter under construction, as of the late 1980s, by Westland and Augusta. Nicknamed the "Iron Bird," the craft is in an experimental stage. The rotor blades, made from composites, permit an advanced aerodynamic airfoil with high-speed blade end. The rotor is equipped with powered blade folding, with provision for a backup manual fold facility.

V/STOL Craft

The concept of taking off and landing vertically inspired early efforts on helicopters just described. Attempts during the early 1900s included: (1) The Gyroplane, designed and built in Europe in 1907 by Bréguet and Richet, managed to lift only a few feet off the ground. The engines were inadequate. Following these experiments, Bréguet decided to concentrate, and successfully, on the design of more conventional airplanes. (2) Frenchman Paul Cornu, while aboard his craft, succeeded in lifting a fragile design off the ground about one foot for a period of about 20 seconds in 1907. (3) Igor Sikorsky, in 1909, constructed a helicopter prototype using an Anzani engine, but the latter proved inadequate to lift the craft off the ground. For several years thereafter, Sikorsky concentrated successfully on the design of seaplanes. (4) In Denmark, Ellehammer, a designer of prior fixed-winged airplanes, managed to lift his helicopter design prototype a few inches above the ground in 1916. This craft was equipped with two coaxial rotors. (5) Frenchman Etienne Oemichen succeeded with his design to take off vertically for a flight of a few hundred yards in mid-1924; (6) The Marquis Raul Patteras Pescara (Spain) in the same year managed a flight of 2,415 feet (736 meters). Pescara continued with other models, but ceased activity upon the successful demonstration of Juan de la Cierva's autogiro which, at that time, appeared to provide the answers being sought from helicopters; (7) In 1930, Italian D'Ascanio's design, piloted by Marinello Nelli, broke the prior distance record, achieving a flight of some 3,535 feet (1077 meters), and at a record height of 59 feet (16 meters) above ground. But probably success for the first practical helicopter designs should go to the Germans with the development of the Focke-Wulf FW-61, which appeared in 1936; and one year later when Igor Sikorsky built and flew the VS-300.

Autogiro. This craft was developed in Spain and flown in 1923. The autogiro consists of a wingless fuselage mounting a pylon which contains a rotating head to which are affixed three or four balanced blades of airfoil section resembling a propeller configuration. The blades are rotated by the action of the relative wind and thus are self-rotating, hence the name of the craft. The autogiro is equipped with a regular power plant and propeller that produce the necessary forward thrust to get the craft in motion and keep it in motion when in flight. The angle of the rotor to the fuselage is controlled by the pilot, and this takes the place of the normal control surfaces of the conventional airplane. The blades rotate at speeds that give an average air velocity over them considerably in excess of the autogiro's airspeed, and so the autogori may be flown at speeds lower than the stalling speed of the airfoil section. This is an advantage, in that it permits the autogiro to land in small fields, with short landing runs, to descend almost vertically, and to approach "hovering" flight. Although autogiro development essentially has been abandoned, primarily because of the successes of later helicopter designs, much of the information gained in autogiro experiments has proven of much value to helicopter pioneers and later VTOL designers.

During the past few decades, a number of V/STOL have been proposed and built, either experimentally or in production. Principal configurations include: (1) *Compound or convertible aircraft*—essentially helicopters with propellers for horizontal flight. At take-off, the engines power the rotors; at altitude, the power is transferred to the propellers. A serious disadvantage of this concept is the substantial drag presented by the rotors during horizontal flight. (2) *Vertiplanes*—large, propeller-equipped airplanes, with a jet engine for effecting vertical take-off. In appearance, past designs of these craft look like an essentially conventional fixed-wing airplane. (3) *Tilt-Prop or Tilt-Wing Airplanes*—essentially airplanes with wings and jets or propellers, either of which can be tilted by 90 degrees during take-off and landing. (4) *Bidirectional Jet Aircraft*—jet-propelled aircraft in which the jet engine(s) can be directed downward for vertical take-off and turned toward the rear for horizontal flight when at altitude.

The AV-8B V/STOL. Developed for the United States Marine Corps, this aircraft is a single-seat, single-turbofan-powered craft for close air support and interdiction missions. The craft is powered by a Rolls Royce Pegasus 11 vectored thrust turbofan, with 21,500 pounds of thrust (95,600 N) without afterburning. The length is 46.3 feet (14.1 meters); height is 11.6 feet (3.5 meters); the wingspan is 30.3 feet (9.2 meters); and the wing area is 230 square feet (21.4 square meters). Take-off distance is 0–1100 feet (0–305 meters). The first flight of the YAV-8B prototype occurred in November 1978. Combat range is 600+ nautical miles (1112 kilometers); ferry range is 2460 nautical miles (4558 kilometers). See Fig. 20. The aircraft can carry armament up to 9200 pounds (4175 kilograms) on several stations: two 30 millimeter cannons; laser or TV-guided weapons; as well as up to four 300-gallon (1136-liter) fuel tanks. The Marine Corps V/STOL close air support uses three types of land sites as well as a variety of ships. Closest to the battle area would be a number of V/STOL forward sites, accommodating two to four AV-8Bs. These sites would normally have fuel and ordnance for turnaround operations only. If a STO (short take-off) strip is not available, the aircraft would operate in the VTOL (vertical take-off and landing) mode. Located about 50 miles (80 kilometers) from the battle-field would be a V/STOL facility with at least a 600-foot (182-meter) strip, providing six to ten aircraft turnaround, support, and maintenance. The V/STOL main base would have at least a 1500-foot (457-meter) strip and all the logistics and support assets for prolonged AV-8B operations.

Fig. 20. U.S. Marine Corps' AV-8B V/STOL aircraft for close air support and interdiction missions. (*McDonnell Aircraft Company; McDonnell Douglas Corporation.*)

The AV-8B's vertical takeoff and landing ability is derived from four exhaust nozzles—two on either side of the aircraft—positioned around the plane's center of gravity. These nozzles can be rotated from the full-aft position, for forward flight, to a full-down position for vertical operations. Within the engine, some rotating parts turn clockwise while others turn counterclockwise. This is necessary to prevent gyroscopic effects which, if all parts rotate in the same direction, can make the aircraft difficult or impossible to control during hover, and during transition from hover to forward flight or from forward flight to hover.

All that is required for take-off or landing is an amphibious assault ship, or a clearing large enough for a 72-foot (22-meter) square aluminum mat, a section of two-lane road, or even a damaged airfield. As with the trend in the design and construction of modern aircraft, much stress is being given to composites and graphite epoxy materials. See Fig. 21.

VTOL Craft

For many years, aircraft designers have been seeking a vehicle that would fly like a rotary wing aircraft in a hover and fly like a fixed wing aircraft in up-and-away flight. The "X-Wing," currently under development by NASA (National Aeronautics and Space Administration), DARPA (Defense Advanced Research Projects Agency), and Sikorsky Aircraft in a joint program, is the most promising concept as of the late 1980s for performing both the traditional airplane and helicopter roles.

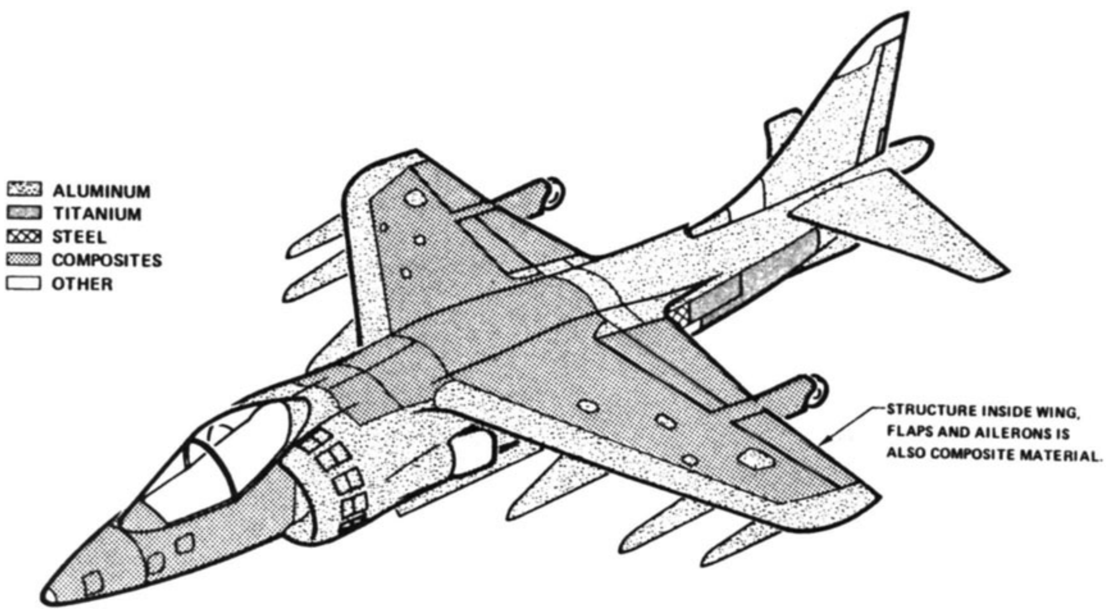


Fig. 21. Types of materials used in the AV-8B V/STOL aircraft. As a percent of structural weight: aluminum, 48.4%; titanium, 8.5%; steel, 14.5%; composites, 23.3%; other materials 5.3%. Pylons not included in these figures. Approximately 1186 pounds (540 kg) of graphite epoxy are used in the structure. (*McDonnell Aircraft Company; McDonnell Douglas Corporation.*)

The unique aspect of the X-wing is that it derives its lift for both hover and forward flight from the same airfoil. In August 1986, the X-wing prototype was delivered by Sikorsky to NASA's Dryden facility at Edwards Air Force Base for flight testing.

As pointed out by Brahney, the most challenging part of X-wing aerodynamics is the conversion between helicopter and fixed wing modes of flight. The retreating side of the rotor disk must continue to support its share of total system lift, even as its airflow is reversing direction. As the rotor/wing slows, the azimuthal segment with the leading edge blowing grows until it encompasses the entire left side of the disk. Thus, on the left side of the disk, what was the leading edge in helicopter flight becomes the trailing edge in the fixed wing mode.

Air is fed out of the compressor into a plenum. The inner wall of the plenum has two rows of 24 valves. The top row supplies the leading edges; the bottom row feeds the trailing edges. The valves and plenum do not rotate with the rotor system. What does rotate is a series of 8 receiver ducts (four for leading edges and four for trailing edges) inside the plenum which receive the valve-modulated air from the plenum and feed it to the individual blades. This rather complex system is described in detail in the Brahney (1986) reference listed.

Additional Reading

See references listed at end of article on **Airplane**.

HELIOPAUSE. This is defined as that distant location in outer space at which the sun's influence ends and is replaced by the interstellar medium. For a number of years, astronomers have estimated that the heliopause occurs between 50 and 100 astronomical units (AU) from Earth. [One AU 5 approximately 93,000,000 miles (150 million km)]. Some experts have revised this estimate to between 85 and 100 AU.

Because of the extremely long distance that has been reached by *Pioneer 10*, the heliopause is gaining interest among astronomers. If the 85 AU estimate is close, then the space probe may reach that distance within the lifetime of many contemporary astronomers. The probe as of 1990 was estimated at a distance of 50 AU. Should the 100 AU figure be correct, *Pioneer 10* should achieve that distance during the year 2010. Although the probe was built to last only 30 months, National Aeronautics and Space Administration officials now feel that the *Pioneer 11* may survive for several more years. It has been noted from both *Pioneer 10* and *Pioneer 11* that cosmic rays appear to increase at about 2 percent for every astronomical unit the spacecraft travels. It is predicted that once the heliopause is reached, the cosmic radiation will be constant with time.

It is interesting to note that *Voyager 2*, if it remains operable, will reach the heliopause in the year 2007. However, expectations for this are not optimistic because of the much greater complexity and vulnerability of the *Voyager 2*.

HELIOSTAT. An arrangement of mirrors, driven by clockwork, used to reflect a beam of sunlight in a fixed direction as the sun moves across the sky. The heliostat is used in the control of some astronomical instruments as well as in some solar energy systems for tracking the sun. See also **Solar Energy.**

HELIOTROPE. See **Bloodstone.**

HELIOTROPIC WIND. See **Winds and Air Movement.**

HELIUM. Chemical element symbol He, at. no. 2, at. wt. 4.0026, periodic table group 18 (inert or noble gases), mp −272.2°C (20 atmospheres), bp −268.93°C (4.2144K), specific gravity 0.124 at 4.2144K. The element has no triple point and can be solidified only by applying high pressure to the liquid phase. Described later, liquid helium undergoes a change in its physical properties at 2.178K, known as the *lambda point*. Solid helium has a close-packed hexagonal crystal structure (subject to further study and confirmation). At standard conditions, helium is a colorless, tasteless, odorless gas. There are two natural isotopes ^3He and ^4He, with ^4He being slightly less than 100% abundant. The boiling point is 3.2K for ^3He. Radioactive ^5He and ^6He have extremely short half-lives. See also **Radioactivity.** The first ionization potential for helium is 24.58 eV; second, 54.14 eV. Other physical properties of helium are described under **Chemical Elements.**

Like the other rare gases, helium exhibits negative chemical properties with ordinary materials under normal conditions. Under the influence of electric glow discharge or electron bomdardment, helium forms compounds with tungsten and other metals, as well as with iodine, sulfur, and phosphorus. In a vacuum electric discharge tube helium shows green to canary-yellow glow. Discovered first in the vapors surrounding the sun by Lockyer in 1868, through the yellow spectral line near the two yellow lines of sodium, then by Ramsay in 1895 in the mineral clevite.

Helium occurs (1) in minerals of uranium and thorium, such as clevites, pitchblende, carnotite, monazite, and also in beryl, (2) in mineral waters (1 part He per thousand of water, in some Iceland waters), (3) in volcanic gases, (4) especially in certain natural gases of the United States. The first discovery of this kind was made in Kansas.

Uses. Industrially, helium is used to provide an inert gaseous shield for arc welding, for growing transistor crystals, in the production of titanium and zirconium, to fill the space between optical lenses in instruments, as the carrier gas in some chromatographic apparatus, as a liquid bath for masers and cryotrons, as a refrigerant for furnishing the low temperature required for superconducting electrical equipment, in lasers, as a diluent gas in deep-sea diving applications, as a heat-transfer medium in gas-cooled nuclear reactors, and as a leak-detecting medium for testing pressure and vacuum equipment. Now, to a rather limited extent, helium is used as a lifting gas for airships and for balloons used in meteorological investigations. Helium is used in aerospace programs in several ways, including its use in propellant tanks as a compressed gas which expands and takes the place of fuel as the fuel is consumed, in ground-support equipment, and in communication satellites for providing the low temperature required for sensitive electronic systems. In medicine, helium sometimes is mixed with oxygen for patients with certain respiratory ailments and also it is mixed with certain anesthetics to reduce the hazards of forming an explosive mixture with air.

Future possible uses of helium have a direct influence on the conservation of helium resources. The known resources are not large in comparison with most other raw materials, and most authorities are of the opinion that conservation measures should be continued. However, most helium demand projections have proved overly optimistic and consequently have lessened the pressure for conservation. Natural gas streams, of which He is but a minor constituent, are produced commercially for sale and consumption as fuel. To separate helium by stripping from natural gas is an expense that is not attractive to the producers of natural gas. Thus, it is evident that helium conservation, under current supply/demand conditions, must stem from government regulation. Even though stripping helium from natural gas is costly, later costs to recover helium lost to the atmosphere from burning He-bearing natural gas would be many times greater.

The history of helium conservation measures dates back to 1925 when the U.S. Congress passed the Helium Act of 1925. Congress amended the act in 1960 to provide for stripping natural gas of its helium, for purchase of the separated helium by the government, and for its long-term storage. In 1971, after about 28 billion cubic feet had been stored (in a federally owned gas field called Cliffside near Amarillo, Texas), the purchase program was terminated by the government, an action that, as reported by Hammel et al., unleashed several lawsuits and not a little acrimony. As of the present, most of the litigation has been concluded, much of the He that could have been saved has been wasted to the atmosphere, and the gas fields supplying the He are almost depleted. However, in the meantime, a new and rich source of He has been discovered in southwestern Wyoming that could ensure adequate supplies for many decades if an appropriate new federal policy on He were developed and implemented. The new field, first explored by Mobil in 1960, led to an initial estimate of 3 to 15 billion cubic feet of He in the Tip Top drilling unit of that field. However, the natural gas from that unit was initially judged unfit for sale as a natural gas fuel. The borehole was cemented shut and the well abandoned. In the early 1980s, it was established by additional drilling that the amount of He recoverable from the Wyoming field is at least 200 billion cubic feet. Left untouched this would represent an excellent long-term helium reserve. More than 90% of this field lies within federal land boundaries.

With increasing incentives because of natural gas pricing (Natural Gas Policy Act of 1978), several firms have plans to drill and develop nearly 250 deep wells, from which 2.8 billion cubic feet of acidic gas per day would be produced. If private developers elect not to conserve the helium from this project (Riley Ridge Natural Gas Project), it is estimated that about 5 billion cubic feet of He would be vented to the atmosphere each year.

Unexpected Uses for Helium. Hammel (1984) points out that new uses for He continue to emerge. These include a 49-meter-diameter He-filled sphere that has been proposed as a lighter-than-air hoist capable of moving loads in excess of 90 tons. Another use is in superconducting magnets for imaging with nuclear magnetic resonance. Proposals for the use of large amounts of He continue to be made in the national security area.

Origin of Helium. As pointed out by Hurley (1954), helium has a geologic occurrence and distribution unique among the elements. It is a product of radioactive disintegration of uranium and thorium within the earth's mantle and crust, but flows to the surface at a rate less than that of its generation, because most of it is driven into crystal structures of rock minerals until released by alpha radiation damage near radioactive concentrations. Mobile helium rising through the crust may then be trapped, along with other gases, beneath relatively impermeable barriers. Nitrogen is almost always associated with helium in natural gases, although this has not been fully explained. Also, carbon dioxide is abundant in some helium-rich gas mixtures.

Liquefaction of Helium. This was accomplished by Onnes in 1908 in Leiden, and Keosom in 1926 succeeded in solidifying helium in the same laboratory. Relatively recently, helium has been solidified at room temperature. The melting pressure at 24°C is 115 kilobars, in complete agreement with the Simon equation. Besson and Pinceaux (1979) developed an original apparatus for the experiment, which allowed loading of the cell at room temperature. Diamond anvil cells were used in the procedure.

Liquid Helium II. Upon cooling, ^{4}He liquefies at atmospheric pressure at 4.216K to form an essentially normal liquid, liquid helium I. On further cooling to the lambda-point, 2.178K at one atmosphere, a change occurs to liquid helium II. The latter has a very low viscosity (hence the name "superfluid") and a very high thermal conductivity, which produce such phenomena as the creeping of a film over the edge of the container, and the fountain effect, in which the liquid sprays out of a capillary. Superfluidity is commonly explained in terms of a two-fluid theory. Thus, London and Tirza attribute the properties of helium II to a mathematical peculiarity in the distribution function of Bose-Einstein statistics, whereby below the λ-point, a finite fraction of the atoms fall into a ground state of zero thermal energy. In this state they would have the properties of a superfluid. However, this theory has not yielded good quantitative predictions of the properties of the aggregate liquid helium. ^{3}He, which follows Fermi-Dirac statistics, does not have a superfluid state.

Landau treats liquid helium by an approach similar to that of the Debye theory of solids. The longitudinal and transverse sound waves, which are the elementary excitations of that theory of solids, correspond in the case of liquid helium to phonons and rotons. The *phonons* are the longitudinal sound waves, while the *rotons* are another type of elementary excitation postulated by Landau to represent the rotational motion of the liquid, because a liquid cannot support transverse waves. The specific heat can be expressed as the sum of contributions from phonons and rotons. Landau derived expressions for these which fit the data and experiments quite closely up to 1.6K.

Feynman developed wave functions to provide an atomistic interpretation of Landau's spectrum of elementary excitations.

The complexity of the helium II problem is apparent at once when one attempts to extend the equations of classical hydrodynamics to this two-component system, in which each component has its own density and velocity. Khalatnikov derived such equations by ignoring terms of second order.

Still another area of investigation has been that of the properties of ^{3}He-^{4}He mixtures. As stated above, ^{3}He exhibits no λ-transition and no superfluidity. It has a critical temperature of 3.35K and a boiling point of 3.2K, against values of 5.2K and 4.216K for ^{4}He.

The most abundant helium atoms, ^{4}He, are bosons, but the ^{3}He atoms are fermions. This has a consequence that liquid ^{3}He does not show superfluidity—a property very probably connected with the Bose-Einstein statistics obeyed by the ^{4}He atoms.

Donnelly and associated researchers (University of Oregon) has observed that in the future liquid helium rather than air may be used in a much down-scaled wind tunnel, perhaps with experiments conducted within the space of an average room versus current, very large wind tunnels. It is envisioned that a tunnel could be filled with superfluid liquid helium, taking advantage of the liquid's absence of viscosity and friction as previously described. Based upon quantum-mechanical factors, a source of heat in the tunnel could cause extremely fast currents to flow in the liquid.

Peterson (reference listed) reported in early 1991 that researchers at Harvard University made what is considered a remarkable prediction regarding the energy-level transitions that occur in a helium atom. The agreement between theoretical calculations and experimental results show that computational methods for constructing a model of a two-electron atom can work, thus bridging the gap between theory and practice.

Chemistry. The most striking properties of helium are its emission as the positively charged (+2) alpha particles in radioactive changes, its formation in radioactive change by uranium-radium and thorium-containing substances, emitting alpha particles, later losing the charge to become helium, and its production artificially by bombardment of lithium or boron with high-velocity protons or alpha rays.

Unlike the other inert gases, helium gives little evidence of compound formation with organic substances. Like neon, but unlike the others, it forms no hydrate. However, it forms compounds much more readily under excitation, due apparently to unpairing of its 1*s* electrons and promoting of one of them to the 2*s* state. The 460 kcal/g-atom of energy is readily obtained by electric discharge or electron bombardment. Under such conditions the helium molecule-ion, He_2^+, with a pair of bonding electrons (1*s*) and a single antibonding electron (1*s*), is formed, as are combinations of the type of HeH^+ and HeH_2^+. In a mercury discharge tube, the compound $HgHe_{10}$ has been found, and with various metallic electrodes corresponding helides, such as the compounds of tungsten, platinum, iron, palladium, bismuth, etc., e.g., WHe_2, Pt_3He, FeHe, PdHe, BiHe, etc., have been formed.

Additional Reading

Andron-Ikashvili, E. L.: "Reflections on Liquid Helium," American Institute of Physics, New York, 1990.

Besson, J. M., and J. P. Pinceaux: "Melting of Helium at Room Temperature and High Pressure," *Science*, **206**, 1073–1075 (1979).

Cohen, E. G. D.: "Quantum Statistics and Liquid Helium-3-Helium-4 Mixtures," *Science*, **197**, 11–16 (1977).

Cook, E.: "The Helium Question," *Science*, **206**, 1141–1148 (1979).

Donnelly, R. J.: "Superfluid Turbulence (Helium)," *Sci. Amer.*, 100 (November 1988).

Epstein, A. W.: "Cool Breeze: A Helium Superwind for Wind-Tunnel Experiments," *Sci. Amer.*, 30 (May 1990).

Hammel, E. F., Krupka, M. C., and K. D. Williamson, Jr.: "The Continuing U.S. Helium Saga," *Science*, **223**, 789–792 (1984).

Lupton, J. E., and H. Craig: "A Major Helium-3 Source at 15°S on the East Pacific Rise," *Science*, **214**, 13–18 (1981).

Ozima, M., and S. Zashu: "Primitive Helium in Diamonds," *Science*, **219**, 1067–1068 (1983).

Peterson, I.: "Helium Theory Gets High-Precision Test," *Science News*, **86** (February 9, 1991).

Staff: "Handbook of Chemistry and Physics," 73rd Edition, CRC Press, Boca Raton, Florida, 1992–1993.

Vollhardt, D., and P. Wolfle: "The Superfluid Phases of Helium 3," Taylor and Francis, New York, 1990.

HELIUM LEAK DETECTION. See **Mass Spectrometry.**

HELIX. A space curve traced on a cylinder or conical surface in such a way that all elements of the surface are cut at a constant angle. A circular helix lies on a right-circular cylindrical surface. In parametric form, its equation is $x = a \cos \theta$, $y = a \sin \theta$, $z = b\theta$ where a, b are constants and θ is the parameter. The thread of a screw is often a circular helix.

See **Conical Surface.**

HELIX FEEDER. See **Feeder (Volumetric).**

HELLBENDER (*Amphibia, Urodela*). A large aquatic salamander, *Cryptobranchus allegheniensis*, of the Mississippi river system. It reaches a length of 18 inches (46 cm) and has a flattened head and body, short legs, and a compressed tail. The gills are concealed, but otherwise it resembles the mudpuppy.

HELLGRAMMITE (*Insecta, Neuroptera*). The large aquatic larva of the dobson fly, *Corydalus*. It lives in running water and is an excellent bait for bass.

HELMHOLTZ EQUATION. An equation of the form

$$n_1 y_1 \tan \theta_1 = n_2 y_2 \tan \theta_2$$

expressing the relation between the linear and the angular magnification at a spherical refracting interface. y_1, y_2 are linear dimensions of object and image, θ_1, θ_2 the angles made by focal rays and axis at object and image points and n_1, n_2 are refractive indices of object and image space. Also called Lagrange-Helmholtz equation. (See, however, the **Abbe Sine Condition.**) A spherical surface cannot satisfy both these equations for finite angles. Hence a spherical surface can never make a perfect image.

HELMHOLTZ FUNCTION. See **Thermodynamics.**

HELMHOLTZ RESONATOR. An enclosure communicating with the external medium through an opening of small cross-sectional area. Such a device resonates at a single frequency dependent on the geometry of the resonator.

HELMHOLTZ THEOREM. The statement that if **F** is a vector field satisfying certain quite general mathematical conditions, then **F** is the sum of two vectors, one which is irrotational (has no vorticity), the other solenoidal (has no divergence).

HELMINTHOLOGY. A biological science dealing with the worms, more particularly parasitic flatworms and roundworms. Since many worms are parasitic, the term parasitology is more commonly used. The study of roundworms is important in agriculture and has resulted in the science of nematology (see **Nematoda**) which is properly a subsidiary of helminthology.

HEMATITE. The mineral hematite, ferric oxide, Fe_2O_3, occurs as thick or thin tabular rhombohedral forms, sometimes in pyramids but rarely in hexagonal prisms. It also assumes botryoidal, columnar and lamellar shapes, and may be granular or compact. Its hardness is 5.6; specific gravity, 5.26; luster, metallic to earthy or dull; color, dark gray to black; earthy forms may be different shades of red; streak, red to red-brown; translucent (in very thin flakes) to opaque. Hematite with a metallic luster is called specular iron.

It is a widely distributed and common mineral, found in igneous, sedimentary and metamorphic rocks as beds and veins, having probably been formed in many different ways under very different conditions. Beautifully crystallized hematite has been found in the Urals of the former U.S.S.R.; Rumania; Switzerland; the Island of Elba; Alsace, France; Cumberland, England. Extremely rich, large hematite ore bodies have been found and are being worked in Minas Gerais, Brazil; Cerro de Mercado, Durango, Mexico; Quebec and Labrador in Canada. The hematite ore deposits which lay along the southern and northwestern sides of Lake Superior in Michigan, Wisconsin and Minnesota have been worked to near depletion. Extensive beds of hematite are found throughout the Appalachian region from New York to Alabama, being mined near Birmingham in the latter state. Hematite occurs in quantity in Nova Scotia and Newfoundland. It is the most important ore of iron, and has other industrial uses in paint manufacture and polishing compounds. The name hematite is derived from the Greek word meaning blood.

HEMATOLOGY. That branch of medicine having to do with the study of the blood, the blood-forming tissues and the diseases of the blood.

HEMATOMA. An accumulation of free blood in the body tissues forming a localized mass. This usually follows an injury in which rupture of blood vessels takes place. See also **Brain (Injury);** and **Cerebrovascular Diseases.**

HEMATOPOIESIS. See **Blood.**

HEMATURIA. The presence of blood in the urine. This condition is found in certain forms of nephritis and with injury, tumors, stones, or calculi in the urinary tract. It is also seen in scurvy and in some cases of severe sepsis.

HEME. See **Cytochromes.**

HEMICHORDATA. A subphylum of the phylum *Chordata* containing only a few primitive marine animals without common names. The genus *Balanoglossus* has lent its name to the forms most commonly seen, although some belong to other genera. They are worm-like animals which live in mud and sand at the bottom of the ocean. The central nervous system is dorsal in this group but it remains partly or wholly at the surface. The notochord is limited to the anterior part of the body and is sometimes connected with the alimentary tract. Gill slits vary from one to many pairs. The group is also commonly named Enteropneusta and rarely Adelochorda.

There are two orders:

Order *Balanoglossida*. Worm-like animals with many gill slits and with a fleshy proboscis before the mouth. *Balanoglossus* and related forms.

Order *Pterobranchia* (*Cephalodisca*). Sessile animals, some solitary and some colonial. One pair of gill slits. A proboscis and branching tentacles lie before the mouth and the intestine is U-shaped. *Cephalodiscus* and *Rhabdopleura*.

HEMI- AND HOLOCELLULOSE. See **Pulp (Wood) Production and Processing.**

HEMICOLLOID. A colloid composed of particles of small size, i.e., ranging from 0.005 to 0.0025 micrometer in length.

HEMIGALES. See **Viverrines.**

HEMIHEDRITY. A term describing crystal symmetry operations, to indicate that only half of a symmetrical structure undergoes modification. For example, if, in truncating a cube, the process is carried out symmetrically on four out of the eight solid angles, the resulting structure exhibits hemihedral symmetry.

HEMIMETABOLA. A division of the insects characterized by incomplete metamorphosis. The immature insect differs conspicuously from the adult in form and is adapted to an entirely different mode of life; in this the group resembles the *Holometabola*. The young have compound eyes, however, and the wings develop externally as in the *Paurometabola*. The group includes the three orders, *Plecoptera, Ephemerida*, and *Odonata*, all with aquatic larvae which are called naiads.

HEMIMORPHITE. This mineral is zinc silicate, $Zn_4Si_2O_7(OH)_2 \cdot 2H_2O$, occurring in tabular and prismatic orthorhombic crystals, although often in massive and fibrous forms. There is a perfect cleavage parallel to the prism; it is brittle with a subconchoidal fracture; hardness, 4.5–5; specific gravity, 3.40–3.50; luster, vitreous; color, white, tending to translucent. Hemimorphite differs from willemite, also a zinc silicate, in that the former contains considerable water which may be driven off when heated to a high temperature.

There are many localities for hemimorphite in Europe, fine specimens having come from Saxony; Sardinia; Cumberland, Alston Moor and Derbyshire, England. It is found in Siberia, Algeria, and Mexico.

In the United States, hemimorphite has been found at Sterling Hill, New Jersey; in Lehigh County, Pennsylvania, and in Virginia, Missouri, Montana, Colorado, Utah, New Mexico, and Nevada.

The mineral is so named because of the tendency to form doubly terminated crystals showing a different grouping of faces at either end. The name is derived from the Greek meaning half and form.

HEMIPLEGIA. Loss of voluntary movement on one side of the body, commonly resulting from damage to the cerebral cortex on the opposite side of the body, or to the nervous pathways leading from it. Transient hemiplegias occur in epilepsy and hysteria but the majority are persistent and are due to hemorrhage or sometimes tumor compressing or destroying the cerebral cortex and associated tracts of nerve fibers.

HEMIPODE. See **Rails, Coots, and Cranes.**

HEMIPTERA. Many hemiptera, an order of insects containing about 21,000 species, are of economic importance. They have piercing and sucking mouth parts and live on the blood or juices of animals or the sap of plants. The wings, when present, are usually distinctive. The basal half is thicker than the terminal, and the tips overlap partially so that the margins of the wings form an X on the back. Metamorphosis is usually gradual. The chinch bug and bedbug are species of economic importance.

Bugs of several families are aquatic and some forms live on the surface of the water, supported by the surface film. The swimming forms are the water boatmen, back swimmers, and giant water bugs and the water striders skate on the surface. One of the last, *Halobates*, is the only marine insect known. Shore forms include the toad bugs. On dry land the order is represented in almost every possible habitat. The main families of *Hemiptera* include:

Belostomatidae	Giant water bugs
Cimicidae	Bed bugs
Coreidae	Squash bug
Corixidae	Water boatmen
Gerridae (also called *Hydrobatidae*)	Water striders
Lygaeidae	Chinch bugs
Miridae (also called *Capsidae*)	Leaf bugs
Nabidae	Damsel bugs
Nepidae	Water scorpions
Notonectidae	Back swimmers
Pentatomidae	Stink bugs
Phymatidae	Ambush bugs
Reduviidae	Assassin or kissing bugs
Tingidae	Lace bugs

HEMITROPIC. A term used by mineralogists for a crystal that appears to be composed of two halves of the same crystal turned partly around.

HEMLOCK TREES. Members of the family *Pinaceae* (pine family), these trees are of the genus *Tsuga*. The trees are sometimes referred to as hemlock spruces or hemlock firs. The hemlocks are evergreen trees, broadly conical. They are well known for their immunity to disease, with the exception of normal decay with age. The trees are quite tolerant of shade. The principal species not listed in the accompanying table include:

Black hemlock	*Tsuga martensiana*
Canadian hemlock	*T. canadensis*
Low weeper form	*T.c. 'Pendula'*
Formosan hemlock	*T. formosana*
Himalayan hemlock	*T. dumosa*
Northern Japanese hemlock	*T. diversifolia*
Southern Japanese hemlock	*T. sieboldii*

The Canadian or Eastern hemlock normally attains a height between 50 and 80 feet (15 to 24 meters), but under favorable conditions can approach 100 feet. This species sometimes has several stems which form a spreading tree. The foliage may be described as feathery or plumelike. The bark is a dull brown. The tree is commonly found in swamps, ravines, rocky woods, and the mountain slopes of cold areas. It is found in some of the eastern mountains up to an altitude of about 2,000 feet (600 meters). The natural range of the tree is from Labrador, Newfoundland, and Nova Scotia westward to Michigan and Minnesota and southward to Delaware and Maryland and on the mountain slopes as far south as Georgia and Alabama. It is found throughout most of New England and is particularly common in the central portions of Maine. Commercially the tree is often called hemlock spruce in the northern states; and spruce pine in the southern states. Timber from the tree is used, but is not considered a high-grade wood. It is of uneven texture, tending to splinter easily. Major uses are for pulp wood, and boxes and crating. In the green condition, Eastern hemlock wood has a moisture content of 111% and weighs 50 pounds per cubic foot (801 kilograms per cubic meter). When air-dried to 12% moisture content, the weight is 28 pounds per cubic foot (448.5 kilograms per cubic meter) or 1,000 board-feet (2.36 cubic meter) weigh 2,330 pounds (1057 kilograms). The compressive or crushing strength of the dry wood, parallel to the grain, is 5,410 pounds per square inch (37.3 MPa); the tensile strength perpendicular to the grain in the green wood is 230 pounds per square inch (1.6 MPa).

The tree makes an excellent hedge and is often used for this purpose in landscaping.

RECORD HEMLOCKS IN THE UNITED STATES[1]

Specimen	Circumference[2]		Height		Spread		Location
	Inches	Centimeters	Feet	Meters	Feet	Meters	
Carolina hemlock (1984) (*Tsuga caroliniana*)	139	353	88	26.8	54	16.5	North Carolina
Eastern hemlock (1979) (*Tsuga canadensis*)	224	569	123	37.5	68	20.7	West Virginia
Mountain hemlock (1955) (*Tsuga mertensiana*)	277	704	113	34.4	44	13.4	California
Western hemlock (1987)[3] (*Tsuga heterophylla*)	270	686	241	73.5	67	20.4	Washington
Western hemlock (1989)[3] (*Tsuga heterophylla*)	316	803	202	61.6	47	14.3	Washington
Western hemlock (1991)[3] (*Tsuga heterophylla*)	291	739	227	69.2	49	14.9	Washington

[1]From the "National Register of Big Trees," The American Forestry Association (by permission).
[2]At 4.5 feet (1.4 meters).
[3]Cochampions.

The Carolina hemlock is essentially exclusive to the Alleghany Mountains. It is a smaller tree, but has many of the characteristics of the Eastern hemlock. Normal height is about 50–60 feet (15 to 18 meters), but can grow higher under favorable conditions. The tree prefers dry, rocky mountain soil as found in Virginia, North and South Carolina, Tennessee, and Georgia.

The Western hemlock is considered the master tree of the genus and can attain a height well in excess of 150 feet (45 meters). The branches are slender and pendulous. The crown is narrow and pyramidal. The needles are dark green. This tree ranges widely from central California northward to Oregon, Washington, and British Columbia on into southern Alaska and eastward to the Rocky Mountains, mainly in Idaho and Montana. Along with the mountain pine, Douglas fir, white fir, and Engelmann's spruce, the Western hemlock makes up a significant portion of the forests of western United States and Canada. The tree is a major source of timber and commercially may be called West Coast hemlock, hemlock spruce, Prince Albert fir, gray fir, Western hemlock fir, or Alaskan pine. The wood has a slight pinkish tinge, is moderately soft, straight-grained, and nonresinous. Select grades are free of knots and suitable for preferred construction uses. However, although the wood is easy to work, it does not plane smoothly. Unfortunately, the wood has frequent dark streaks from heart rot, particularly common in the older trees. In the green condition, Western hemlock wood has a moisture content of 74% and weighs 41 pounds per cubic foot (657 kilograms per cubic meter). When air-dried to 12% moisture content, the weight is 29 pounds per cubic foot (465 kilograms per cubic meter); or 2,420 pounds (1098 kilograms) for 1,000 board-feet (2.36 cubic meter). The compression or crushing strength parallel to the gram is 2,990 pounds per square inch (20.6 MPa) for the green wood; 6,210 pounds per square inch (42.8 MPa) for the dried wood. The tensile strength perpendicular to the grain for the green wood is 310 pounds per square foot (1513 kilograms per square meter) and about the same for the dried wood. Commercially, hemlock timber is commonly mixed with Douglas fir. The bark of the Western hemlock contains 22% tannin. However, most hemlock-bark extract is obtained from the Eastern hemlock.

As will be noted by the names given in the prior list of hemlock species, the hemlocks also occur in Asia at similar latitudes. See also **Conifers.**

HEMOCHROMATOSIS. See **Anemias; Liver.**

HEMOCYANIN. An oxygen-absorbing substance in the plasma of the blood of the crayfish and many other arthropods. It is a clear material in the blood, but turns blue when removed and allowed to stand for a time. It serves a purpose similar to hemoglobin such as is found in higher forms of animal life.

HEMOGLOBIN. The main function of the hemoglobin molecule is oxygen transport. The hemoglobin molecules from each species of organism which has been examined differ in the sequence of amino acids in their polypeptide chains unless they are very closely related. Chimpanzee and human hemoglobins are apparently identical. Sometimes two or more different kinds of hemoglobin are found simultaneously in the same organism. These structural variations may give rise to differences in the physiological properties which help to determine the efficiency of oxygen transport by the blood from lungs or gills to the tissues. Hemoglobin also plays an important role in carbon dioxide transport. See also **Blood.**

Vertebrate hemoglobins are usually composed of four polypeptide chains of two types, called α and β. The molecules can, therefore, be described as $\alpha_2\beta_2$. An iron porphyrin moiety, *heme*, is associated with each chain. Evidence indicates that combination of the heme with oxygen results in structural changes in the protein to which it is bound. Studies of single crystals of horse and human hemoglobins by x-ray diffraction show that removal of oxygen from the iron atoms of the four hemes results in a separation of the β-chains from one another; the relative positions of the α-chains do not appear to change. Although the molecular basis is not fully understood, the consequences are important. It is certain that any change in the mutual relationships of the polypeptide chains will alter the environment of many amino acid residues. These environmental changes are probably responsible for the degree of oxygenation to the oxygen pressure; and the dependence of the oxygenation upon pH and upon carbon dioxide concentration.

Mutations which alter the amino acid sequence can occur in either the α or the β-chain of the adult. However, most mutations are deleterious and changes in the α-chain would be more severely selected against in a process of natural selection because any change in the α-chain would affect the sensitive fetus, whereas changes in the β-chain would affect only the adult. This means that evolution tends to favor changes in the β-chain over changes in the α-chain. These considerations indicate that molecular adaptation of hemoglobin, at least in mammals, may involve changes more in the β-chain than in the α-chain.

Hemoglobins can be dissociated into their α- and β-subunits. Not only are hemoglobins capable of dissociating into their polypeptide subunits, but certain hemoglobins are also capable of polymerization. Many reptiles and amphibians and certain mice possess hemoglobins which polymerize to form double molecules $(\alpha_2\beta_2)_2$ and sometimes triple or quadruple molecules. Many hemoglobins from invertebrate animals have very large molecular weights and are composed of a large number of subunits—as many as 180 in some species. The nature of the forces holding these large aggregates together is under study.

The amino acid sequences of hemoglobins have been extensively altered by mutation during evolution. Data on the amino acid sequences of the chains from a variety of mammalian and other vertebrate hemoglobins show that the sequence can be varied extensively without drastic change in function. There appears to exist a hierarchy in the functional importance of different parts of a protein. Substitutions in different segments of a polypeptide chain may, according to the type and position of the substitution, exhibit a spectrum of effects, ranging from detectable to catastrophic. For example, the single substitution of valine for glutamic acid in the 6th position of the β-chain in human sickle cell hemoglobin results in a large decrease in the solubility of deoxygenated hemoglobin within the red cells. The hemoglobin, by forming a gel, distorts the red cell shape ("sickle") in such a way that flow through the capillaries is retarded. Such drastic consequences do not result if the substitution is lysine rather than glutamic acid (hemoglobin C). Histidine in position 63 of the human β-chain has an essential role stabilizing the ferrous state of the heme iron. Substitution by tyrosine (in hemoglobins "M") results in the loss of this stability because the ferric iron can form a strong linkage with the —OH group of tyrosine. Such a substitution results in a complete loss of capacity to combine reversibly with oxygen.

The foregoing are radical substitutions. Most effective substitutions appear to be relatively conservative and do not drastically affect the oxygen transport function. Therefore, the number of differences between homologous chains appears to be related not to functional differences, but to the time which has elapsed since the chains diverged from a hypothetical polypeptide ancestor. The mean number of differences between the hemoglobin chains of man, horse, pig, rabbit, and cattle is approximately 11. The common ancestor of these mammals may have existed some 80 million years ago. Thus, approximately 11 effective mutations per chain occurred in 80 million years, or 1 substitution per chain in 7 million years. Zuckerkandl and Pauling, using standard probability theory, have used this figure to estimate the time at which the different human hemoglobin chains (α, β, γ, and δ) are believed to have arisen by gene duplication. These estimates are shown in the accompanying table.

Estimates like these indicate that hemoglobins are very old and that it may be possible to find relatives of vertebrate hemoglobins in invertebrate animals. They also suggest that the gene duplication believed to

DIVERGENCE OF HEMOGLOBIN CHAINS WITH TIME

Type of Chain Divergence	Number of Differences	Estimated Time since Divergence
β-δ	10	35 million years
β-γ	37	150 million years
β-α	76	380 million years
(α-β)-myoglobin	~135	650 million years

be responsible for the divergence of the α- and β-chains took place in the Devonian period at the time of the appearance of early amphibians and the dominance of fish.

The suggested relationship between numbers of differences and evolutionary time is not wholly secure. It assumes uniformity in the rate of effective amino acid substitution, but this rate may be neither uniform with time, nor uniform in different parts of the polypeptide chain. Differences in the rate of effective substitution along the polypeptide chain may be due not only to restrictions imposed by the required tertiary structure, but also to differences in the rate at which various parts of the DNA or the gene mutate. The evolution of hemoglobin may be contrasted with that of cytochrome *c* in which approximately 50% of the molecule appears to have remained invariant during the time yeast and man have evolved.

HEMOGLOBINURIA. The presence of hemoglobin in the urine. This occurs when red cells of the blood are destroyed at such a rate that the hemoglobin set free cannot be disposed of by the normal processes, but appears unchanged in the urine. Myohemoglobin, the pigment of muscle cells, may similarly appear in urine, especially after extensive crush injury, from mismatched blood transfusion, from allergy to the bean, *vicia faba* and in certain rare conditions, e.g., after exposure to cold in certain persons whose blood contains a hemolytic agent active only when the blood is cooled (Donath-Landsteiner reaction), in certain otherwise normal persons after exercise (march hemoglobinuria) and in a rare type of hemolytic anemia (Marchiafava-Micheli anemia) in which the hemoglobinuria occurs only in sleep. See also **Kidney and Urinary Tract.**

HEMOLYMPH. The blood of higher invertebrates, consisting of a clear plasma and white cells but without red cells. Respiratory pigments are dissolved in the plasma. It contains a lower percentage of water than the blood of more primitive forms.

HEMOLYSIS. See **Blood.**

HEMOLYTIC ANEMIAS. See **Anemias.**

HEMOPHILIA. A hereditary blood condition in which the blood fails to coagulate; an abnormal tendency to bleed. Transmitted by females, but occurs in severe form only in males. This familial blood disease has been recognized for hundreds of years. Because of the frequency of the disease in many of the royal families of Europe, particularly those of Spain and Russia, the incidence of hemophilia has changed world history. Hemophilia in women is practically unknown. A woman may carry the genetic factor producing hemophilia without having any of the symptoms; she may display the symptoms if each of her parents carried this factor for the disease. It behaves as a sex-linked Mendelian recessive. See **Heredity.** A female capable of transmitting the disease does so to about two-thirds of her male children, while two-thirds of her female offspring are conductors of the disease. See also **Sex-Linked Inheritance.**

Hemophilia is almost always apparent in the first year of life, and is generally recognized without difficulty because of its prior occurrence in the family. On rare occasions, it occurs in families which have no history of the condition. Therefore, unusually severe bleeding from a seemingly minor injury should be checked out by a physician. The hereditary nature of hemophilia should serve as a warning to members of the families in which the disease has occurred.

Hemophiliacs do not usually die from the first severe bleeding because of their reserve stores of blood cells. Subsequent hemorrhage may prove fatal. Patients who receive no medical treatment in cases of bleeding seldom live beyond their 20th year, while those who obtain proper care have an excellent chance for a long life.

The immediate treatment of patients with hemophilia is frequently self-administered. The victim should administer clot-stimulating materials, as previously prescribed by a physician, applying them directly to any cut or scratch. The usual methods for stopping blood flow have little or no effect.

If bleeding cannot be stopped by applying such substances, then an injection of antihemophilic factor VIII, the special clot-forming protein that is missing from the blood of hemophiliacs, is required. Potent doses of this protein can be prepared by freezing, thawing, and then centrifuging fresh plasma. This procedure is to be contrasted with massive plasma transfusions which may have to be repeated and carry the risk of hepatitis. VIII concentrate can be administered quickly by syringe and is especially valuable for the hemophiliac who may need an emergency operation.

Fortunately, for the hemophilic patient, there may be remissions in the disease, during which time nearly normal clotting activity will occur for weeks, or even years. A life of moderate activity with some precautions, prompt attention to bleeding, and VIII injections when necessary are the measures that will increase life span. A new dimension of concern for the hemophiliac today is the risk of being transfused with blood containing the AIDS virus. Much tighter control of blood quality, of course, has been instituted since the first awareness of AIDS.

HEMOPTYSIS. The bringing up of blood from the larynx, trachea, bronchi or lungs. The commonest cause is pulmonary tuberculosis; carcinoma of the bronchus is a frequent cause; it may also occur in any chronic bronchial or pulmonary disease and certain varieties of heart disease, especially mitral stenosis, in which the pulmonary blood pressure is persistently raised; and in aneurysm of the aorta. Other diseases which may predispose hemoptysis include polyarteritis nodosa (a subacute or chronic, remittent, disseminated vascular disease characterized by focal necrotizing inflammation of the walls of medium- and small-sized arteries and arterioles); Weil's disease (a severe form of leptospirosis); and wool sorter's disease. See also **Leptospirosis.**

HEMORRHAGE. Bleeding. Escape of blood from the vessels. Anemias caused by sudden blood loss as in traumatic injury are generally normocytic, that is, the cells are of normal size, but reduced in number. When the blood is lost over a longer period of time, as from bleeding hemorrhoids, peptic ulcer, in hookworm disease, and in excessive menstrual bleeding (menorrhagia), a microcytic anemia may result. Following hemorrhage, body fluids seep into the blood which restore it to its former volume; consequently, dilution of the blood occurs, and anemia may result. It may require some time for the body to manufacture the necessary red cells and other substances necessary to return the blood to normal. The symptoms of such a blood-loss anemia include a general weakness, dizziness, and faintness. In more severe cases, there may be vomiting and a great thirst, the heart rate may be rapid, and the breathing weak and shallow.

The first step in the treatment of persons with a posthemorrhagic anemia is to stop the loss of blood. Blood transfusions may be given to return the blood to its proper volume before excessive dilution occurs. In milder hemorrhages, however, the body may be able to restore the lost blood without transfusion. This is often accomplished by ample rest and a good diet, including adequate amounts of iron and protein necessary for red cell building.

Hemophilia, a rather rare hemorrhagic disease, is described under **Hemophilia.** Other conditions exist in which unusually large amounts of blood may be lost. In many such cases, bleeding may take place into the skin, as in a bruise. This symptom is referred to as *purpura.*

Essential thrombocytopenic purpura is a disease characterized by hemorrhage, and caused by a deficiency in the number of blood platelets. The spleen may be responsible for this disease by destroying the blood platelets. Corticosteroid therapy helps control the bleeding, and in most patients, is regarded as a desirable precaution prior to removal of the spleen.

Purpura and excessive bleeding may occur in persons suffering from deficiency of vitamins C and K. Some newborn infants contract a hemorrhagic disease which once was frequently fatal; the victims now recover rapidly when treated with vitamin K. Purpura may occur in persons receiving antitoxin treatments, or as a symptom of snakebite poisoning, or with some types of food poisoning. The taking of certain drugs may bring about abnormal bleeding. Purpura is occasionally a symptom of such varied conditions as meningitis, scarlet fever, severe measles, chronic kidney disease, endocrine disorders, liver disease, macrocytic anemias, allergies, typhus fever, and a specific bacterial heart disease. The symptom disappears in each case when the primary cause is removed.

HEMORRHOIDS. See **Arterial and Venous Disorders.**

HEMOTOXIN. See **Snake.**

HEMP. The fibers of the hemp plant, *Cannabis sativa*, of the family *Cannabinaceae* (hemp family) are coarse and rather harsh and much less pliable than flax fibers. They are dark colored and not easily bleached without damage. The fibers are used mainly for making rope and coarse twine, warp of carpet, belt and upholstery webbing, and wherever strength and durability without concern for appearance are of importance. Short fibers of hemp, called tow, are used in packing joints in pipes, for pump packing, and for stuffing upholstery. The woody waste from hemp fiber is sometimes used in the manufacture of certain papers. See also **Rope.**

Hemp is obtained from the stem pericycle of a tall hollow-stemmed annual which is a native of central and western Asia. In cultivation, the slight branching, which characterizes the plant, is considerably reduced by planting thickly. The plants grow from 5 to 16 feet high. They have digitately compound dark green leaves and small inconspicuous flowers, which are of two kinds, occurring on different plants. The staminate flowers appear in small axillary clusters on male plants, and the pistillate flowers are borne in leafy spikes on female plants. The fruit, an achene, is a hard ovoid structure, often called hemp seed. Hemp grows best in regions having a warm humid growing season of about 5 months. The plants grow rapidly, soon shading the ground so effectively as to suppress other plants, and thus plantings of hemp have been used as a means to eradicate weeds. When the staminate flowers are mature, the plants are ready for harvest. To delay after that is not desirable, since the male plants die soon after flowering. After flowering, the fibers become coarser. Harvesting and the treatment of the plants after harvesting are similar to the procedures used with flax plants. See **Flax.** The hemp plants are cut off or pulled up, denuded of leaves, roots and tops, and tied in bunches and left to dry for about 2 weeks. They are then immersed in water to ret. In retting, the intercellular substance of the stems is acted upon by bacteria and softened so that the fibers are readily cleaned of surrounding tissues. Scutching removes the woody tissue, after which the rough hemp fibers are hackled, or drawn over coarse combs which pull out the fibers.

In recent years, the cultivation of hemp has been subject to controls because marijuana is prepared from the dried leaves and flowers of the plant, which then are smoked in the form of cigarettes as a narcotic. The species *Cannabis indica* is usually used in this connection. See **Marijuana.**

HENNA SHRUB. Of the family *Lythraceae*, the *Lawsonia inermis* is a small shrub native to Africa and Asia and cultivated in tropical countries. Well known since the time of the early Egyptians as a red dye for hair, nails, hoofs of animals, etc., the leaves of this plant are powdered and made into a paste which is then used as a dyeing medium. The small flowers of the plant are inconspicuous, but fragrant.

HEPARIN. A complex organic acid (mucopolysaccharide) present in mammalian tissue; a strong inhibitor of blood coagulation. Precise chemical formula has not been fully established, but the formula $(C_{12}H_{16}NS_2Na_3)_{20}$, with a molecular weight of 12,000, has been suggested for sodium heparinate. The drug is derived from animal livers or lungs. Heparin is used in deep venous thrombosis therapy. It is also used in rodenticides which cause internal hemorrhaging. Pets exposed to such poisons must receive immediate treatment with the administration of vitamin K, also sometimes called the antihemorrhagic vitamin. See **Anticoagulants; Vitamin K.**

HEPATITIS. See **Liver; Virus.**

HERBICIDE. A substance that kills or interferes markedly in the life cycle of certain plants and is used with other control chemicals, such as fungicides and insecticides, to increase the yield and quality of crops. In addition to eliminating or greatly stunting the growth of those plants (weeds) that compete with crops for water and soil nutrients, herbicides achieve a number of objectives. Usually lumped under the phrase *weed control*, the advantages of herbicides include: (1) Eliminate weeds that serve as harboring places for insects which attack crop plants. See accompanying illustration. (2) Eliminate perennial plants that may serve as hosts for survival and build-up of virus diseases. An example is the corn stunt virus, which overwinters in johnson-grass rhizomes. Insects able to carry virus diseases may feed on weeds and move to crop plants, causing infection by damaging virus diseases. (3) Eliminate weeds that serve as traps for moisture. Easy availability of moisture encourages fungus diseases that can be spread easily from weeds to crop plants by wind movement of fungus spores. (4) Eliminate honey-suckle, kudzu, and other plants that grow on fences and that are severely damaging to fencing and other minor structures because of the sheer weight of their foliage. Damaged fences adversely affect livestock production. (5) In so-called "no-till" planting, herbicides are used exclusively, eliminating the mechanical removal of weeds by cultivating equipment. The advantages of herbicides in this regard are time and labor savings.

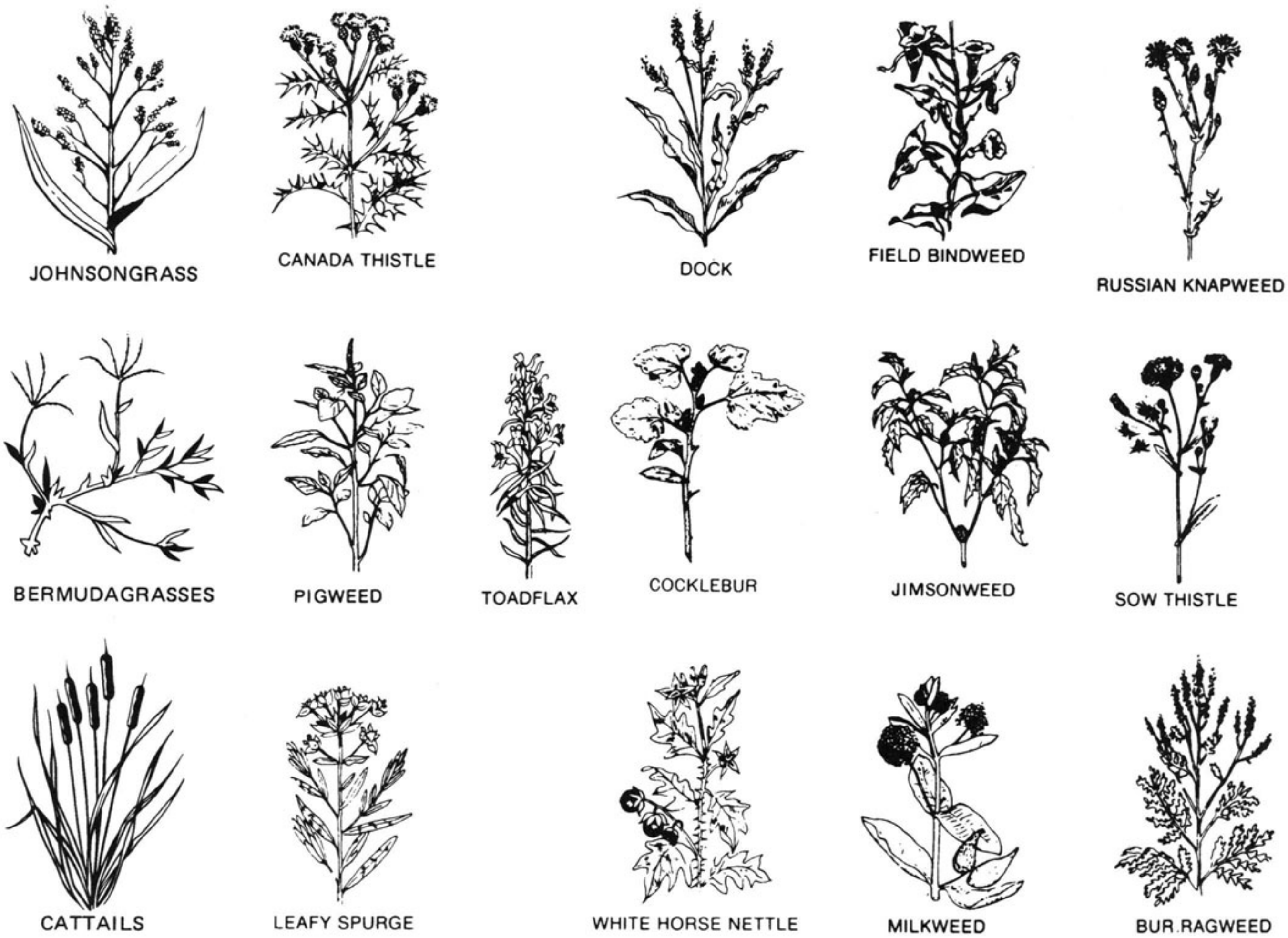
Classification of Herbicides

There are several ways in which herbicides can be grouped.

Target plant selectivity is a measure of the effectiveness of a herbicide against a range of plants to be destroyed. *Nonselective* herbicides are not difficult to create. There are hundreds of chemicals that will kill just about any living plant within range. These are extremely *widespectrum* substances, not only destroying or stunting both broadleaf and grasslike weeds, but woody plants as well. Of course, some control over these very powerful chemical substances can be exerted by regulating concentration. Dilute applications may result in desired defoliating, for example, without fully destroying a stand of plants, such as trees. *Broad-spectrum* and nonselective herbicides are sometimes regarded as the same, but more generally, broad-spectrum refers to a compound that does not differentiate between broad- and narrowleaf plants. A *selective* or *narrow-spectrum* herbicide is customized to make this selection and, in fact, to differentiate even more closely. In operating with crop-rotation programs, it is usually advantageous to select herbicides with relatively narrow spectrums of effectiveness so that later crops may not be adversely affected by any residues from a prior crop. Herbicide manufacturers continue in their research toward the development of crop-specific control chemicals. Just one example—a herbicide to control wild oats in connection with wheat production.

Timing. Herbicides are usually designated for a *pre-plant* or *pre-emergence* use or for *post-emergence* application. The terms tend to be self-descriptive—pre-emergence signifying the use of the herbicide on the land prior to the cracking stage or emergence of weeds or desired crop above the soil line. The herbicide, possibly in granule or liquid form, may be incorporated into the soil a number of weeks before planting, in which case, the term pre-plant is used. For effective control over the growing season, some land areas or crops may have to be treated a number of times between emergence and harvest. Thus, the term post-emergence. Several days must elapse between application and harvesting to avoid contamination of the crop when gathered.

Stability. A number of factors determine the stability of a herbicide. For example, some of the control chemicals decompose (and thus become ineffective) when applied at temperatures in excess of about 90°F (32.3°C), or during long periods of intense sunlight. Most herbicides are more efficient when applied on cool, partly cloudy days. In the case of many other herbicides, the presence of moisture (after an irrigation, rain, or during a generally wet period) greatly reduces or destroys their effectiveness. The presence of certain chemicals also affects application success. Some control chemicals are adversely affected by any mixture with acidic materials; others by the presence of alkaline materials, or certain metals, such as iron or copper. Essentially, these materials interact with the original chemical composition and alter it so that, as a result, instead of applying an effective herbicide, for this purpose the material may be essentially inert. Careful preparation, particularly of emulsions for spraying, cannot be overemphasized. If the stability



Important target grasses and weeds that can be controlled by herbicides. (*The Dow Chemical Company.*)

and effectiveness of a herbicide is long term, it can be designated as a *persistent herbicide*, meaning effectiveness over a period of several months. Those substances that break down within several days to a few weeks (biodegradable in a sense) are *nonpersistent herbicides*. This is an important factor in selecting a herbicide. Some control chemicals can essentially sterilize a plot of land for a period of years, and should one's objectives change after such an application, neutralization or removal of the substance from the affected area can be costly and quite difficult.

All control chemicals can be categorized as either *contact-type* or *systemic* substances. Although these designations are more commonly applied to insecticides, they also are operable in terms of herbicides. In a contact-type substance, the killing action is largely limited to the area of actual contact between chemical and plant (for example, a defoliant that damages a bush or plant without completely destroying the whole plant). In a systemic substance, contact of the substance with part of the plant is progressively spread throughout the plant as, for example, by the plant's vascular system.

Physical Form. A large number of herbicides are available in several forms, including granules, powders and dusts, wettable powders, emulsifiable concentrates, slurries, etc. Some of these are factory-prepared; others can be prepared locally by the user. Sprays and dusts are widely used for foliar applications, whereas dusts and granules may be preferred for soil applications.

Chemical Structure. As with fungicides, insecticides, and other pesticide chemicals, herbicides are usually complex, often synthetic organic chemicals and of a widely varying composition, ranging from carbamates, to anilides, to organic acids, salts, etc. Chemical make-up is discussed further a bit later in this entry.

Nomenclature of Herbicides. There are well over 100,000 pesticides and agricultural control chemical formulations; perhaps 25% of these fall into the sphere of herbicides. As with insecticides, although the basic chemicals used in the formulation of herbicides may number in the several hundreds, many thousands of possible formulations arise from the various physical formats offered, as well as minor differences provided by many manufacturers in brand name products. Each manufacturer markets products under trade names—names that are essentially coined for their marketing charisma and infrequently connoting much about the content or purpose of the product. Thus, there are scores of equivalent (or essentially equivalent) products, adding to the difficulty of selecting these chemicals. Unfortunately, from this standpoint, the generic chemical names of the majority of herbicide chemicals are long and complex and essentially meaningless to persons who are not well versed in organic and biochemistry. Helpful listings of this type can be found in the "Foods and Food Production Encyclopedia" (D. M. Considine, editor), Van Nostrand Reinhold, New York, 1982. There are also a number of frequently revised directories of control chemicals and considerable information available from various government agencies and universities. See list of references at end of this entry. This situation of nomenclature is quite similar to that which applies to generic and trade name drugs and pharmaceuticals.

Chemistry of Herbicides

Aromatic Carboxylic Acids. Considerable research has gone into investigations of the physiological activity of these acids on plants, including benzoic, phenylacetic, and naphthoic acids. Among the benzoic acid derivatives, the greatest activity is shown by those compounds containing substituents in the 2, 3, and 6 positions; and only to a slightly lesser degree, by those substituents in other positions. Included among commercial herbicides in this category are: 2,3,6-trichlorobenzoic acid; 2-methoxy-3,6-dichlorobenzoic acid; 2,5-

dichloro-3-nitrobenzoic acid; and 2,5-dichloro-3-aminobenzoic acids. Slightly less active are: 2-bromo-3,5-dichlorobenzoic acid; and 2,3,5-triiodobenzoic acid.

Substituted phenylacetic acids have a high activity. Considerable activity is shown by monohalogen-substituted acids. Introduction of a second halogen does not markedly affect the degree of activity.

Aryloxyalkylcarboxylic Acids and Derivatives. Research has shown that the physiological activity of phenoxyacetic acid toward plants increases when a halogen atom is incorporated into the molecule. The strongest effects are displayed by fluorine and chlorine. The position of the substituent also affects physiological power, with the 4-halophenoxyacetic acids displaying the greatest activity. It is interesting to note that the activity of this compound is about ten times greater than the case of the 2-isomer. And, the activity is further reduced in the 3-chlorophenoxyacetic acid. There are numerous herbicides in this chemical structural category, including 2,4-D, 3,4-D, and MCPA (4-chloro-2-methylphenoxyacetic acid). It is also interesting to note that while MCPA is very effective as an agricultural control chemical, the very closely related compound, 4-chloro-2-chloromethyl-phenoxyacetic acid is not of great value.

Derivatives of Carbamic Acid. Whereas the aryl esters of *N*-methylcarbamic acid find wide application as insecticides, the alkyl esters are strong herbicides against monocotyledonous weeds. Their actions against dicotyledonous plants are much weaker. Because of these differences, these herbicides are effective in controlling monocotyledonous weeds in such crops as carrot, cotton, and sugar beet. Research has shown that the esters of naphthylcarbamic, diphenylcarbamic, and other polycyclocarbamic acids are not effective herbicides. It has been found that the arylcarbamic acid ester derivatives of unsaturated alcohols are stronger herbicides than the corresponding esters of saturated alcohols. The carbamates have an ability to form hydrogen bonds with the chlorophyll molecule or proteins of plants, accounting for their effective herbicidal activity.

Derivatives of Thio- and Dithiocarbamic Acids. Research has indicated that the derivatives of the thiocarbamic acids are good penetrants of plants, moving easily through the xylem. Among this structural class of herbicides, the *S*-alkyl-*N*-dialkylthiocarbamates are the most effective. Most of these compounds are selective herbicides against annual grasses and a few dicotyledons. They have been applied successfully in connection with such crops as bean, beet, other vegetables, and sugarcane. The thiocarbamates are usually mixed with the soil as pre-emergence herbicides. In terms of effectiveness, this usually decreases as the number of carbon atoms in the ester radical increases, particularly in excess of five carbon atoms. The activity also decreases when the total carbon atoms in the alkyl radicals on the nitrogen atom is greater than six.

Derivatives of Urea and Thiourea. A great deal of investigation has gone into the effectiveness of these compounds and, as the result, a number of urea derivatives have found use as effective herbicides as well as growth regulators. This is particularly true among the trialkylureas that contain simple and complex hydrocarbon radicals. Examples include: 3-(3,4-dichlorophenyl)-1,1-dimethylurea (Diuron); 3-(3,4-dichlorophenyl)-1-methoxy-1-methylurea (Linuron);1,3-dimethyl -3-3 (2 benzothiazoyl) urea (Methabenzthiazuron); 3-(4-bromophenyl)-1-methoxy-1-methylurea (*Metobromuron*); and *N*-benzyl-*N*-(dichloro-3,4-phenyl)-*N*,*N*-dimethylurea (*Phenobenzuron*).

The salts of aryldialkylureas tend to be more active than the ureas.

Thiocyanates and Isothiocyanates. At one time, ammonium thiocyanate was a commonly applied, nonselective, contact-type herbicide and desiccant. The compound is less important now because of the development of other organic herbicides that do not decompose so readily.

Sulfuric and Sulfurous Acid Derivatives. Earlier, sulfuric acid was applied as a herbicide and still finds some use as a desiccant for potato plant tops prior to mechanical harvesting. The primary drawback of sulfuric acid, not experienced with more recently developed herbicides, is the large amount of acidity which it adds to the soil. Ammonium sulfamate continues to find use as an effective herbicide in some areas, both for the elimination of weeds and as a sterilant for soil. This compound hydrolyzes in soil to form ammonium sulfate, a source of ammonia and nitrogen.

Several other classes of organic chemicals, including heterocyclic compounds, are represented among the scores of herbicides commercially available. As of the late 1970s, in terms of tonnage usage, the Food and Agriculture Organization (United Nations) listed the following major categories: MCPA, 2,4-D, 2,4,5-T, triazines, carbamates, and urea derivatives. It should be stressed here that regulations pertaining to the use of herbicides vary widely from one country to another—and from one time period to another. The proliferation of new products tends to offset those prior compounds that have been banned in some countries, or where usage has been severely curtailed.

HERBS. See **Composite Family.**

HERCULES. A large constellation lying between Lyra and Corona Borealis. Hercules contains no strikingly bright stars, and hence is somewhat difficult to locate. Once found, however, it is a fertile field for a small telescope. In 1934, this constellation received considerable notice because of the brilliant nova that appeared in it just before Christmas. Perhaps the most interesting object within it is a remarkable star cluster, which was first noted by Halley, in 1714. Although this cluster can be distinguished as such in a telescope of only 2-inch aperture, it requires a telescope larger than 6 inches to appreciate the magnificence of the object. (See map accompanying entry on **Constellations.**)

HEREDITARY MECHANICS. The field of mechanics involving boundary conditions extending over continuous intervals of space and time and demanding integrals for their representation. For example, in the application of stress to a deformable elastic medium, the final strain at any instant depends not only on the stress at that instant but on the whole previous stress to which the medium has been exposed. Analytically,

$$\delta(t) = kX(t) + \int_{t0}^{t} \theta(t, \tau)X(\tau)\, d\tau$$

where δ is the final strain at time t, $X(t)$ is the instantaneous stress at time t and the integral represents the effect of the stress heredity of the system. The quantity $\theta(t, \tau)$ is called the coefficient of heredity. The above equation may be considered an integral equation for the evaluation of X when δ is known.

HEREDITY. The transmission of developmental potentialities from one generation of living things to the next and following generations through the natural process of reproduction. The materials of the parent bodies from which a new individual develops are its actual heritage. During its own embryonic development, the potentialities of this heritage are expressed in the structural characteristics of the new body, normally like those of the parents or those of a more remote generation of ancestors. This fact leads to the statement that the organism inherits certain characters; although this may not be precisely true, the interpretation is permissible for ordinary purposes of description.

Genetics is that branch of biology which deals with the phenomena of heredity and the variations between parents and offspring. See also **Genes and Genetics.**

Work of Mendel. The first steps in genetics were taken by plant hybridizers of the eighteenth and nineteenth centuries, chiefly in Europe, and culminated in the experiments of Gregor Johann Mendel, a monk at Brŋo, Czechoslovakia, then Brünn in Austria. Mendel's results were published in 1866 and lay almost unnoticed until 1990, when they were corroborated by three scientists in the birth of modern genetics. The published report of Mendel's work repeated the significant observations of his predecessors and added a simple mathematical analysis that had not been previously expressed. As a result of the importance of this work, the term Mendelian heredity is applied to the established fundamentals with which subsequent discoveries have been correlated.

Mendelian heredity depends on three fundamental concepts: (1) The organism is a mosaic of unit characters capable of separate hereditary transmission. (2) A unit character may mask a related unit character completely when the potentialities for the development of both are present in the same individual. This principle is called dominance, and the masked character is said to be recessive. (3) Unit characters may be segregated during reproduction, regardless of the combinations in which they have been associated.

To these concepts, later scientists in the field added that the association of different related unit characters in one individual may result in the development of both in different parts of the body, in a mosaic inheritance, or in an intermediate condition.

Some characters, particularly of a quantitative nature, are due to multiple genes. Such characters must be studied by statistical methods. They were the foundation of another attempt to formulate laws of inheritance made by Sir Francis Galton, from which we retain the law of ancestral inheritance and the law of filial regression. The former indicates that each parent contributes one-quarter of the total heritage of the individual, each grand-parent one-sixteenth, and so on in a rapidly diminishing percentage. The law indicates the great reduction of the possibility of a hereditary character reappearing after a lapse of generations. Filial regression is the tendency of extreme parents to produce offspring less extreme than themselves. Thus tall parents beget tall children, but usually shorter than themselves. Galton studied human inheritance and in addition to his mathematical analyses, so necessary in this field, took the initial steps in proposing deliberate control, which led to the science of eugenics.

Modern gene science has added vast amounts of quantitative information, strengthening the recognition that hereditary potentialities are resident in the chromosomes of body cells and that definitely located genes within these chromosomes are the determiners through which specific unit characters are brought to expression. The behavior of chromosomes has been found to be in harmony with the transmission of characters by Mendelian heredity. Since nothing was known of chromosomes during Mendel's life, this correlation had to await further advances in cytology.

Mendel's chief contributions were derived from the study of garden peas, in which he observed seven pairs of unit characters, all similar in behavior. He noted, for example, that seed colors included two unit characters, yellow and green. When he crossed parent plants of the two strains the resulting hybrid seeds were entirely yellow, indicating the dominance of this color over green. He then inbred the hybrids, and in their offspring both yellow and green seeds appeared in the ratio of three yellow to one green. Related unit characters of this kind are said to be alleles or allelomorphs. It is now known that their genes occupy the same position in the paired chromosomes of the cells, while only one can be represented in the single chromosome of a germ cell. Since each parent contributes one chromosome to each pair in its offspring, it may also contribute one gene of an allelic pair. The one parent plant contributed a gene for yellow, the other for green, and through dominance the offspring were yellow. Segregation, however, enabled these hybrids to transmit either yellow or green during their reproduction, and through random fertilization all possible combinations of these determiners were established. The characters are commonly represented by symbols, using a capital letter for the dominant and a small letter for the related recessive, as Y and *y* for yellow and green, respectively. For the pair of characters mentioned, the following diagram is representative:

Parental generation (P):	YY	yy
Germ cells:	Y	y
Hybrids of first filial generation (F_1):	Yy	
Gametes of F_1 generation	Y	y

and their combinations in the F_2 generation, in a Punnett square:

	Y	y
Y	YY	Yy
y	Yy	yy

The YY and *yy* individuals in this diagram are homozygous, and the Y*y* individuals are heterozygous. Since all YY and Y*y* individuals look alike, due to the dominance of Y, they belong to the same phenotype, but since their hereditary potentialities are different they belong to different genotypes. The *yy* individuals from hybrid parents are known as extracted recessives. There are twice as many heterozygotes as homozygotes of either kind in this 3:1 ratio because similar individuals in this category result from reciprocal combinations of genes, half of the individuals receiving the dominant from one parent and half from the other. Examples of this kind, involving only one pair of allelic characters, are known as monohybrids.

Additional complexity arises in dihybrids, trihybrids, and polyhybrids of still more characters through the free reassortment of the unrelated pairs of alleles. Thus peas from smooth yellow seeds crossed with others from wrinkled green seeds, a dihybrid combination, produce only yellow smooth seeds in the F_1 generation, but when inbred these plants give rise in the F_2 generation to the four possible combinations: smooth yellow, smooth green, wrinkled yellow, and wrinkled green, in the ratio 9:3:3:1. The reason is evident in the following diagram:

	SY	Sy	sY	sy
SY	SY SY	Sy SY	sY SY	sy SY
Sy	SY Sy	Sy Sy	sY Sy	sy Sy
sY	SY sY	Sy sY	sY sY	sy sY
sy	SY sy	Sy sy	sY sy	sy sy

In this diagram, each pair of symbols above and at the left side represents the contribution of one parent in one of its germ cells, and in the small squares the possible combinations from the two parents are shown. Dominance prevails as in the monohybrid.

In a trihybrid, free reassortment results in an F_2 ratio of 27:9:9:9:3:3:3:1. The number of phenotypes is always a power of two indicated by the number of pairs of alleles under consideration. See also **Cell (Biology).**

Studies of Fruit Fly. The study of heredity in animals has shown that these principles are applicable in that kingdom as well as in plants, but relatively few animals are sufficiently prolific to demonstrate complex ratios. The fruit fly, *Drosophila melanogaster*, has been the most productive of all genetic subjects, whereas man and the domestic animals yield very limited Mendelian data.

Modern genetics, largely from studies of the fruit fly, has disclosed many principles as corollaries of simple Mendelian heredity. The more important are as follows:

Multiple alleles: More than two unit characters may be related to each other as alleles. In such cases only two of the series may be present in any one individual, and dominance is in a graded series, as may be determined by experimental results.

Multiple genes: More than one gene may be necessary for the production of a single unit character. If two genes are essential for its appearance and either alone is incapable of expression, they are said to be complementary. If one expresses itself alone, a gene that modifies this expression is supplementary. If two are capable of producing the same effect whether present singly or in combination, so that the resulting character is absent only from homozygous recessives, they are said to be duplicate genes. In all cases, recombination of the genes during reproduction follows the same course as in simple Mendelian heredity, but the resulting phenotypic ratios differ because fewer unit characters are involved.

Lethal genes: Some genes completely inhibit development or modify it in such a way that the individual dies. They also modify the usual ratios of associated characters.

Linkage: Some characters, although not allelic, are inherited in definite groups; they are said to be linked. Modern genetics shows that linkage is due to the presence of genes for the linked characters in the same chromosomes.

Crossing over: Linkage relations are sometimes interrupted in a limited number of individuals, permitting some reassortment of normally grouped characters. This change is due to the breaking of paired chromosomes in synapsis and the reunion of their fragments in new combinations to form similar chromosomes, sometimes with new combinations of genes.

Translocation: This change is a shifting of the relations of genes in the chromosomes, due to looping, fusion, and rupture, or to the attachment of fragments to other chromosomes. It may result in the duplication of genes within a chromosome or in a change in the serial arrangement of the included genes.

The inheritance of sex has also been shown in many cases to depend on a simple chromosomal mechanism. Males of many species have an X chromosome without a synaptic mate or with a Y chromosome mate that is evidently abortive. The females of such species have two X chromosomes. In the formation of germ cells all eggs receive an X chromosome while half of the sperm cells receive an X chromosome and half a Y or none. Random combination of these cells restores the XX combination in one-half and X or XY in the other, thus producing half females and half males. Other investigations have shown that the quantitative balance between the sex and other chromosomes is the active factor in conditioning the differentiation of the sexes.

This disclosure also explains the phenomenon of sex linkage. Genes lying in the sex chromosomes, mostly in the X chromosomes but a few in the Y, are inevitably transmitted and expressed in some definite relation with sex; hence they are said to be sex linked. Such characters need have no active sexual role.

Plant and Animal Breeding. The findings of genetics have been of value in plant and animal breeding. Although the improvement of cultivated plants and domestic animals by selection preceded by many years the formulation of scientific principles of heredity, the discovery of these principles has made possible much more precise and efficient procedure in the establishment of useful strains. Hybridization and selection together are the chief means of improvement. Applied by scientists they have brought about many modifications of living things and have disclosed many facts concerning heredity. Corn (maize) has been studied in detail and subjected to many experiments, both practical and purely scientific. Tomatoes, radishes, various cereals, and flowers of many species have also commanded attention. More has been done with plants than with animals because the domestic animals are less amenable to experiment. From the practical point of view plants are more satisfactory subjects because desirable hybrid strains may often be propagated by cuttings, grafting, and other asexual methods which avoid the segregation that is inevitable in sexual processes. Only rigid selection can establish desired hybrid combinations in plants or animals that must be produced sexually.

The study of human heredity depends on studies of families. Genealogical records have furnished a large amount of valuable material and the records of public institutions have been equally useful to the geneticist. Such records are not to be compared with scientifically assembled experimental data, but they leave no doubt that the principles of heredity worked out in the study of other organisms are also applicable to humans.

The clearest evidences of human heredity are found in the behavior of simple structural defects, such as the appearance of extra digits (polydactylism), the fusion of bones in the digits (symphalangism), and shortness of the fingers (brachydactylism). These defects are transmitted as Mendelian unit characters allelic to normal structure. Red-green color-blindness (vision) is one of the most striking examples of inheritance in man. It is a sex-linked recessive allele of normal vision. Both X chromosomes of the female must carry the gene for the defect if she is to be color-blind, whereas the male may be color-blind if he receives such a gene in his one X chromosome. Females may be heterozygous carriers of the defect, with normal vision; males are either strictly normal or defective. In this type of inheritance the male always receives the genes for his characters from his mother; therefore a carrier mother may have some color-blind sons. A color-blind man and a genotypically normal woman cannot produce color-blind children, but all of their daughters are carriers. On the other hand, a color-blind woman and a normal man will produce carrier daughters and color-blind sons. Hemophilia is inherited in a like manner, except that the recessive genes for hemophilia are lethal in the homozygous condition in the female. See also **Hemophilia.**

Pigmentation of the skin is controlled by multiple-factor inheritance. Since variations of skin color within a race are not always readily identified and may be partly environmental, knowledge of skin color inheritance has had to come mainly from study of black-white marriages. When a black person without white ancestry marries a white person without black ancestry, their children are typically intermediate in color, or mulattoes. Children from the marriage of a typical mulatto to another typical mulatto may vary in skin color from the black of the black grandparent to the light color of the white grandparent. It has been estimated that the color differences in blacks and whites are controlled by from two to four pairs of alleles. It is possible for a white-skinned person of black-white ancestry to have all the genes of the white genotype. Children from such a person married to a white or similar near-white should be all-white. Children from the marriage between two near-whites are seldom much darker than their parents, and some would have light skin color. If a near-white marries a white, their children are usually no darker than their near-white parents; there is no well-established evidence that a very dark or black child could be born to them.

Albinism is a rare inherited condition in which the skin, hair, and eyes lack the melanin pigment normally present. It results from a biochemical deficiency in which specialized skin cells called melanocytes are unable to synthesize melanin from the amino acid tyrosine. See also **Albinism.**

There is evidence that some allergies may be inherited. Inherited weaknesses in the tissues may make it easier for some antigens to enter the body of certain persons. See also **Allergy.**

Diseases of genetic origin are discussed in the entry on **Gene Science.**

Extranuclear Inheritance. The existence of cytoplasmic genes was suggested as long ago as 1909 when the first examples of non-Mendelian inheritance were described by Correns and Baur. However, the demonstration that chloroplasts, mitochondria, and the kinetoplasts of trypanosomes contain specific DNA of their own came as a surprise to most biologists. It is now recognized that organelle DNAs are present in the cell in small amounts, perhaps 1-10% of the total cellular DNA. Organelle DNAs are also distinct entities, as indicated by average nucleotide compositions different from nuclear DNA. All organelle DNAs examined thus far consist of covalently closed circles and exhibit autonomous replication. Although the functions of such organelle DNAs remain largely unknown, it appears that ribosomal RNAs and most if not all tRNAs of chloroplasts and mitochondria are transcribed from the corresponding DNAs. Specific proteins either coded by organelle genes or synthesized with the organelle have been more difficult to identify. The importance to the cell of organelle DNA and the resultant extranuclear inheritance of genes present in this DNA is illustrated by the petite mutants of yeasts. These mutants contain an altered mitochondrial DNA which results in lack of mitochondrial respiratory function. Thus, to survive, petite mutants must utilize an alternative source of energy such as anaerobic fermentation of carbohydrates. Genetic analysis has established that inheritance patterns of the defect are consistent with cytoplasmic inheritance.

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HERMAPHRODITE. An animal with functional reproductive organs of both sexes.

HERMAPHRODITISM. A condition characterized by the presence of both ovarian and testicular tissue. Because of overactivity of the adrenal glands, excessive hormones can be produced. In some patients with overactive adrenals, there is an excessive development of fat, accompanied by sexual disturbances. The symptoms vary according to age and sex. If the disease develops during fetal life and the child is a female, a form of hermaphroditism, or dual sexuality, may result, in which the clitoris is enlarged and resembles the penis. Other signs of masculinization accompany this condition. Sometimes a true hermaphrodite may appear to be a normal female, but who is found at surgery to possess testes in the groin region. Only about a dozen cases of true hermaphroditism in the human race have been reported. This term signifies the presence of all of the functioning genital organs of both sexes in one individual. The reported cases were claimed to have both testicles and ovaries present. However, the ability to impregnate as well as to conceive has never been reported in one individual.

Many cases of pseudo-hermaphroditism have been seen. In this condition the genital organs, internal or external, do not conform either totally or in part with the sexual glands (testicles or ovaries) present. In the male hermaphrodite, testicles are present but may be abdominal in position. The penis is small and more nearly resembles a large clitoris; the scrotum is divided by a cleft resembling the female labia with a small short vagina. Uterus and tubes are not present.

The female hermaphrodite has a large clitoris more like a small penis, rudimentary vagina, a uterus and ovaries. Various in-between stages may be present, given a very bizarre picture where the sex can only be determined by microscopic study of sex characteristics shown by the nuclei of the tissue cells. Such cells may be examined by biopsy of the skin, and a definite decision as to sex given with accuracy of a high degree; where biopsy is not desired or facilities are not available, the nuclei of epithelial cells scraped from the inside of the mouth, or even of polymorphonuclear leucocytes in the blood will furnish a slightly less reliable answer. Such sexing should be done as soon as possible after birth in any infant in whom the identity of the sex organs appears dubious; by this means mistakes in naming and upbringing can be avoided. Where such a decision as to sex is not made in very early life, it is probably wise to bring up the child according to the sex which seems most apparent and defer final decision until the onset of puberty, when the development of sex consciousness may reveal psychological orientation to one sex or the other.

Dewald et al. (Mayo Clinic and Mayo Foundation), using chromosome heteromorphisms and blood cell types as genetic markers, demonstrated chimerism in a chi46, XX/46,XY true hermaphrodite. The pattern of inheritance of the chromosome heteromorphisms indicated that this individual was probably conceived by the fertilization, by two different spermatozoa, of an ovum and the second meiotic division polar body derived from the ovum and subsequent fusion of the two zygotes. A chimera may be defined as an individual with two or more genetic cell types resulting from the fusion of different zygotes. As described by Dewald et al., "Chimeras can be readily classified as wholebody or partial chimeras according to their mode of origin. Partial chimeras can arise by placental cross-fertilization between dizygotic twins, maternal-fetal transplacental exchange, transfusions, or grafting. Because of lack of suitable studies, the origin of wholebody chimeras is less clear. Theoretically, they can arise by (1) early fusion of different embryos, (2) fertilization of an ovum and any polar body by two different sperm and subsequent fusion of the zygotes, (3) fertilization of a haploid ovum or polar body and subsequent fusion with a diploid polar body or ovum, or (4) fusion of a diploid sperm with an embryo." Most reported chimeras have sexual abnormalities, such as clitoral hypertrophy or true hermaphroditism. More detail will be found in Dewald, G., et al.: "Origin of chi46, XX/46,XY Chimerism in Human True Hermaphrodite," *Science*, **207**, 321–323 (1980).

HERMITE EQUATION. A second-order differential equation

$$y'' - 2xy' + 2ny = 0$$

where n is a constant. The Hermite polynomials (see **Generating function**) are solutions. The equation occurs in the quantum mechanical problem of the harmonic oscillator. (See also **Weber Equation,** from which the Hermite equation can be obtained by a change of variable).

HERNIA. At one time commonly called rupture, a hernia is an abnormal protrusion of a part or organ through the containing wall of its cavity. In common usage, the term hernia usually applies to the abdominal cavity and implies a covering or sac over the protrusion. There are two main classes of hernias: (1) *congenital* hernia in which the sac was present before birth; and (2) *acquired* hernia in which the sac is formed after birth and pushes through an opening in the muscle wall which failed to close at birth, or that was formed following an incision. A large percentage of acquired hernias result from injury or strain, such as those hernias which occur when a person lifts a heavy object. Hernias may occur in the groin, the navel, the membrane separating the abdominal and chest cavities (diaphragm), in surgical incisions, and elsewhere. All herniation takes place through a normal opening, or through an opening that should have been eliminated at some period of development, or through an opening which had closed and then reopened in later life.

The hernial sac has a mouth, a neck, and a body. The mouth connects with the abdominal cavity and is called the hernial ring; the body is the pouch or sac that projects outside the abdominal wall; and the neck connects the mouth and body of the sac.

The contents of the sac might be any of the abdominal organs, in whole or part; loops of the intestine are commonly found in hernias. The sac and its contents are subject to injury which can lead to serious complications. The skin surface is vulnerable to blows, falls, pressure, irritation from binders or trusses, or may become inflamed, infected, or abscessed. From within, the contents of the sac are prone to strangulation when the blood supply is cut off by a narrow or constricted hernial ring; gangrene may set in if treatment is not sought promptly.

Hernias are considered *reducible* or *irreducible*. Reduction may be spontaneous; for example, sac contents may return unaided to the abdominal cavity when the patient lies flat. If the patient remains untreated, however, a reducible hernia may become irreducible, that is, the contents of the sac can no longer be returned to the abdominal cavity. Irreducibility may be caused by increased size of the hernia, formation of adhesions, or development of a small or constricted hernial ring. Hernias of enormous size, hanging down to the knees, have been reported. An irreducible hernia is a constant source of danger.

Hernias occurring in the groin are either *inguinal* hernias or *femoral* hernias. Inguinal hernias account for about 92% of all hernias. Superficially, inguinal and femoral hernias look alike because the bulge is in the groin. However, they differ anatomically. Inguinal hernias slip through the normal openings for the passage of nerves or organs of the reproductive system. Femoral hernias occur through the passageway for nerves and vessels to the thigh.

Normally, the deep and shallow layers of muscles and ligaments on the abdominal wall protect these normal openings against herniation. With rise of intra-abdominal tension, as by straining, coughing, or lifting, the muscles contract and flatten like a shutter in a normal situation. But if the muscles and/or other protective structures of these openings are weak, the shutter action fails and an increase of intra-abdominal tension may push part of the abdominal organs through the opening into the preformed sac, and thus a hernia is begun. Successive incidents of tension increase the size of the sac by forcing additional intra-abdominal tissue into it.

Hernias of the navel are called *umbilical* hernias. The navel is an opening that should close in the process of development. After birth, it is a scar formed of interlaced muscle fibers of the contracted umbilical ring. Sometimes a defect occurring before birth prevents its closing, and the baby is born with a hernia, or may soon acquire one. In adults between 25 and 40 years of age, obesity and pregnancy are the most common predisposing causes of this form of hernia.

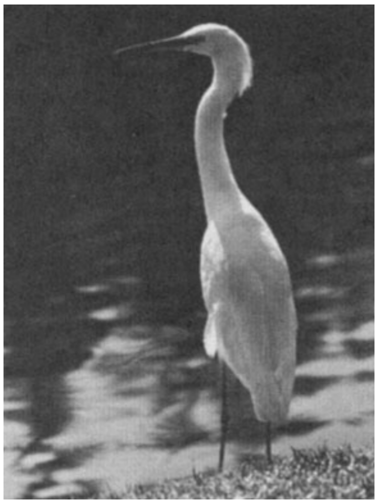
Hiatus hernia is described under **Esophagus.**

Obese women are the most frequent subjects of hernia in the site of a surgical incision. Some incisional hernias are caused by failure of the layers of deep muscle and fascia to knit firmly after surgery. Blood clot, infection, exudate, and swelling in the line of incision, as well as increased intra-abdominal tension, also are factors favoring herniation. The neck of the incisional hernia is a firm ring of scar tissue. Because of the large hernial ring, these hernias are difficult to control by a truss. Large incisional hernias may cause invalidism unless surgical relief is obtained.

The treatment of a hernia patient can be accomplished by a mechanical device (truss), or surgery. Most authorities agree that a truss is a makeshift which is acceptable only when surgery would be hazardous. Improved techniques have made possible the surgical repair of hernias which not many decades ago would have been irreparable. It is often necessary to close the opening with a fascial graft, or an inert foreign material, such as polypropylene mesh.

HERON (*Aves, Ciconiiformes*) Long-legged wading birds (*Aves*) with a sharp slender beak and when adult with plumes or a crest. They live chiefly on fish.

Herons are found throughout the world. The most widely known North American species are the great blue heron, *Ardea herodias*, the green heron, *Butorides virescens*, and the egret, *Egretta*. The last is a white bird which bears beautiful plumes known as aigrettes during the breeding season. It was once threatened with extinction through the use of these plumes as ornaments for hats, but the remaining birds are adequately protected. See illustration.



Egret. (*National Audubon Society; Grant M. Haist.*)

These birds are remarkable for the down which they produce. It is exceptionally light and fluffy and grows all over the breast, rump, and flanks. The birds roost in tall trees during the day and feed mostly at night. They have long legs and long beaks. Sometimes they will reach a height of about 20 inches (51 centimeters). Some species are gray with black on head and neck. However, there is a wide variation both in size and coloration. See also **Ciconiiformes.**

HERPES SIMPLEX VIRUS DISEASES. Four major herpesviruses cause infections in humans: (1) Herpes simplex; (2) varicella-zoster; (3) cytomegalovirus; and (4) Epstein-Barr virus. These are among the most widespread of all human pathogens and, characteristically, tend to follow cycles of dormancy and activity within an individual, such cycles often extending over long periods. A long span of dormancy may be interrupted by a flareup resulting from unusual physical or psychological stress. There is no known effective treatment for achieving their full eradication. Incidence of herpesvirus infections tend to occur more frequently in immunosuppressed patients. See **Immune System and Immunology.** These viruses are described further in the entry on **Virus.** See also **Cancer Research.**

There are two types of herpes simplex virus (HSV), with multiple strains of each type. Type 1 infects mucous membranes of the oral cavity, perioral skin, eyes, and skin above the waist. Type 2 usually causes a genital infection, an infection which is the second most common venereal disease found in the United States and a number of other countries. It has been estimated that between 30 and 90% of young adults carry antibody to one or both types of herpes simplex virus. Most infections from Type 1 HSV occur during childhood, but may occur any time during life. Type 2 usually does not appear before puberty and the commencement of sexual activity. The incidence of Type 2 antibody peaks by age 35 years. Type 2 antibody can be found in 20–35% of the general population. Certain occupational groups, such as health professionals and prostitutes, are at greater risk for HSV infection—simply because of the greater number of possible contacts with the virus. However, the infection is found in all segments of society.

Both types of herpes simplex virus appear to be spread by close contact between infected and susceptible individuals. Incubation period ranges from 2 to 20 days. Even in persons with antibody to the virus, second infections are observed. Virus may be shed from the oral or genital mucous membranes. Where there is adequate immunity, the shed virus may produce either subclinical infection or observable clinical disease. In persons who are highly immunosuppressed (as in cases of persons who have received organ transplants), the infection may be widespread and not heal for many months.

The most serious infection of Type 1 HSV is herpes keratitis, which can lead to destruction of the cornea. Other primary infections of Type 1 include stomatitis, pharyngitis, tracheobronchitis, and dermatitis. Type 2 usually involves vulvovaginitis or balanitis. Some clinical studies have shown that, in addition to causing oral, ocular, and genital lesions, HSV infections may involve visceral sites, such as the throat, lungs, esophagus, brain, meninges, liver, spleen, and pancreas. It has been observed that organ transplantation and the widespread use of cancer chemotherapy have increased the frequency of herpes simplex visceral infection.

The neonate is seriously threatened in cases of maternal genital infection. Such infections can be fatal in about half of the cases of the newborn. Studies have shown that even when the mother is asymptomatic, there is a risk to the infant. Some authorities have suggested that the incidence of neonatal Type 2 disease could be lowered by performing a cesarean section in symptomatic women.

Herpes simplex virus obtained from infected sites grows easily. The virus also can be demonstrated in tissues by fluorescent antibody staining. A fourfold rise in complement-fixing antibody titer supports the diagnosis, but is not itself diagnostic. Intranuclear inclusion bodies may be observed in tissues from patients who are infected with herpes simplex as well as in tissue from patients infected with varicella-zoster virus or cytomegalovirus.

Therapy depends on the site involved. For herpes keratitis, vidarabine, trifluridine, and idoxuridine are licensed. Topical acyclovir is effective for initial genital herpes and for localized mucocutaneous lesions in immunocompromised patients. Intravenous acyclovir has been approved for herpes simplex infection in immunocompromised patients and for initial genital herpes infection in immunocompetent patients that is sufficiently severe to require hospitalization.

Every effort should be made to deter transmission of the virus. Medical personnel and others close to infected patients, such as sexual partners, should avoid direct contact with lesions. Asymptomatic shedding, unfortunately, limits the effectiveness of effort to prevent spread.

HERPETOLOGY. The study of amphibians and reptiles, often mistakenly believed to be a study of reptiles only and snakes in particular. See also **Snakes.**

HERPOLHODE. The curve along which the cone traced out by the angular velocity vector intersects the invariable plane tangent to the momental ellipsoid and perpendicular to the angular momentum vector, in the case of a rotating rigid body not subject to any external torque. The concept is useful in studying the dynamics of a rigid body.

HERRING (*Osteichthyes*). Of the order *Clupeiformes*, herring are a characteristic fish group of the oceans, but also include a number of species that inhabit tropical fresh water. They form schools and are found near shores as well as in the open sea. Many of them are migratory. Herrings can be distinguished from other species by several characteristics—there are no rayed canals on the gill cover bones; lateral line pores are absent; there are keel scales along the medial line of the belly. Noteworthy skull characteristics include a suprabranchial organ with unknown function which joins the fourth and fifth gill arches; there is little dentition in the mouth, since most species feed on plankton; and there are no teeth on the parasphenoid (a bone at the base of the skull).

Commercially important herring are of the suborder Clupeoidei and most statistics report the catch of clupeoids without distinguishing specific subtypes. About 25 herring genera, with some 100 species, live in the sea. Until recent times, herring was considered the most important commercial fish catch.

Atlantic Herring. This fish (*Clupea harengus*), shown in Fig. 1, is one of the most important of commercial fishes in the northeastern Atlantic Ocean. Originally, it was presumed that the entire herring popu-

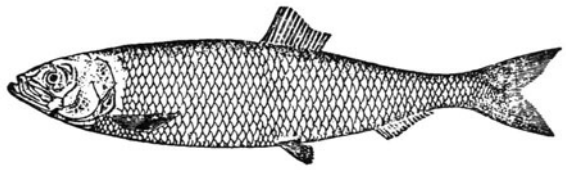


Fig. 1. Atlantic herring.

lation in the northeastern Atlantic Ocean was a unified group, inhabiting the region from the Arctic Ocean to the English Channel. From here, this group presumably migrated extensively to the north and south (and back) during the course of a year. But in recent years structural differences in herring caught in different regions have been noted. Based upon these and other studies, herring researchers met in Copenhagen in 1956 and proposed the following classification:

Class A Herring—Known as the *Atlanto-scandian herring*, which inhabit the open Atlantic Ocean and which spawn on the Atlantic coasts of northern Europe in mid-winter, spring and possibly in early summer. These fishes are of appreciable size and are characterized by an intermediate number of vertebrae (57 or more).

Class B Herring—Known as *shelf herring*, which inhabit the North Sea on the shelf west of the British Isles and in the transition zone between the North and Baltic Seas. These fishes spawn between August and January along the coasts. They reach a smaller size and have an intermediate number of vertebrae.

Class C Herring—Distributed inside coastal waters of the North Sea, in the traditional area between the North and Baltic Seas, and in the Baltic Sea. They spawn in shallow water during winter and spring. Body size and vertebral number are smaller than in the Class B herring.

Class D Herring—Found in the most northeastern part of the Atlantic Ocean and which, at one time, were classified with the Pacific herring.

The eggs of the Atlantic herring have a diameter of about 1 to 2 millimeters. In general, the winter-spring spawners have a relatively lower fertility and larger eggs, while the opposite is true of the summer-fall spawners. Winter-spring spawners lay from 22,000 to 40,000 eggs, whereas the summer-fall spawners lay from 48,000 up to 70,000 eggs. Freshly hatched larvae are found in tremendous masses on the spawning grounds and vicinity. They are transparent and very slender. Even when the fish is only about 0.8 inch (2 centimeters) in length, the distinguishing characteristics of the herring can be observed.

The availability of suitable plankton is the most important determinant for the development and ultimate survival of the larvae. The prey must be as close as 0.2 inch (0.5 centimeter) from the herring larvae in order for it to be perceived and eaten. The larvae feed only on moving organisms.

Summer and fall spawners of the North Sea reach sexual maturity in the third or fourth year, at which time they have a length of about 9.5 inches (24 centimeters). Life expectancy of the summer-fall spawners is from 12 to 16 years, while late-winter spawners of the Norwegian coast may live from 23 to 25 years.

Herrings have been found in schools ranging from hundreds to thousands of individuals, in all sizes from young to sexually mature adults. The individuals within a particular school are generally of equal size and age. It is not known how long they remain together. Studies have indicated that herring in the North Sea spend the day at the floor. With the beginning of dusk, they ascend to depths of some 100 to 165 feet (30 to 50 meters) into a warmer-temperature zone. Light plays an important role in this movement. During darkness, they seem to remain scattered at the higher level, but with the onset of dawn, they collect and return to the floor. In recent years, herring behavior at darkness has been studied closely. Soviet researchers have observed daily activities from a submarine and report that the Atlanto-scandian herring spends the night motionless at the surface of the water in an oblique position, as if sleeping. They become active shortly before dawn and begin their move to greater depths. Other researchers believe that schooling behavior ceases at night.

In recent years, the behavior of commercial fishes has been studied intensively with a view toward developing better fishing methods. One finding of these studies has been that vision plays a significant role in herring. Experiments have shown that herring do not avoid plastic sheets if they are transparent. Herring in which the eyes were covered could not perceive a net, while those with sight did detect the obstruction. During the day, a wall of air bubbles can act as an obstacle to herring, but they will swim right through it at night. Herring can also detect noises and vibrations, to which they respond with fright behavior. This has been confirmed with echolocation tracking. When a ship moves toward a herring school, the school sinks to a depth of many meters. Fright can also be induced in an aquarium by tapping on the wall.

During a few months of the year, herring can be found in certain regions in tremendous quantities, while at other times these same areas are completely devoid of the fishes. On the other hand, herring can be caught in some places throughout the year, but the catch varies from year to year. Marking studies have shown that Atlanto-scandian herring migrate between feeding grounds off Iceland and spawning grounds on the Norwegian coast. It has been established that this major herring population has 3 major growth and development areas, i.e., in the Norwegian fjords, the Barents Sea, and in the southern and eastern parts of the ocean off northern Europe. Young herring which have developed in the fjords migrate to the sea at an age of 2 to 3 years, where they meet those herring which have been developing there.

Herring feed on plankton, which is not simply filtered, but selected—a phenomenon found by stomach content investigations and aquarium studies. Some researchers indicate that herring select food visually and then again test it in the mouth. Materials that are useless or of bad taste are immediately rejected.

Three stocks of herring are taken in Icelandic waters. Two of them are of Icelandic origin and the third of Norwegian origin. The Newfoundland herring fishery is entirely coastal, particularly concentrated in the west and south coast regions. The fish are caught during winter and spring with gillnets and purse seines. The industry in Newfoundland has increased by 200% to 300% during the past 20 years.

Pacific Herring. This fish (*Clupea pallastii*) inhabits the coasts of the northern Pacific Ocean from the Bering Strait to Korea and in the Arctic Sea to the mouth of the Lena River. On the North American coast, its distribution extends from California to Nome, Alaska. Herring in the White Sea and from Cape Kanin to the Kara Sea are very similar to the Pacific herring. This species differs from the Atlantic herring by the smaller number of vertebrae, among other features. Generally, the eggs are laid in brackish water on plants. Pacific herring form spawning groups whose distribution is limited to very specific narrow zones. They apparently do not migrate to a great extent. On the Asiatic coast, 10 spawning groups are known in the region from Korea to the Sea of Okhotsk. Of these, the *Hokkaido-Sakhalin herring* is commercially the most important. The principal herring fisheries on the Pacific coast of the United States are in the bays and channels of southeastern and central Alaska. A general downward trend of the herring catch in Alaska, as contrasted with British Columbia, commenced in the early 1950s.

Sprat. This fish (genus *Sprattus*) is closely related to the herring. Six well-defined species have been identified. The majority are found in the southern hemisphere. Length ranges up to about 8 inches (20 centimeters). Coloration of the best known species (*Sprattus sprattus*) resembles herring and is irridescent. These fishes are found in the northern hemisphere on the European coast from Tromsö to the Baltic Sea and the Bay of Biscay, as well as in the Mediterranean and in the bordering waters of the Black Sea. Sprats do not undertake long migrations like herring. They generally stay near the coast and in river mouths. In the Baltic Sea, the sprat is found in water with low salt content. The sprat apparently avoids areas far from the coast. It reaches sexual maturity at an age of 2 to 3 years. Spawning takes place some distance from the coast and occurs in the North Sea from April to July; in the Kattegat and Skagerrak from May to June; in the Baltic Sea from May to August. Sprats, like herring, are commercially important fishes.

Sardines. Also related to sprats and herrings and of large commercial importance are the *Sardinops* sardines, of which there are at least 5 significant species: (1) *Pacific sardine* (*Sardinops caerulea*); (2) *South American sardine* (*S. sagax*); (3) *Japanese sardine* (*S. melanosticta*); (4) *Australian sardine* (*S. neo-pilchardus*); and (5) *South African sardine* (*S. ocellata*). Commercially, the latter species is of the least commercial importance.

The *Pacific sardine* is of large commercial importance for the United States and is found on the east coast of the Pacific Ocean from Baja California to British Columbia. The species lives in schools near the surface of the water and spawns between January and June, chiefly in March and April. Spawning takes place on the high seas off Baja California and southern California, as much as 300 nautical miles (556 kilometers) from shore. The larvae hatch 3 to 4 days after spawning and they migrate to the coast at a length of 3 to 5 inches (7.5 to 12.5 centimeters). They are caught in great masses and used as bait for tuna. At a length of about 6.5+ inches (17 centimeters), they leave the feeding grounds off the coast and meet the adults swimming on the open sea. Sexual maturity is attained at a length between 6.5 and 9.8 inches (17 and 25 centimeters), which occurs at an age of 2 to 3 years. The species can reach an age of 13 years. In California waters, the sardine catch has decreased dramatically since the late 1930s.

Sardine Ecology. The California sardine fishery has become the classic example of an ecologically complex community modified by an intensive fishery. A simple matter of overfishing might, on theoretical grounds, be overcome by abstention from fishing for an appropriate period, but apparently an ecologically related and evidently competitive species, the anchovy, has occupied the gap left by the exploitation of the sardines (the gap was evidently increased by harvesting during a period of conditions unfavorable for reproductive success). In the absence of a similar market for anchovies, a reduced technology for processing sardines, and legislative restrictions on harvesting anchovies, the situation reached the stage where a sardine fishery of nearly any magnitude further decreased the stock. Possibly this imbalance could be redressed by an unpredictable alteration in natural conditions in favor of the sardine, but this does not appear likely. It is interesting to note that the Pacific sardine supported the largest fishery in the Western Hemisphere in the early 1930s (exceeding over 1 billion pounds; 0.45 billion kilograms) taken from California waters, as compared with lower catches of just a few million pounds annually in recent years. From an economic standpoint, much of the loss of production of California and Oregon sardines has been compensated by large increases in menhaden catches in the south Atlantic Ocean and off the Gulf states.

The *South American sardine* and the *South African sardine* have increased in production since World War II. The South African sardine, sometimes referred to as the *pilchard* (*Sardinops ocellata*), has a wide geographical distribution and is known from St. Lucia Bay (north of Durban) to Bahia dos Tigres on the Angolan coast. The main commercial concentrations are limited to the Walvis Bay region, the waters off St. Helena Bay, and the area between Cape Point and Cape Agulhas. The species is normally found within 25 miles (40 kilometers) of the coastline, but occasionally schools have been reported up to 80 miles (129 kilometers) offshore.

The South African pilchard is a fast-growing fish and reaches sexual maturity at the age of about 2.5 years, by which time it attains a length of some 8.25 inches (21 centimeters). The main spawning seasons are spring and early summer. Spawning occurs offshore and three main grounds have been identified—those off Walvis Bay; near St. Helena Bay; and east of Cape Point. The pilchard is a filter-feeder, its diet consisting of both phytoplankton and zooplankton. Tagging experiments have established that there is periodically an influx of pilchards from the Walvis region into Cape waters.

The *Japanese sardine* is a warm-water species which attains a length of about 11.5 inches (29 centimeters). Distribution is chiefly in a temperature of from 59 to 79°F (15 to 26°C). Spawning grounds are off the south coast of Japan and Korea at some distance from the coast. Japanese sardines spawn from December to May on the high seas at a water temperature of 55 to 68°F (13 to 20°C). The number of eggs varies between 27,000 and 84,000. After spawning, a migration to the north takes place on the far eastern coast, where a large fishing industry has developed. These sardines feed chiefly on plankton. The annual catch (mainly by Japanese, Russian, and Korean fishers) varies considerably from year to year.

True Sardine. Only one species belongs to the genus of the true sardines, the *Sardina pilchardus*, also called *pilchard*, and not to be confused with the South African fish previously described. See Fig. 2. The true sardine reaches a length of about 11.8 inches (30 centimeters), but is generally from 9 to 9.8 inches (23 to 25 centimeters) long. Commonly, the larger sizes are called pilchards, while the smaller fishes are called sardines, the latter ranging between 5 and 6 inches (13 and 16 centimeters) in length. Distribution is on the coasts of west and southwest Europe and north Africa, from southern Ireland, the southern part of the North Sea, and the Kattegat in the north to Madeira and the Canary islands in the south. Distribution also includes the northern parts of the Mediterranean and bordering waters. There are two subspecies. The spawning period of the pilchard is rather extended. Off the Iberian peninsula, spawning takes place from February to March; in the North Sea, from July to August; off the coast of west Britanny, November to June; in the Mediterranean, September to May; and in the Black Sea, July and August.

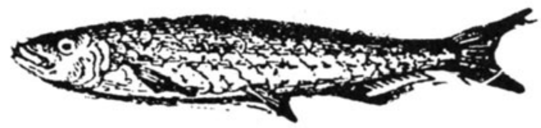


Fig. 2. True sardine.

The distribution of the true sardine is approximately limited by the 68°F (20°C) isotherm. Until 1930, the Strait of Dover was apparently the northern limit for pilchards. Those found in Norwegian waters and in the Kattegat only occurred in small numbers. Since that time, the northern population has increased significantly. In the late 1930s, large quantities of sardine eggs were reported off the East Frisian islands. In late 1940s, the first large quantities were found in the area near Amrum island (in the North Frisians). Climatic changes are probably responsible for the extension of the northern limit.

Adult pilchards feed primarily on zooplankton. The catch of pilchards has been increasing progressively over a number of years and is very important to Portugal, Morocco, Spain, France, and the former Yugoslav Republics, most of the catch being used in production of canned sardine oil.

Shad. This fish of the genus *Alosa* is also related to the herring. Shads have a compressed upper body and the keel scales form a sharp keel. The teeth in the jaw are either small or absent altogether, and in adults there are no vomerine teeth. There are four species in the north Atlantic Ocean, Mediterranean, and in the northern Pacific Ocean. These species migrate into fresh water. The best known species in Europe is the shad (*Alosa alosa*). The length of the shad exceeds 27.5 inches (70 centimeters) and the jaw protrudes forward. The scales are not as lightly attached as in the herring. Distribution is on the European coast from Norway to the Iberian peninsula, and along the north African coast to Morocco, as well as the western part of the Baltic Sea and the Mediterranean. In March, shad migrate from the sea to spawn in the rivers well upstream. Earlier, they were found in the Neckar River, Germany. During recent years, shad have disappeared from much of their original habitat, largely as the result of pollution.

The *American shad* (*A. sapidissima*) is found mainly in the Atlantic Ocean, from the Gulf of St. Lawrence to Florida. In 1871, the species was introduced to the Sacramento and Columbia Rivers in the western United States and, by 1876, shad were caught off Vancouver island. Since that time, the species has spread along the entire coast of the Pacific Ocean from southern California to Alaska and Kamchatka. It is a prevalent fish in the California rivers. Since the earliest settlements in North America, the American shad has been an important and valuable commercial fish, while the Alabama species has enjoyed much regional acclaim. Over the years, however, the shad population has decreased.

See also **Fishes.**

HERRINGBONE BEAR. See **Helical Gearing.**

HESSIAN. A functional determinant, related to the Jacobian and defined for six variables by the equation

$$H(F) = \frac{\partial(u, v, w)}{\partial(x, y, z)} = \begin{vmatrix} F_{xx} & F_{xy} & F_{xz} \\ F_{xy} & F_{yy} & F_{yz} \\ F_{xz} & F_{yz} & F_{zz} \end{vmatrix}$$

where u, v, w are differential coefficients of another function, $F(x, y, z)$ and $u = \partial F/\partial x = F_x$, $v = F_y$, $w = F_z$; $\partial^2 F/\partial x^2 = F_{xx}$, etc.

It can be generalized for any number of variables. The Hessian of two binary quantics is a covariant; hence, it is useful in studying the invariants of algebraic functions. As an example, the Hessian of a quadratic is its discriminant.

See also **Determinant.**

HESSIAN FLY (*Insecta, Diptera*). One of the worst pests of wheat. It is a small two-winged fly *Mayetiola* (*Phytophaga*) *destructor*, a member of the gall-gnat family, which was introduced into the United States in the Revolutionary period. The larva lives between the base of a leaf and the stem of the wheat plant and either kills or weakens the plant so that no grain develops. Other cereals are attacked to some extent.

Fall plowing and burning stubble aid in destroying many insects. The most effective means of avoiding damage to winter wheat is to sow late enough to avoid the attack of most of the adults. They live no more than ten days and the date of emergence is known for various regions; hence, late planting subjects the crop only to the light infestation due to the eggs deposited by the relatively few flies which emerge late. Phorate is an effective chemical control.

HESSITE. A mineral telluride of silver, Ag_2Te, with some gold, crystallizing in the monoclinic system at normal temperatures; isometric system above 149.5°F (65.3°C). Crystalline form not obvious at normal temperatures. Hardness, 2–3; specific gravity, 8.24–8.45; color, gray with metallic luster; opaque. Named after G. H. Hess (1802–1850).

HETERODYNE. This term is used in communications terminology as an adjective or a verb, but in either case it concerns the beating together in an electrical circuit of two frequencies to produce new frequencies which are the sum or difference of the original ones. When two voltages of different frequencies are applied simultaneously to a circuit containing a non-linear impedance, for example, one in which the signal current varies as the square or higher power of the input signal voltage, the output of the circuit will contain new frequencies, among them one equal to the sum and another equal to the difference of the applied frequencies. Either one or both of these may be selected by properly tuning or filtering the output.

HETEROGAMY. The occurrence or union of male and female gametes of different size and structure; anisogamy. The alternation of two sexual generations, one true sexual, the other parthenogenetic.

HETEROMORPHOSIS. Deviation from normal form. Malformation or deformity and also less extreme departures incidental to slightly different conditions in the animal or its environment.

HETEROPOLYACIDS. Acids derived from two or more other acids, under such conditions that the negative radicals of the individual acids retain their structural identity within the complex radical or molecule formed. The term heteropolyacids is usually restricted to complex acids in which both radicals are derived from oxides, such as phosphomolybdic acid.

HETEROSPORY. The production of two distinct types of spores by a plant, in contrast to homospory, which is the production of only one type of spore. The two kinds of spores produced in heterospory are known as microspores and megaspores. The microspores are very small and grow into the male gametophyte. The megapores are much larger and form the female gametophyte. All of the seed plants have heterospory and a few of the minor subphyla of vascular plants do also. The *Lycopsida* is one of these subphyla. See also **Lycopsida.**

HETEROZYGOUS. Bearing two allelic genes of a different nature. The opposite of homozygous, which means to bear allelic genes of the same kind. For instance, if a person is homozygous for the recessive gene for albinism, the person will bear two such genes, represented as *aa*. The person will be an albino. A person who is heterozygous, however, will bear one gene for normal pigmentation and one gene for albinism, represented as *Aa*. A person can also be homozygous for the dominant gene, *AA*. Both heterozygous and homozygous dominant persons will have normal pigmentation.

HEULANDITE. The mineral heulandite is a monoclinic zeolite whose crystals are often quite suggestive of orthorhombic forms. Its chemical composition is probably $(Na, Ca)_{4-6}Al_6(Al, Si)_4Si_{26}O_{72}\cdot 24H_2O$; strontium may be present. Heulandite has one good cleavage; is brittle with a conchoidal fracture; hardness, 3.4–4; specific gravity, 2.18–2.22; luster, vitreous to pearly; color, white to gray, red or brown; streak, white; transparent to translucent. Occurs chiefly in cavities in basaltic rocks with other zeolites, but may be found in granites, pegmatites, gneisses, and schists. Famous localities are in Iceland, India, the Harz Mountains, Italy, Switzerland, Scotland, Nova Scotia; and in the United States at Bergen Hill and West Paterson, New Jersey. This mineral was named for the English mineralogist Heuland.

HEXACTINELLIDA. The glass sponges, constituting a class of the phylum *Porifera*. The spicules of the skeleton are silicious and of six-rayed form. Many of the species have a large central cavity, resulting in a tubular or vase-like form, and when freed of organic matter appear to be made of spun glass. These sponges are found in deep water in the ocean. Venus' flower basket, *Euplectella*, and the glass-rope sponge, *Hyalonema*, are the most common examples.

HEXADECIMAL NUMBER. In computer design, the hexadecimal (radix 16) numbering system is used as a convenient method for representing large binary numbers, which often consist of long strings of zeros and ones. The latter are difficult to handle in a computer. Each hexadecimal digit stands for four binary digits.

Hexadecimal notation calls for the use of 16 symbols to represent 16 number values. Inasmuch as the decimal system provides only 10 number symbols (0 to 9), six additional marks thus are needed to represent the remaining values. The letters A, B, C, D, E, and F are used for this purpose. As shown in the accompanying table, the list of hexadecimal

COMPARISON OF DECIMAL,
HEXADECIMAL, AND BINARY NOTATION

Decimal	Hexadecimal	Binary
0	0	0000
1	1	0001
2	2	0010
3	3	0011
4	4	0100
5	5	0101
6	6	0110
7	7	0111
8	8	1000
9	9	1001
10	A	1010
11	B	1011
12	C	1100
13	D	1101
14	E	1110
15	F	1111
16	10	10000
17	11	10001
18	12	10010
19	13	10011
20	14	10100
21	15	10101
22	16	10110
23	17	10111
24	18	11000
25	19	11001
26	1A	11010
27	1B	11011
28	1C	11100
29	1D	11101
30	1E	11110
31	1F	11111

symbols is comprised of 0, 1, 2, 3, 4, 5, 6, 7, 8, 9, A, B, C, D, E, and F, in ascending sequence. From the table, note that upon reaching decimal 16, the hexadecimal symbols are used up and hence a "1 carry" must be placed in front of each hexadecimal symbol during its second cycle, i.e., from decimal 16 to decimal 31.

Binary numbers are converted to hexadecimal notation simply by dividing the numbers into groups of four binary digits, commencing from the right, and replacing each group by the corresponding hexadecimal symbol. Where the left-hand group is incomplete, zeros are filled in as required. This is illustrated by the following example.

$$\begin{aligned} 111110011011010011 &= 0011/1110/0110/1101/0011 \\ &= 3\ E\ 6\ D\ 3 \\ &= (3E6D3)_{16} \end{aligned}$$

Hexadecimal numbers are best understood in terms of expansion in powers of 16. In the case of hexadecimal number 2CA.B6, for example, when decimals are substituted for hexadecimal symbols, it is evaluated as

$$\begin{aligned} & 2 \times 16^2 + 12 \times 16^1 + 10 \times 16^0 + 11 \times 16^{-1} + 6 \times 16^{-2} \\ &= 2 \times 256 + 12 \times 16 + 10 \times 1 + 11/16 + 6/256 \\ &= 512 + 192 + 10 + 0.6875 + 0.0234375 \\ &= 714 + 0.7109375 \\ &= (714.7109375)_{10} \end{aligned}$$

HEXAMINE. $(CH_2)_6N_4$, formula weight 140.19, white crystalline solid, mp 280°C, decomposes at higher temperatures. Also known as hexamethylenetetramine, methenamine, and urotropine, the compound is soluble in H_2O and only very slightly soluble in alcohol or ether. Although used to some extent in medicine as an internal antiseptic, the primary use of hexamine is in the manufacture of synthetic resins where the compound is a substitute for formalin (aqueous solution of paraformaldehyde) and its NaOH catalyst. Hexamine also is used as an accelerator for rubber.

On a commercial scale, hexamine is manufactured from anhydrous NH_3 and a 45% solution of methanol-free formaldehyde. These raw materials, plus recycle mother liquor, are charged continuously at carefully controlled rates to a high-velocity reactor. The reaction is exothermic. The reactor effluent is discharged into a vacuum evaporator which also serves as a crystallizer. The hexamine crystals then are washed, dried, and screened. Average yield of the process is about 96% conversion of ingredients to produce hexamine.

HEXAPODA. Synonymous with Insecta.

HIATUS HERNIA. See **Esophagus.**

HICKORY AND WINGNUT TREES. Of the family *Juglandaceae* (walnut family), hickory trees are of the genus *Carya* and are one of the most, if not the most distinctly North American tree. They are relatively unknown on the other continents. One authority aptly describes the hickories as walnuts with greater height and grace. The principal species are indicated in the accompanying table. It should be appreciated that the tree dimensions given in the table represent record specimens and that the average tree, most likely growing under somewhat more adverse conditions, will not attain such dimensions.

The bitternut or swamp hickory (*C. cordiformis*) has a light-brown or gray-brown thin bark. The fissures are shallow. The leaves are compound. The leaflets are of a deep yellow-green color, somewhat lighter

RECORD HICKORY TREES IN THE UNITED STATES[1]

	Circumference[2]		Height		Spread		
Specimen	Inches	Centimeters	Feet	Meters	Feet	Meters	Location
Bitternut (1975)[3] *Carya cordiformis*	174	442	120	36.5	80	24.4	Virgnia
Bitternut (1982)[3] *Carya cordiformis*	149	378	137	41.8	115	35.1	Michigan
Bitternut (1991)[3] *Carya cordiformis*	177	450	115	35.1	108	32.9	Tennessee
Black (1980) *Carya texana*	103	262	135	41.1	66	20.1	Texas
Carolina (1988) *Carya ovata*	100	254	114	34.7	51	15.5	North Carolina
Mockernut (1989) *Carya tomentosa*	140	356	156	47.5	70	21.3	Mississippi
Nutmeg (1985) *Carya myristiciformis*	132	335	145	44.2	80	24.4	Alabama
Pignut (1985) *Carya glabraavar glabra*	157	399	190	57.9	78	23.8	North Carolina
Red (1982) *Carya glabra var odorata*	142	356	140	42.7	62	18.9	Tennessee
Sand (1982)[3] *Carya pallida*	114	290	114	34.7	86	26.2	North Carolina
Sand (1980)[3] *Carya pallida*	138	351	94	28.7	86	26.2	New Jersey
Shagbark (1984) *Carya ovata*	132	335	153	46.6	56	17.1	South Carolina
Shellbark (1986) *Carya laciniosa*	174	442	105	32.0	123	37.5	Virginia
Water (1991) *Carya aquatica*	162	411	135	41.1	88	26.8	Virginia
Pecan (1983) *Carya illinoensis*	286	726	130	39.6	90	27.4	Mississippi

[1]From the "National Register of Big Trees," The American Forestry Association (by permission).
[2]At 4.5 feet (1.4 meters).
[3]Cochampions.

underneath. The ovoid fruit is about an inch long and is contained in a thin husk. The tree prefers a rich woodsy environment, but will tolerate a variety of soils. The tree ranges from southern Maine and western Quebec westward to the Great Lakes and Minnesota and south through Nebraska, Kansas, Oklahoma, and Texas. It ranges eastward and south to Florida. The tree is found commonly only in southern New England, with only occasional representation in Vermont and New Hampshire. The tree attains its greatest height in the mountains of the Carolinas. As compared with other hickories, the wood is considered inferior.

The nutmeg hickory (*C. myristicaeformis*) prefers alluvial soil. It is found in the southeastern states and westward through Arkansas. The fruit is a little over an inch long and contained within a thin husk. The shell is very hard; the kernel is not edible. The wood is strong, hard, and is of a light-brown color.

The mockernut or bigbud hickory (*C. tomentosa*) occurs in southeastern Canada and the eastern United States. It tends to be a very tall tree with a round head. The dark-green leaves are long, with from five to nine toothed and pointed leaflets. The male catkins are from 3 to 5 inches (7.6 to 12.7 centimeters) long.

The water hickory or bitter pecan (*C. aquatica*) is generally a tree of the coastal plain, ranging from Virginia southward to Florida and then westward into Texas. The tree is capable of attaining great heights and is generally quite slender. The bark is an ashen gray, thin, and often quite shaggy on older trees. The leaves are compound. Leaflets have sharp points, of a deep yellow-green color, with slightly lighter coloration underneath. The wood is considered inferior as compared with other hickory species.

The shagbark or shellbark hickory (*C. ovata*) is a tree of stature and beauty and of great utility. It is capable of attaining great height, as evident from the accompanying table. It is valued for its wood and nuts. See Fig. 1. The bark is a pale-brown/gray and very shredded and shaggy—hence the name. Often, the bark will hang loosely in strips of a foot or more in length. The branches are pendulous and the foliage is a deep green. The leaves are large, from 4 to 6 inches (10.1 to 15.2 centimeters) in length. The staminate catkins occur in clusters of three and are green. The fruit may be described as globular in shape, with a very thick husk. The nut is white, thin-shelled and the kernel is sweet. This is the most important of the hickory nuts marketed, not of course including the pecans to be described shortly.

The shagbark hickory ranges from the Saint Lawrence River valley southward into Maine and generally following the Appalachian Mountains into the southeastern states. The tree ranges westward through Michigan and Minnesota and southward into Kansas, Oklahoma, and Texas. The tree does very well in certain parts of New England and particularly well in the Piedmont region of North Carolina. The wood is well known for its hardness, density, toughness, and close-grain. It is of a pale-brown color and remains the preferred wood for quality tool and implement handles and other heavy-duty applications. In the green state, the wood has a moisture content of 57% and a weight of 63 pounds per cubic foot (1009 kilograms per cubic meter). After air-drying to 12% moisture content, the weight per cubic foot is 51 pounds (817 kilograms per cubic meter) and 1,000 board-feet (2.36 cubic meter) of nominal sizes weigh 4,250 pounds (1927 kilograms). Crushing strength of the green wood when compression is applied parallel to the grain is 4,570 pounds per square inch (31.5 MPa); of the dry wood, 8,970 psi (61.9 MPa). The wood has 30% greater strength than white oak and double the shock resistance.

The species *C. illinoensis* is well known for its production of pecans. See Fig. 2. The tree ranges through much of the eastern United States. In particular, these trees are extensively cultivated in the southern states for their nut crop. The trees have huge branches and are capable of attaining a height of 100 feet (30 meters) or more. The head is rounded. There are 11 to 17 toothed, pointed leaflets. The tree was introduced to Europe many years ago and does well in the central and southern parts of France. For top production of pecans, the tree requires hot summers.

Wingnut trees are of the genus *Pterocarya*, deciduous, with large alternate, pinnate leaves. The leaflets are toothed. There are unisexual flowers in separate catkins appearing on the same tree. The trees bear small, winged nuts which occur on long hanging spikes. The trees are fast-growing and not too particular about soil. The Caucasian wingnut tree (*Pterocarya*) occurs in the Caucasus. It can attain a height of 100 feet (30 meters), is broad and spreading, with large oblong leaves from 8 to 14 inches (20.3 to 35.5 centimeters) in length. The *P. stenoptera* occurs in China and also can attain a height of 100 feet (30 meters). The

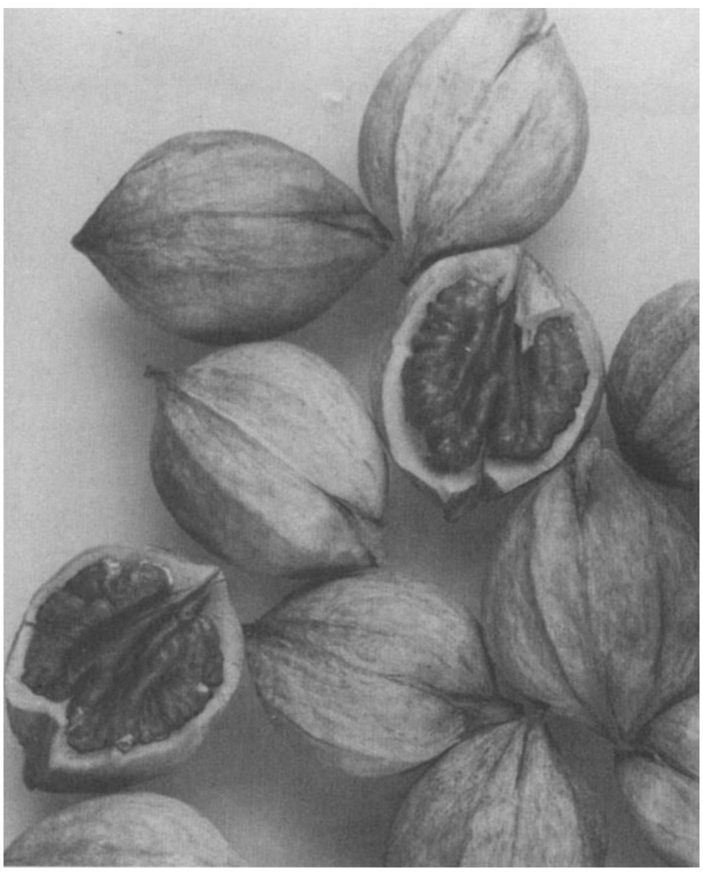


Fig. 1. Shellbark hickory nut. (*USDA photo.*)

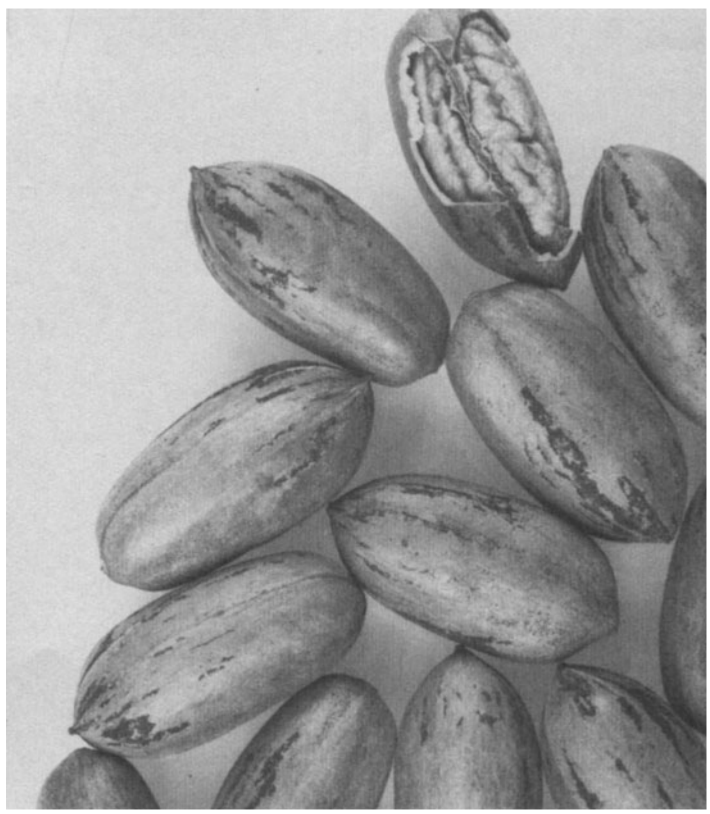


Fig. 2. Pecans. (*USDA photo.*)

P. x rehdrena is a hybrid of the two aforementioned species and is a shorter (up to 40 feet (12 meters) in height) broad-domed tree.

HIGH. See **Atmosphere (Earth).**

HIGH FIDELITY. The quality of a sound reproducing system such that the acoustical characteristics of the reproduced sounds (usually musical) match as closely as possible the characteristics of the original sounds when made under their normal conditions. Thus a high fidelity reproduction of a symphonic work should sound the same to the listener as if he were present in a concert auditorium, listening to the orchestra directly, even though the sounds used in the recording were actually transcribed in a recording studio with extremely artificial acoustical characteristics.

HIGH-G ACCELEROMETER. See **Acceleration Measurement.**

HIGH-LIFT DEVICES. See **Aerodynamics.**

HIGH LIMITING CONTROL. See **Control Action.**

HIGH-PASS FILTER. See **Filter (Communications System).**

HIGH-PRESSURE TECHNOLOGY. See **Pressure.**

HIGH TECHNOLOGY. A buzz term of the 1980s and early 1990s used mainly by the lay media to identify relatively new, complex, and sophisticated scientific and industrial pursuits, such as late-generation computers, electronic components, gene science, medical research, lasers and advanced optics, automation, communication systems, etc. It is frequently abbreviated as *high tech*. Use of the term was intended to connote a marked distinction between these more recent activities that are aligned with the information needs of society and the long-established, but more prosaic and heavy industries that are geared mainly to the material needs of society. In the case of high tech, the field is predominantly white collar, intellectual, clean, and nonpolluting, versus the older industries with their large blue-collar workforce. Many of the glowing promises of high tech have been much slower in reaching fruition than initially forecast. In retrospect, it is interesting to note that each generation of the past could have made its claims to high tech, as witness the light bulb, the telephone, radio and television, the automobile, the airplane, etc., which in its day probably were more revolutionary than most of the high-tech claims of the current generation. The high tech of today is based upon the high tech of yesterday.

HIGH-TEMPERATURE RESEARCH. See **Solar Energy.**

HIGHWAY BANKING. See **Superelevation.**

HILDEBRAND RULE. The entropy of vaporization, i.e., the ratio of the heat of vaporization to the temperature at which it occurs, is a constant for many substances if it is determined at the same molal concentration of vapor for each substance.

HILL'S DETERMINANT. See **Mathieu Equation.**

HIP. The joint at the attachment of the human thigh to the body. Also, the adjacent portion of the thigh where it merges with the buttocks and less commonly the corresponding part of the leg in various animals.

Congenital dislocation of the hip, caused by improper development during the fetal life, is thought to be a heritable condition. Females are much more likely to be afflicted than males, and the dislocation may be of one or both hips. The condition is often difficult to diagnose before the child begins to walk, although it is during infancy that treatment is most useful. The first symptom may be a more pronounced rotation of the femur than in normal infants. When only one hip is affected, the creases in the infant's thighs may not be symmetrical. Upon starting to walk, the individual may develop a limp and marked lordosis (forward curvature of the spine). Later, there is a shortening of the thigh and a wide space between the thighs when the child stands with feet together. Usually there is no pain associated with the dislocation until adulthood and that is usually a low-back pain resulting from the lordosis. Treatment involves a long tedious procedure employing casts and weights to gradually correct the dislocation. Surgical treatment may be required if soft tissues have developed in the space to which the head of the femur is to be restored.

Accidental dislocation of the hip, as with other bones, causes pain and limitation of movement. Nerves also may be severely injured. In hip dislocations, the major motor nerve may be paralyzed; and if nutrient-supplying blood vessels are torn, the head of the thighbone may become necrotic, soft, and die, or osteoarthritis may develop.

A total hip replacement is one of the most successful orthopedic procedures. Approximately 150,000 hip replacements are performed annually in the United States. A major goal of such procedures is the relief of pain. What is referred to as a total hip replacement may have a limited life, depending largely upon the kind of exercise and work that are customary for a given patient. Breakage and wear of the artificial joint theoretically are potential problems, but the most significant problem is that of the prosthesis loosening.

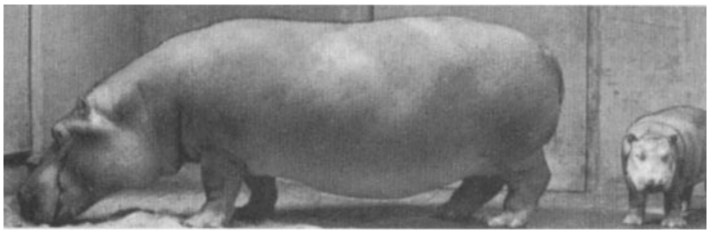
An artificial replacement hip joint.

Forces generated in a total hip replacement during normal walking are about 1.5 to 3 times body weight. With activities, such as running and jumping, these forces can reach 4 to 6 times body weight. An artificial hip replacement joint is shown in the accompanying figure. Ultimately, most patients can participate in numerous activities, including walking, bicycling, golf, hiking, swimming, rowing, and cross-country skiing. High-impact activities, however, should be avoided, and these include running, jumping, heavy lifting, and contact sports.

HIPPOCAMPUS. See **Nervous System and the Brain.**

HIPPOPOTAMUS (*Mammalia*, *Artiodactyla*). The group of *Hippotamines* is one of the smaller in the order *Artiodactyla* (even-toed hoofed animals). There are two extant species: (1) the greater Hippopotamus (*Hippopotamus*); and (2) the Pigmy Hippopotamus (*Choeropsis*).

The greater hippopotamus is a large, rather commonly occurring animal of the rivers of tropical Africa. The body is large, barrel-shaped, bulky, with short strong legs, and a very broad muzzle. The beast may attain a length of about 14 feet (4.2 meters) and a height of about 4 feet (1.2 meters). Four tons is usually the figure quoted for the larger specimens. They are almost hairless, and of a gray-black coloration with white underneath. Hippopotamuses are largely aquatic in habits and live entirely on vegetation, including both water plants and terrestrial species. They have been known to do extensive damage to crops in their roving. The animals feed mostly at night. The greater hippopotamus is the largest of the living nonruminating even-toed mammals. They are fast swimmers and can run about as fast as humans on land. They can issue loud grunts and bellows. Multiple births are rather uncommon. The gestation period is about eight months. The baby animal is known as a calf. See accompanying photo.



Female hippopotamus and baby. (*A. M. Winchester.*)

In some regions, the hippopotamus has been a staple food among tribesmen who harpoon the animal from small canoes. The manner in which the animal is prepared for consumption by some tribesmen is rather offensive to most people. After the body of the animal is dragged to the shore of the river, it is allowed to "ripen" under the hot tropical sun for a few days. The animal is then ripped open and the natives tear apart the softened carcass, gorging themselves on the rotten flesh.

Normally, the hippopotamus has a reasonably good disposition, but is proprietary concerning staked out stretches of the river. Usually, the animal will move out of the way of boats, floating just beneath the surface, but watching the passerby with use of the periscopic eyes which are just above the water line. However, the animal has been known to attack boats for unknown reasons, stomping and chewing the boat and occupants. Considering the size of the creature's mouth, the bite of a hippopotamus is no less than ghastly and usually terminal. The animal also is unpredictable when encountered on land, particularly at night.

The pigmy hippopotamus attains the size of a large pig. It has a large mouth equipped with fang-like teeth. For habitat, the pigmy hippopotamus prefers small lakes and rivers, ponds and stagnant pools and seldom wanders far from its aquatic habitat. The animal cannot afford to stay out of water very long because the skin is equipped with very large pores and, unless kept moist, the skin cracks easily. Apparently, the animals use these pores for absorbing water into their system, rather than taking all of their liquid input by mouth.

Naturalists always have suspected a relationship between the *Hippopotamines* and the *Suines*, but fossil remains have failed to yield evidence for this connection. And thus they remain in a separate classification. It is interesting to note that the Romans regarded them as pigs of a very special nature.

HIRSUTISM. Abnormal growth of hair, particularly on the face of women. Although not well understood, causative factors appear to include a predisposition to the condition by inheritance, variations in endocrine activity, and imbalance of the metabolic processes. The condition usually does not appear until middle age. Treatment essentially is of a cosmetic nature.

Hirsutism is one of the principal features of polycystic ovary syndrome. Control of this hirsutism is extremely difficult. Best results have been obtained with therapy directed to suppressing adrenal and ovarian functions. Hirsutism is also seen in Cushing's syndrome (hyperfunction of adrenal cortex). See also **Androgens.**

HIRUDINEA. The leeches, a class of segmented worms (phylum *Annelida*), well known for their habit of sucking blood. Marine and freshwater species are known, and in the moist tropical forests terrestrial species occur. They often attach themselves to bathers.

The members of this class are distinguished from other annelids by the following characteristics: (1) The body is relatively short, usually with 32 segments. (2) The external segments are annuli, numbering from 2 to 14 to each metamere. (3) Each end of the body bears a sucker. (4) The mouth is usually provided with three toothed plates or jaws. (5) The alimentary tract is provided with an enormous pouched crop in which blood is stored prior to digestion. (6) The anus opens dorsally to the posterior sucker. (7) The coelom is partially obliterated by a peculiar mesenchymal tissue. (8) At the anterior end of the ventral nerve cord, several ganglia (ganglion) are fused to form a large mass.

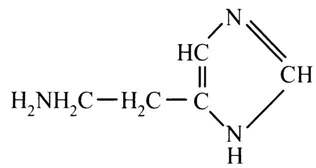
Leeches were once extensively used in medicine for letting blood and are still of minor importance for this purpose. Otherwise they are of no importance to man save as an occasional annoyance. They eat small aquatic animals as well as the blood of vertebrates, and some species are entirely predacious.

Two orders are recognized:

Order *Rhynchobdellida*. With a protrusible proboscis, colorless blood, and no jaws. Marine and freshwater.

Order *Gnathobdellida*. With jaws and red blood. No proboscis. Freshwater and terrestrial. The medicinal leech belongs to this order. It is native to Europe but is naturalized in ponds and streams of the eastern United States.

HISTAMINE. A powerful vasodilator which is released in anaphylactic (hypersensitivity to protein) shock and occurs in blood and tissues in minute amounts.

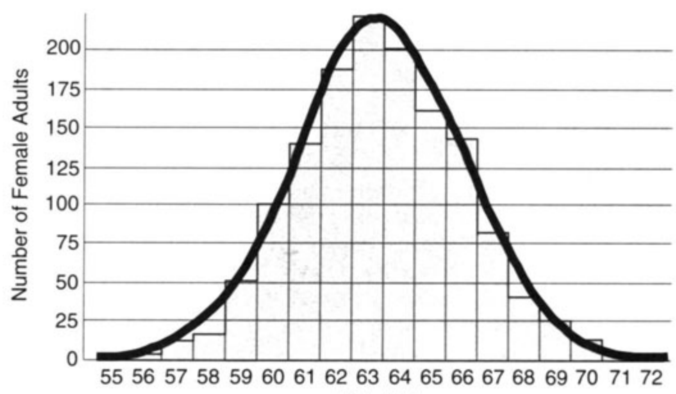


The injection of 1 microgram intravenously in humans is said to bring about a sharp drop in blood pressure. Its close relationship to histidine is emphasized by the fact that the amino acid can be decarboxylated by certain intestinal bacteria to produce it.

Histamine is a product of the degradation of histidine and is liberated by injury to the tissue, or whenever a protein is decomposed by putrefactive bacteria. Histamine's biological role is both positive and negative. The production of excessive histamine gave rise to the formulation of drugs for countering such excesses. See also **Antihistamine.** Histamine can cause pulmonary edema of noncardiac etiology. Histamine causes both constriction of bronchial smooth muscle and edema of bronchial mucosa by increasing the permeability of small bronchial veins. Histamine is one of several humoral mediators that affect bronchial tissue. Most cells contain and release histamine. Histamine stimulates connective tissue regeneration by producing edema.

HISTIDINE. See **Amino Acids.**

HISTOGRAM. A histogram is a graphical representation of a grouped frequency distribution. Rectangles are formed by using the class interval as the base and the frequency of the class as the height. Equal areas represent equal frequencies. See accompanying figure.



Histogram showing heights of female adults prepared from a survey of approximately 1400 women. Normal distribution curve is fitted to the histogram. This familiar bell-shaped curve typifies numerous empirical distributions found in biology.

HISTOLOGY. The science that deals with the minute structure of living things. Microscopic morphology. The study of the structure and functions of cells is the special province of cytology, leaving the study of special forms of cells and their association in tissues and organs as the field of histology, but histology necessarily includes much cytological matter.

The science is made up of two subordinate fields, general and special histology. In the former are considered the specialization of cells in the multicellular body and the characteristics and classification of the tissues in which they are grouped. The details of minute structure of the organs and organ systems are the materials of the latter. This field of histology is necessarily extensive and detailed, even in the study of a single species.

Histology recognizes five principal kinds of animal tissues, epithelium, nervous tissue, mesenchymal (connective and supporting) tissues, muscular tissue, and vascular tissue. All organs are made up of these components.

The tissues of plants are not so easily separable, but plant histologists recognize epithelial tissue, vascular tissue, supporting tissue, and parenchymous tissue. Plant tissue is studied more by its location than by the particular kinds of tissue.

HISTONES. Basic proteins which occur in the nuclei of both plant and animal cells. They are less basic than the protamines, having isolectric points at about pH 11. Some investigators restrict the term *histone* to only those basic proteins anatomically and chemically associated with DNA (deoxyribonucleic acids). The close associations of histones with DNA led to the hypothesis that histones might play a role in the control of genetic expression at the cellular level. Advances in molecular biology have permitted more detailed mechanisms for such control to be proposed. Histones, by blocking some areas of the DNA molecule, may permit only part of the DNA base sequences to act as templates for the formation of messenger RNA. Thus histones, by controlling messenger RNA formation, may ultimately control protein biosynthesis within the cell. Or, the primary role of histone may be structural, histone being essential for stabilizing the DNA helix, for the integration of DNA strands into more complex chromosomal structures, and for fixing and maintaining during cell division chromosomal changes occurring during differentiation and development. The foregoing two concepts are not mutually exclusive, i.e., histone may fix chromosomal structure in a specific configuration in which the position of the histone molecules also limit RNA formation. The possibility that histones play a role in genetic mechanisms suggests the possibility that histone changes may initiate or accompany early cellular changes, leading to the formation of tumors. Further investigation is needed to ascertain whether or not tumor histones differ from those of corresponding normal tissues. See also **Cell (Biology).**

HISTOPLASMOSIS. An intracellular mycosis of the human reticuloendothelial system resulting from inhalation of the dimorphic fungus *Histoplasma capsulatum* which has large (8 to 20 micrometers diameter) spores with thick capsules. The fungus is found worldwide in soils and predominantly in those which have been enriched by chicken droppings; minor epidemics have followed the cleaning out of chicken coops or aviaries. In the United States, cases are mainly reported from rural areas of the Mississippi and Ohio River valleys. In humans, two forms appear: (1) the disseminated form with ulcerative lesions in skin and mucous membranes, and (2) the pulmonary form which simulates tuberculosis. The condition often is difficult to differentiate from tuberculosis. Infants are highly susceptible and the disease can be fatal in the young. In the limited form, the lesions appear most commonly around the mouth, on buccal mucous membranes, and on the penis.

All ages are vulnerable to attack, with males reporting disease more frequently than females. One asymptomatic or benign infection bestows a lasting cellular and humoral immunity. Drugs which have been effective in some cases include 2-hydroxystilbamidine and amphotericin B, the latter administered intravenously. In the late 1970s, increasing interest was shown in the use of immunotherapy in selected patients with progressive disease who are receiving conventional therapy and have a demonstrable immune defect. Initial experience with transfer factor therapy has shown some promise and research continues.

A. C. V.

HISTOSOLS. See **Soil.**

HITTORF PRINCIPLE. An application of the Paschen law. The Hittorf principle states that discharge between electrodes in gas at a given pressure will not always occur between the closest points of the electrodes if the distance between these points corresponds to a point to the left of the minimum of the ignition potential curve.

HIVES. See **Urticaria.**

H LINES. A contour along which the electromagnetic field strength is constant with respect to some reference plane.

HOARFROST. See **Precipitation and Hydrometeors.**

HOATZINS (*Aves, Galliformes*). A very strange bird, which at first sight looks like a small curassow, belonging to the suborder *Opisthocomi*. There is only one species, the Hoatzin (*Opisthocomus hoazin*). The size is approximately that of a crow, the length is 60 centimeters ($23\frac{1}{2}$ inches) and the weight is about 800 grams (28 ounces).

Hoatzin. (*Sketch by Glenn D. Considine.*)

The hoatzins lay claim to a special position among all birds: first, as a result of their specialized diet, and second, because of the ability of their young, while they are still very undeveloped, to climb about a network of branches on all fours with the aid of the primary wing feathers which have talons. Young hoatzins have particularly long and movable first and second digits; each has a strong claw which retrogresses later. The oldest bird so far known, *Archaeopteryx* from the Jurassic period, also had such flexible fingers with claws. We assume that it, too, used them to climb around in trees. Therefore, young hoatzins look primitive when they move around in the branches like reptiles. They not only climb, but also swim and dive with all fours when danger impels them to drop into the water.

Old hoatzins, in contrast, avoid the water and almost never touch the ground. Yet they, too, give the impression of being "primitive" when they flit about in tree branches or awkwardly fly short distances. But that has to do with their diet, in which their crop plays a peculiar part.

Hoatzins primarily eat leaves of various arum types which they pick or from which they tear off large pieces with their beaks. They form the pieces into a ball in their mouths and swallow these large chunks. The leaves are ground into a fine mash in their huge crops, which are extremely muscular, with horny ridges, and are divided into several sections. The mash then passes through the small gizzard and the short intestine. The crop is fifty times as large as the gizzard and represents 13% of the entire weight of the bird. In no other bird is the crop comparatively as large. See also **Galliformes.**

HOB. A milling cutter with form-type teeth of helicoidal shape and with profiles such that conjugate surfaces on cylindrical parts may be machined by rotating the work and the hob at a constant velocity ratio. Hobs are extensively used for cutting spur gears, and hobbing is the only really precise method of cutting heavy-duty worm wheels. Two types of gear hobs are commonly used; the radial or infeed type, and the tapered or tangential feed hob. The latter is superior, particularly for hobbing worm gears with high helix angles and high pressure angle. Hobbing processes are also used for spline cutting, and for generating ratchet teeth.

The term hobbing is also used to designate a method of die sinking, in which a hardened master punch, a duplicate of the part to be formed, is pressed into an unheated die blank so that the shape of the hob is reproduced in the die impression. This method of producing die cavities is simpler than die sinking by cutting away the material, since it is considerably easier to machine the surface of the hob than to machine the die cavity. It is also advantageous in the production of multiple die cavities, since a single hob can be used for a series of duplicate dies. The process is also referred to as "hubbing."

HOBBING. See **Worm Gearing.**

HOCK (or Hough). The joint at the attachment of the foot and the leg in animals which walk on the toes (digitigrade or unguligrade), commonly applied to domestic animals. It corresponds to the ankle joint of other species. Also the back of the human knee.

HODGKIN'S DISEASE. A malady characterized by a painless localized enlargement of lymph nodes, usually beginning in one side of the neck, but occasionally in the axillary or inguinal-femoral region. On examination, the mass is found to be a discrete, rubbery, painless lymphadenopathy, frequently surrounded by enlarged lymph nodes. Some patients have an intermittent evening fever alternating with afebrile periods sometimes lasting days or weeks. Pruritis is usually general and when severe is a characteristic symptom. Profound anemia develops in some cases at the onset of the disease, but more commonly is seen during the course.

The cause of the disease is unknown, but is probably of viral origin because high titers to Epstein-Barr virus are found in the sera of victims. Patients with Hodgkin's disease have a defect in delayed hypersensitivity and, in general, the more advanced the clinical extent of the disease, the more complete is the loss of immunological reaction. Because of this defect in cellular immunity, Hodgkin's patients are particularly susceptible to viral and bacterial infections.

Diagnosis is by examination of excised lymph tissue; histology shows destruction of nodal architecture with proliferation of abnormal reticulum cells and the development of characteristic Reed-Sternberg giant cells. The reticulum cells are monocytic macrophages having surface receptors for crystalline fragments of immunoglobulin and a tendency to ingest IgC. Four variants are distinguished by histology: *lymphocyte predominant*, in which the infiltrate consists mainly of small lymphocytes with a small number of histocytes; the *nodular sclerosing form*, in which tumor nodules are separated by collagenous connective tissue; the *mixed cellular form*, presenting a diffuse architectural replacement with the infiltrate containing conspicuous granulocytes and plentiful Reed-Sternberg cells; and the *lymphocyte-depleted form*, in which there are seen large numbers of H.D.-2 atypical reticulum cells, often pleomorphic, with bizarre mitoses and Reed-Sternberg cells.

Surgical excision, irradiation, and chemotherapy all have a place in the treatment of patients with Hodgkin's disease. Surgical excision can be used when the condition is localized, followed sometimes by local irradiation and/or chemotherapy. X-irradiation alone is valuable for localized disease. Massive doses in the early stages can produce dramatic results, including the rapid disappearance of masses and long remissions. Nitrogen mustard has been beneficial in patients with disseminated disease. Other drugs which have been used include chlorambucil, cyclophosphamide, and vinblastine sulfate.

Hodgkin's disease has a bimodal, age-specific incidence rate in the United States and northern Europe. There is a high rate between the ages of 15 and 34 and after the age of 50 years. The first age mode appears to be absent in Japan. Hodgkin's disease in children under 10 years of age is seen much more frequently in some of the developing countries of Latin America and the Middle East than in the United States and is found in boys from 8 to 10 times more frequently than in girls. In the United States, the incidence of Hodgkin's disease is about 30 cases per million population per year, but the incidence varies widely by sex, age, and socioeconomic status.

The clinical course of the disease can be extremely variable. In addition, almost all patients receive treatment that may profoundly affect the course of the disease. Sometimes the treatment results in apparent cure, and sometimes it produces complications that become difficult to separate from the disease itself. However, in time, nearly all patients with untreated or uncontrollable Hodgkin's disease develop increasingly severe systemic symptoms. High continuous fever, drenching night sweats, malaise, fatigue, anorexia, and weight loss characterize the terminal picture.

Hodgkin's disease no longer can be considered inevitably fatal. No matter what the stage of disease, patients now have the potential for cure, although the probability of cure ranges between 25 and 90%.

See also **Immune System and Immunology.**

Additional Reading

Canellos, G. P., et al.: "Chemotherapy of Advanced Hodgkin's Disease with MOPP, ABVD, or MOPP Alternating with ABVD," *N. Eng. J. Med.*, 1478 (November 19, 1992).

Davis, T. H., et al.: "Hodgkin's Disease, Lymphomatoid Papulosis, and Cutaneous T-Cell Lymphoma Derived from a Common T-Cell Clone," *N. Eng. J. Med.*, 1115 (April 23, 1992).

Hancock, S. L., Cox, R. S., and I. R. McDougall: "Thyroid Diseases after Treatment of Hodgkin's Disease," *N,. Eng. J. Med.*, 599 (August 29, 1991).

Melby, J. C., and A. L. Vickery, Jr.: "A 27-Year-Old Woman with Hodgkin's Disease and an Adrenal Mass," *N. Eng. J. Med.*, 400 (February 7, 1991).

Urba, W. J., and D. L. Longo: "Medical Progress: Hodgkin's Disease," *N. Eng. J. Med.*, 678 (March 5, 1992).

R. C. Vickery, M.D., D.Sc., Ph.D., Blanton/Dade City, Florida.

HODOGRAPH. In general (mathematics), the locus of one end of a variable vector as the other end remains fixed. A common hodograph in meteorology represents the vertical distribution of the horizontal wind.

HOGBACK. Ridge-like topographic features, the result of the differential erosion of highly tilted hard and soft strata. The steeper, or dip-slope, side is developed on the harder or less soluble formation. while the gentler slope is developed on the opposite side, on the softer rocks.

HOIST. Any device for lifting materials, weights, articles, etc., may be called a hoist. Hoists often compose a part of other apparatus whose purpose may extend to movement of material other than vertically. For example, the bridge crane incorporates within it a hoist for vertical lift. The energy required for lifting is derived ultimately from a number of various sources. For example, in the hoisting field one finds such varied power sources as compressed air, internal combustion engines, hydraulic power, steam and electric power. The pneumatic drives may be either a direct lift supplied by air acting on a piston connected directly to the load, or it may be employed in compressed air engines, whose crankshaft is geared to the hoisting apparatus. In the internal combustion engine type hoist, the gasoline engine is generally used for the light-capacity hoist, and the Diesel engine for heavier hoists. It has the advantage over other drives for portable service, such as locomotive cranes, and power shovels.

The essential parts of a hoist are a rope or chain which is wrapped around a drum or drive sheave. A hook, grapnel magnet, or other device for handling the load is attached to the free end. The rotation of the drum winds up the rope, thus shortening the distance between the drum and the load. If the drum is fixed in position over the load, naturally the load must be hoisted. To drive the drum, one of the power supplies just mentioned is connected with the drum through a suitable speed-reducing, torque-increasing mechanism. A gear train is often used. These component parts when supplied with a brake controlling the speed during lowering of weights, are the essential elements of all hoists except the direct-acting.

HOLLERITH. Pertaining to a widely used system of encoding alphanumeric information onto cards (described by American National Standard ANSI X3.26–1970). The term Hollerith cards is synonymous with punch cards. Such cards were first used in 1890 for the United States Census and were named after Herman Hollerith, their originator.

HOLLYHOCK. See **Malvaceae.**

HOLLY TREES AND SHRUBS. Of the family *Aquifoliaceae* (holly family), genus *Ilex*, there are numerous species of hollies and many hybrids and cultivars, making both nomenclature and generalization difficult. The plants may be deciduous or evergreen. They often are spiny with leathery leaves. The flowers frequently are white, usually polygamous. They bear small fruit and can withstand full sun or partial shade. The hollies tend to be more resistant to pests than most plants. Some of the important varieties include:

American holly	*Ilex opaca*
Azorean holly	*I. perado*
Chinese holly	*I. pernyi*
Dahoon holly	*I. cassine*
English holly	*I. aquifolium*
Highclere hybrid holly	*I. × altacierensis*
Longstalk holly	*I. pedunculosa*
Posshmhaw holly	*I. decidua*
Tarajo holly	*I. latifolia*

Depending upon height, the American holly may be considered a shrub or a tree. The plant can range from about 15 to 30 feet (4.5 to 9 meters) in height, although as shown by the accompanying table, under favorable conditions, the plant can develop into a very sizeable tree. The foliage may be described as being of a bronze-green or olive-green color. The leaves are glossy and quite spiny, but less so than the English species. The leaves are from 2 to 3 inches (5 to 7.6 centimeters) in length. The fruit is a scarlet red, sometimes (in the 'Xanthocarpa') a bright yellow, and a little over $\frac{1}{4}$ inch (0.6 centimeters) in diameter. It is berrylike on short stems and often clings to the plant throughout most of the winter months. With proper care, there are numerous areas in the United States where the plant does quite well. The natural occurrence generally follows the coastal regions. Sheltered locations are preferred. The plant ranges from Massachusetts southward into Florida and westward to the Mississippi Valley south of lower Indiana and Illinois.

As the name suggests, Azorean holly is found on the Azores and also on the Canary Islands. This species is a small evergreen tree with dark green foliage and deep red berries. *I. pernyi* is found in central and

RECORD HOLLY TREES IN THE UNITED STATES[1]

Specimen	Circumference[2]		Height		Spread		Location
	Inches	Centimeters	Feet	Meters	Feet	Meters	
DAHOON HOLLY							
Ilex cassine (1975)	34	86	72	21.9	22	6.7	Florida
Myrtle holly (1972) (*Ilex myrtifolia*)	67	170	46	14.0	35	10.7	Florida
HOLLY							
American holly (1991)[3] (*Ilex opaca*)	135	343	55	16.8	51	15.5	Virginia
American holly (1987)[3] (*Ilex opaca*)	119	302	74	22.6	48	14.6	Alabama
Carolina holly (1989) (*Ilex ambiqua*)	74	36	25	6.7	18	5.2	Florida
Gallberry holly (1973) (*Ilex coriacea*)	10	25	27	8.2	12	3.7	Virginia
Silver Varigated holly (1977) (*Ilex aquifolium*)	75	191	40	12.2	22	6.7	Oregon
Tawnberry holly (1990) (*Ilex krugiana*)	40	102	55	16.8	22	6.7	Florida
POSSUMHAW HOLLY							
Ilex decidua (1981)	36	91	42	12.8	52	15.8	South Carolina
WINTERBERRY HOLLY							
Common winterberry holly (1991)[3] (*Ilex verticillata*)	10	25	17	5.2	14	4.3	Mississippi
Common winterberry holly (1991)[3] (*Ilex verticillata*)	10	25	19	5.8	12	3.6	Virginia
Mountain winterberry (1989) (*Ilex montana*)	30	76	28	8.5	36	11.0	New York
YAUPON HOLLY							
Ilex vomitoria (1964)	49	124	45	13.7	40	12.2	Texas

[1]From the "National Register of Big Trees," The American Forestry Association (by permission).
[2]At 4.5 feet (1.4 meters).
[3]Cochampions.

westward China and is a narrow tree of pyramidal form that can rise to a height of about 30 feet (9 meters). The leaves are small, leathery, and of a lustrous dark green color. The tree bears clusters of small red berries.

The Dahoon holly is found in the southeastern United States and is characterized by a somewhat heavier trunk than found on most hollies. The leaves are evergreen and narrow, about 2 to 4 inches (5 to 10 centimeters) in length. Their color is dark green. The flowers and fruit are similar to the American holly. The plant ranges from the southern part of Virginia to Florida and along the Gulf coast west to Louisiana.

English holly occurs naturally in western Asia, northern Africa, and southern Europe, as well as the British Isles from which it derives its name. However, several forms of this holly are not so hardy in England as they may be on the continent. Because of the numerous hybrids, a great variety of leaf and fruit colorations, as well as other characteristics of the shrubs and trees, is obtainable. Some of the more important varieties include: Perry's weeping silver holly ('Argenteomarginata Pendula'), which has silver foliage and lots of berries; the silver milk-boy ('Argenteo-Medico Picta'), which grows to a height of about 30 feet (9 meters) and has dark green, spiny leaves with cream-colored spots in their central portion; the golden queen ('Aurea Regina'), with yellow-edged dark green leaves; and the silver hedgehog holly ('Ferrox Argentea'), which has leaves featuring white spines and margins, and ranging up to 15 feet (4.5 meters). Other varieties include the 'Bacciflavia,' the 'Crispa,' the 'Elegantissima,' the 'Ferox,' and the 'Hastata.'

The highclere hybrid hollies are known for their vigor. They have quite large evergreen leaves. They are known for their toleration of industrial and seaside environments. They range in height from small bushes to trees of 50 feet (15 meters). These hybrids were obtained by crossing the English holly with the Azorean holly. Some of the more important varieties include: 'Camellifolia,' a tree of conical contour, characterized by a purple bark, almost spineless evergreen leaves that are purple when young, later turning a dark green; the 'Golden King,' which has green leaves with yellow edges, nearly spineless; the 'J. C. van Tol,' almost spineless leaves of dark-green color and produces large quantities of berries; the 'Lawsoniana,' which has large leaves with yellow borders and marbleized centers; the 'Purple Shaft,' which is known for its vigor and large quantities of berries; and the 'Silver Sentinel,' which has mottled leaves that are flat and almost spineless.

The longstalk holly is found in Japan. It ranges from a shrub to a small tree of about 30 feet (9 meters) in height. The plant has evergreen leaves and small red fruits.

The Possumhaw holly is found in the southeastern United States. It is also sometimes referred to as the swamp holly. Normally, it is a small shrub or tree, but as shown by the accompanying table, the plant can attain very respectable dimensions under favorable conditions. The leaves are a lustrous deep green, deciduous, and from $1\frac{1}{2}$ to 3 inches (7.6 centimeters) in length. The flowers are similar to those of the American holly. Its natural range is between the Atlantic coast and the Appalachian Mountains south of Virginia and into western Florida and westward to Arkansas, Missouri, and Texas.

The Tarajo holly is found in Japan. This species can attain a height of 60 feet (18 meters) or more and features the largest leaves of any holly. The leaves are evergreen of a dark green color, yellow underneath. The fruit occurs in large numbers of orange-red clusters.

HOLMIUM. Chemical element symbol Ho, at. no. 67, at. wt. 164.93, tenth in the Lanthanide Series in the periodic table, mp 1,474°C, bp 2695°C, density 8.795 g/cm^3 (20°C). Elemental holmium has a close-packed hexagonal crystal structure at 25°C. The pure holmium is silver-gray in color, slow to tarnish or oxidize at room temperature in normal atmospheres. Even at relatively high temperatures, the metal is slow to oxidize. Under a vacuum of about 10 torr, holmium will react when hot with water vapor, CO_2, NH_3, and hydrocarbons. Holmium is soft and can be worked by conventional equipment. There is one natural isotope of holmium, ^{165}Ho, and 18 artificial isotopes have been produced. The natural isotope is not radioactive. In terms of abundance, holmium is present on the average of 1.2 ppm in the earth's crust, ranking ahead of bismuth, antimony, cadmium, and mercury in potential availability. The element was first identified by P. T. Cleve and J. L. Soret in 1879. The metal has a low acute-toxicity rating. Electronic configuration $1s^22s^22p^63s^23p^63d^{10}4s^24p^64d^{10}4f^{10}5s^25p^65d^16s^2$. Ionic radius Ho^{3+} 0.894 Å. Metallic radius 1.766 Å. Other important physical properties of holmium are given under **Rare-Earth Elements and Metals.**

Holmium occurs in apatite, xenotime, and yttrium and heavy rare-earth minerals. The element of a purity of 99.9% can be obtained through organic ion-exchange techniques. Supplies of holmium are available commercially as the result of yttrium production. To date, the applications for holmium have been very limited. When added to orthoferrites, it has shown promise for use in electronic circuits. Uses in semiconductors, lasers, thermoelectric devices, phosphors, and ferrite bubble devices currently are being studied.

See references listed at ends of entries on **Chemical Elements;** and **Rare-Earth Elements and Metals.**

NOTE: This entry was revised and updated by K. A. Gschneidner, Jr., Director, and B. Evans, Assistant Chemist, Rare-Earth Information Center, Energy and Mineral Resources Research Institute, Iowa State University, Ames, Iowa.

HOLOCRYSTALLINE. The term applied by petrologists to igneous rocks composed entirely of crystals; in contradistinction to igneous rocks which are partly or entirely composed of natural glass, such as obsidian.

HOLOENZYME. See **Coenzymes.**

HOLOGRAPHY. The technique of holography is similar to photography in many respects, yet it is fundamentally different. With photography, one generally records, by means of lens and film, the two-dimensional irradiance distribution in the image of an object. With holography, one records not the optically formed image of an object, but the object wave itself. This wave is recorded (frequently on photographic film) in such a way that a subsequent illumination of this record, called a *hologram*, reconstructs the original object wave. A visual observation of this reconstructed wavefront then yields a view of the object which is practically indiscernible from the original, including three-dimensional parallax effects. The process was discovered by Gabor[1] (England) in 1948. It was then identified as a two-step method of optical imagery. During the past couple of decades, holography has become widely known and a limited number of practical uses for it have been developed. This later progress is attributed to the general availability of the laser, with the outstanding temporal and spatial coherence of its light. Much of the work in adapting the laser to holography was carried out by Upatnieks and Leith (University of Michigan) during the early 1960s.

With reference to Fig. 1, one starts with a single, monochromatic beam of light that has originated from a very small source. This single beam is split into two components, one of which is directed toward the object and the other to a suitable recording medium, most commonly a photographic emulsion. The component that is incident on the object is scattered by it, and this scattered radiation, now called the object wave, impinges on the recording medium. The wave that proceeds directly to the recording medium is called the *reference wave*. Since the object and reference waves originate from the same source, they are mutually coherent and form a stable interference pattern when they meet at the recording medium. The detailed record of this interference pattern constitutes the hologram.

Types of Holograms

When the hologram is illuminated with a beam similar to the original reference wave, it modulates the phase and/or amplitude of the illuminating wave in such a way that the transmitted wave divided into three

[1]For which he received the Nobel Prize in physics (1971).

illuminating beam so that all of the rays of the reference beam are reversed in direction. In this way, an undistorted, real, three-dimensional image of the object scene appears in front of the hologram, as shown in Fig. 1(c).

Holograms may be recorded with diverging, parallel, or converging reference beams. If care is taken to maintain the recording geometry during reconstruction, it is possible to form holograms with an arbitrary reference beam, the only requirement being that it be coherent with the object beam.

Color holograms can be produced by recording three separate holograms on a single photographic plate, each in a different color. Subsequent illumination with a three-color beam yields three separate wavefronts, one in each of the three colors representing the portion of the object corresponding to that color.

Holograms also can be made that can be viewed in reflection. This is done by allowing the reference and object beams to enter the recording medium from opposite sides. The fringes formed are planes lying approximately parallel to the plane of the hologram. When such a hologram is illuminated by a beam similar to the reference wave, a reflected wave is formed which exactly duplicates the object wave. The image is viewed in reflected light. This type of hologram can be illuminated with white light. The interference planes filter the light by acting as a $\lambda/2$ multilayer interference filter, in the same way as in Lippmann color photography.

One of the most striking aspects of the modern hologram is the three-dimensional image that it is capable of producing. The three-dimensional image indicates that there is a large amount of information contained in a single hologram—much more than is contained in a conventional photograph of the same size. Because of the many perspectives available, the hologram is well suited to display purposes. With a hologram, one can present all of the observable characteristics of a three-dimensional object clearly and concisely. Complex molecular or anatomical structure can be simply presented with a single holographic image, with little chance of error or misinterpretation on the part of the viewer. Thus holograms may reduce the number of conventional drawings or photographs to illustrate a single object. It has been proposed that the use of holograms in textbooks would be an aid to readers, particularly in fields where three dimensions are important. Holograms can be made to be viewed with a small penlight and a colored filter.

Applications of Holography

Early applications of holography, essentially prior to the early 1980s, were more of a novel than scientific nature. In 1984, Chang pointed out that holography is more than simply a curiosity related to three-dimensional photography. It is a technology involving the precise structure of light waves, with advanced implications for solutions to engineering problems. Engineering applications of holography utilize the interference patterns created by superimposing holographic images from a target made under slightly different conditions. The patterns can be examined visually or the data can be digitized for computer analysis. In realtime holography, the object is viewed through a hologram of itself. Double-exposure holography involves recording the interference patterns obtained from the same target before and after distortion. Time-average holograms, employed for vibration analysis for example, are made by exposing the plate while the object is driven in resonance.

As described by Chang, some of the more recently developed applications for holography include: (1) Analysis of dimensional instability when an object is stressed. Small distortions resulting from the applications of forces, changes in environment, and other factors yield interference fringes in the superimposed images equivalent to strain contour lines. The approach has been used in connection with thermally induced changes in large-dish antennas at very low temperatures, for observing the performance of miniature gyroscopes, and for examining deformations due to pressure in a vessel or pipe. (2) Checking for cracks in weldments by seeking fringe discontinuity across a seam. (3) Determining voids in layered objects, including the inspection of composite aircraft components, clutch plate facings, multiple-layer circuit boards, tires, O-rings, antique paintings, nuclear fuel rods, and detecting the delamination in the composite blades of a helicopter, among others. See Fig. 2. (4) Vibration analysis of such components and subsystems as turbine blades, loudspeakers, rocket castings, and automobile engines. See Fig. 3. (5) Studies for flow visualization in connection with air foils, plasmas, and combustion flames—as an alternate to Schlieren photography. (6) Studies of biological and crystal growth. (7) Image processing in connection with pattern recognition in robotics and in finding defects in parts, such as integrated circuits. (8) Monitoring materials properties, such as index of refraction. (9) Analysis of particle size distribution and movement, of interest in improving combustion efficiency and for checking particulate contamination of food and drug products. (10) Use of holograms instead of lens systems for transforming light beams and images and thus use in optical scanners, diffraction gratings, and pattern generators, among others. (11) Microscopic and interferometric studies where objects may be only several microns in diameter—without depth-of-field limitations.

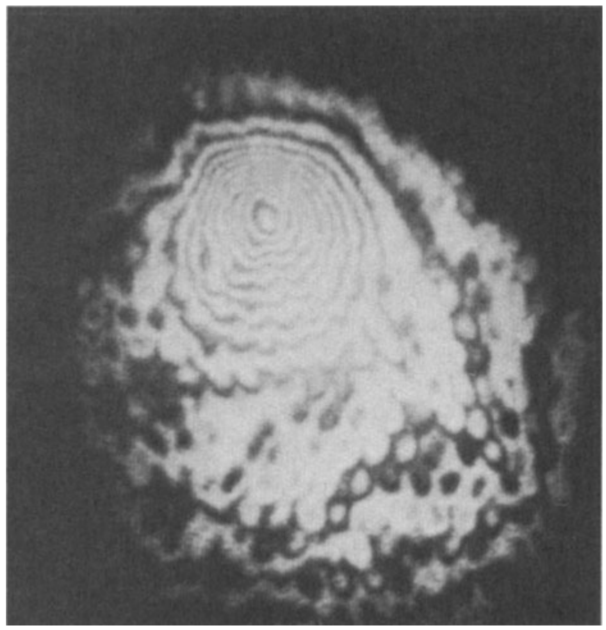


Fig. 2. Holographic interferometry used here to determine delamination in the bonding between skin and honeycomb structure of a composite helicopter rotor blade. *(After M. Chang, Newport Corp.)*

As the potential for holographic techniques become better known, more sophisticated uses are being uncovered. One such application is picosecond holographic-grating spectroscopy. As reported by Wiersma and Duppen (1987 reference listed), interfering light waves produce an optical interference pattern in any medium that interacts with light. This modulation of some physical parameter of the system acts as a classical holographic grating for optical radiation. When such a grating is produced through interaction of pulsed light waves with an optical transition, a transient grating is formed whose decay is a measure of the relaxation time of the excited state. Transient gratings can be formed in real space or in frequency space, depending on the time ordering of the interfering light waves. The two gratings are related by a space-time transformation and contain complementary information on the optical dynamics of the system. The status of a grating can be probed by a delayed third pulse. which diffracts off this grating in a direction determined by the wave vector difference of the interfering light beams. This generalized concept of a transient grating can be used to interpret many picosecond-pulse optical experiments on condensed-phase systems. In their paper, Wiersma and Duppen illustrate some low-temperature experiments. The impact of nonlinear photon-interference spectroscopy on the field of transient-grating and more generally on four-wave mixing spectroscopy is currently significant and is expanding.

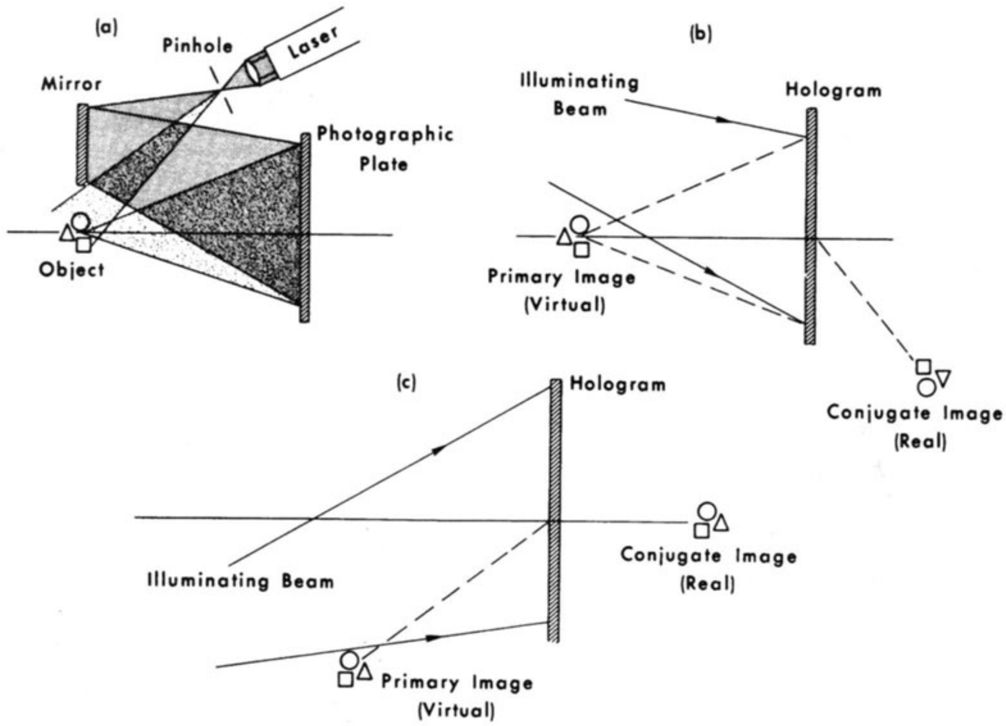


Fig. 1. A typical holographic arrangement: (a) Recording the hologram; (b) reconstructing the primary object wave; (c) reconstructing an undistorted conjugate wave.

separate components, one of which exactly duplicates the original object wave.

If the two interfering beams are traveling in substantially the same direction, the recording of the interference pattern is said to be a *Gabor hologram* or *in-line hologram*. If the two interfering beams arrive at the recording medium from substantially different directions, the recording is a *Leith-Upatnieks* or *off-axis hologram*. If the two interfering beams are traveling in essentially opposite directions, the recorded hologram is said to be a *Lippmann* or *reflection* hologram, first invented by Denisyuk.

Electromagnetic radiation is most commonly used, although acoustic radiation can be used. The most common electromagnetic radiation employed is light, but holograms have also been recorded successfully with electron beams, x-radiation, and microwaves.

Holograms can be classified by the way they diffract light. In an *amplitude hologram*, the varying irradiance distribution of the interference pattern is recorded as a density variation of the recording medium. In this type of hologram, the illuminating wave is always partially absorbed, i.e., the illuminating wave is *amplitude-modulated*. In the *phase hologram*, a *phase modulation* is imposed on the illuminating beam which, in turn, results in diffraction of the light. Phase modulation occurs when the optical path (thickness × index) varies with position. A phase hologram results from either relief-image or index variation, or both.

Either phase or amplitude holograms can be classified further as *Fresnel holograms* or as *Fraunhofer holograms*. Generally speaking, if the object is reasonably close to the recording medium, say just a few hologram or object diameters distant, the field at the hologram plane is the Fresnel diffraction pattern of the object. A hologram recorded in this manner is termed a *Fresnel hologram*.

If the object and hologram are separated by many object or hologram diameters, the field at the hologram due to the object alone is the Fraunhofer diffraction pattern of the object. A hologram recorded in this manner is termed a *Fraunhofer hologram*.

Any of these holograms types may be recorded as either a *thick* or a *thin* hologram. A thin hologram is one for which the thickness of the recording medium is thin compared to the space between the recorded interference fringes. A thick or volume hologram is one in which the thickness of the recording medium is of the order of or greater than the spacing of the recorded fringes.

Conceptually, the simplest form of an off-axis hologram is one for which the object is just a single, infinitely distant point so that the object wave at the recording medium is a plane wave. If the reference wave is also plane, and incident on the recording medium at an angle to the object wave, the hologram will consist of a series of Young's interference fringes. These recorded fringes are equally spaced straight lines running perpendicular to the plane of incidence. Since the hologram consists of a series of alternating clear and opaque strips, it is in the form of a diffraction grating. When the hologram is illuminated with a plane wave, the transmitted light consists of a zero-order wave traveling in the direction of the illuminating wave, plus two first-order waves. The higher diffracted orders are generally missing or very weak, inasmuch as the irradiance distribution of a two-beam interference pattern is sinusoidal. As long as the recording is essentially linear (irradiance proportional to final amplitude transmittance), the hologram will be a diffraction grating varying sinusoidally in amplitude transmittance, and only the first diffracted orders will be observed. One of these first-order waves will be traveling in the same direction as the object wave. This is the reconstructed wave.

Holographic Recording

The recording of a hologram and the subsequent reconstruction is shown in Fig. 1. In Fig. 1(a), the laser beam is first expanded and then divided by a mirror, which directs part of the beam directly onto the photographic plate; the rest of the light is reflected from the object. After processing, the hologram plate may be replaced in its original position (Fig. 1(b)), and the object removed. The light diffracted by the hologram forms, in part, the same wavefront that was originally scattered by the object. A viewer looking through the hologram will see an undistorted view of the object, just as if it were still present.

In addition to the *virtual* or *primary image*, a real, or *conjugate image* will be formed on the observer's side of the hologram. This image will appear unsharp and highly distorted, and it will also be inverted in depth, i.e., reversed front to back, as shown in Fig. 1(b). However, a distortion-free real image can be formed by changing the position of the

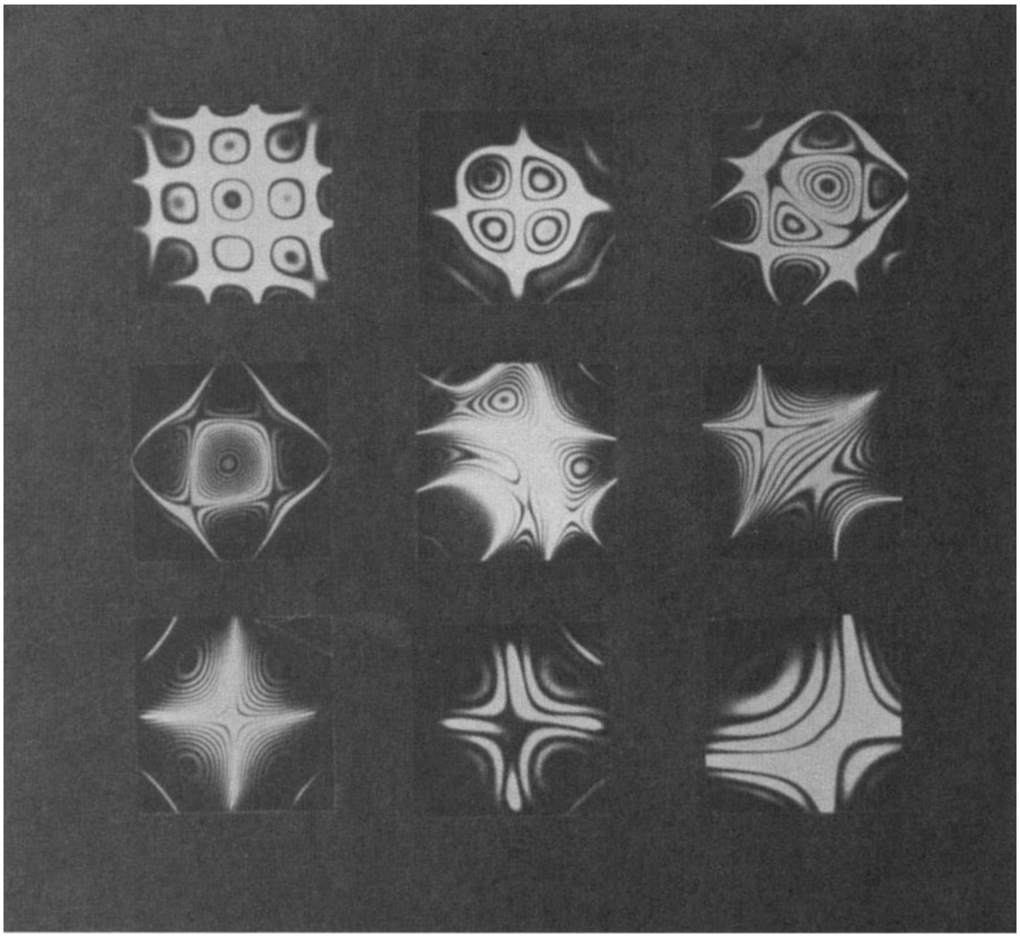


Fig. 3. Examples of resonant modes of plate as obtained through time-averaged holographic interferometry. (*After M. Chang, Newport Corp.*).

HOLOHEDRAL CRYSTAL. A crystal in which the full number of faces are developed, corresponding to the maximum and complete symmetry of the system. See **Mineralogy.**

HOLOTHUROIDEA. The sea cucumbers, a class of the phylum *Echinodermata*.

These animals differ from other echinoderms in several particulars: (1) The principal axis is elongated and the animal rests on its side. (2) The body wall is soft because of the reduction of the calcareous ossicles. (3) A branching respiratory tree extends from the alimentary tract into the body cavity.

Sea cucumbers are used as food in the Oriental region. They are dried for the market and in this form are called trepang or bêche-de-mer.

The class includes five orders:

Order *Aspidochirota*. Tropical species with shield-shaped tentacles. In shallow water.
Order *Elasipoda*. Benthonic species of deep water.
Order *Dendrochirota*. Shallow water species with branching tentacles.
Order *Molpadonia*. Burrowing species. Tentacles unbranched or slightly branched.
Order *Apida*. (*Synaptida, Paractinopoda*.) Burrowing species without respiratory trees.

HOLOTYPE. A term used by biologists and paleontologists to mean the specimen to which all others should ultimately refer to determine the species. The holotype does not necessarily have to be the originally described species (type) and frequently is not.

HOMEOSTASIS. Maintenance of the steady state. As applied to living organisms, this refers to the many adjustments which are constantly being made to keep the organism in a rather constant environment internally in spite of the fact that there may be many variations in external environment. Life within individual cells can continue only within a rather narrow range of conditions, and each form of life possesses many self-regulating systems whereby it can maintain a favorable internal environment in spite of the great variations in its surroundings. When a person goes from bright sunlight into a dark room the eyes undergo certain changes as a result of automatic internal adjustments which permit the eyes to function in spite of the greatly reduced light intensity. A person living in arctic regions of the north and persons on a tropical beach have an internal temperature that does not vary more than a fraction of a degree. Internal thermostatic adjustments regulate the body temperature to keep it at such a constant level. The human brain must have a blood supply at a constant pressure; a slight drop in pressure brings a "blackout" and too great a pressure will cause the bursting of capillaries and a "stroke." Homeostatic mechanisms change the beat of the heart and the force of the blood to the head and thus regulate the blood pressure at a constant level. We sometimes say that these mechanisms maintain the steady state.

Frequently, homeostatic regulations are by means of the negative feedback mechanism. As an example the male hormones, androgens, of vertebrate animals inhibit the production of gonadotropin from the pituitary gland. Gonadotropin, on the other hand, stimulates androgen production by the testes. In this way, the level of androgens in the blood is kept within rather close tolerances. A castrated animal will show a sudden rise in the gonadotropin in the blood and urine due to the removal of the androgens which usually serve as a control. (See Fig. 1). See also **Hormones.**

A good example of homeostasis in plants concerns the maintenance of the water balance in the leaves. The leaf must maintain a steady state of water concentration in spite of great variations in the amount of water available in the soil and variations in the humidity and tempera-

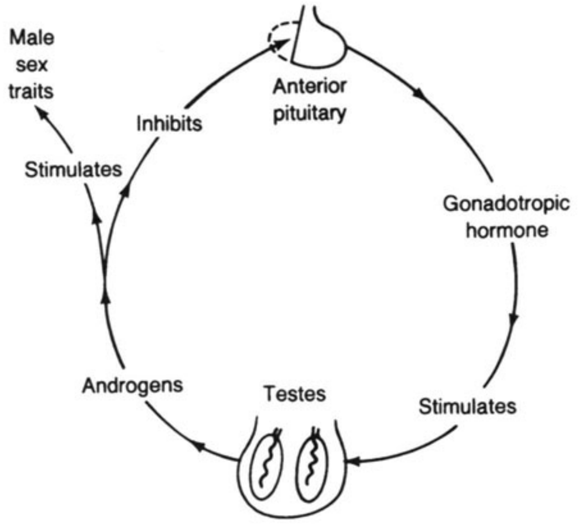


Fig. 1. Homeostatic regulation of male hormone. (*A. M. Winchester.*)

ture of the air, which affect the loss of water from the leaves through transpiration. The leaves have tiny stomata which admit air to the leaf. This air is needed to supply the carbon dioxide for photosynthesis, but it can also carry moisture from the leaf. The guard cells surrounding the stomata minimize the loss of water by opening and closing the stomata in accordance with the amount of water in the leaf. The guard cells have a tough inner portion that, as the cells swell with turgor pressure, becomes more convex and opens the stoma in between. When the guard cells lose turgor pressure, the inner portions become less convex and the stoma is closed, thus preventing the loss of more water when the water level drops in the cells. The guard cells also function according to the usage of carbon dioxide during photosynthesis. As the carbon dioxide level drops, some of the starch is converted to sugar. This increase of solutes within the cell causes the cell to absorb more water from the surrounding cells and the stoma is opened. At night when no carbon dioxide is being used, the sugar concentration is low and the stoma is closed. The concentration of carbonic acid, from the carbon dioxide in the cells, is the factor which activates the starch-splitting enzymes. With low carbonic acid, there is an inactivity of the enzymes, with high acid, the enzymes are most active. (See Fig. 2.)

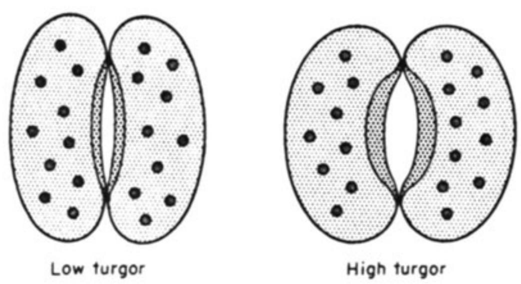


Fig. 2. Guard cell closing. (*A. M. Winchester.*)

There can also be homeostasis of a genetic nature. The balance of genes within a gene pool is homeostatically regulated. Suppose a certain harmful gene is continually being added to a population through mutation. As the gene is expressed it will reduce the reproductive potential of the individual and there will be a gradual elimination of the gene from the population. Soon the input through mutation is exactly balanced by the outgo through genetic death and the gene remains at a stable level in the population. If we increase the mutation rate, say, by radiation, then the concentration in the gene pool will increase. If we reduce the rate of elimination by medical means, we can also increase the gene pool.

Ecological homeostasis concerns the balance of nature. When certain plants which serve as food for herbivores increase, the number of herbivores increases. This, in turn, results in an increase of the predators. A balance is established that will vary according to variations in environmental factors which cause an increase or decrease of any one of the organisms in the complicated food web.

HOMEOTYPE. A term used by biologists and paleontologists for a specimen which has been identified by an authority by comparing it with the type.

HOMOCENTRIC RAYS. Rays having the same focal point. (It may be at infinity; in other words, the rays may be parallel.)

HOMOCLINE. Group of strata which dip in one and the same direction. Never a complete structure and usually representing the limb of an anticline or syncline.

HOMODYNE RECEPTION. In this system, used in connection with radio reception for suppressed-carrier systems of radiotelephony, the receiver generates a voltage which has the original carrier frequency. This is combined with the incoming signal. The term zero-beat reception is also used.

HOMOGENEOUS (Mathematics). The term has several meanings. Inhomogeneous is the opposite of homogeneous.

A function $f(x_1, x_2, \ldots, x_n)$ is homogeneous in all of its variables if, for any parameter t, $f(tx_1, tx_2, \ldots, tx_n) = t^n f(x_1, x_2, \ldots, x_n)$. The exponent n is the degree or order of the function. The behavior of such a function is known as the Euler theorem on homogeneous functions.

The term is used with two meanings for a differential equation:

1. A first order equation, $y' = M(x, y)/N(x, y)$ is a homogeneous equation if M, N are homogeneous functions of the same degree.
2. The general equation, $f(x, y, y', y'', \ldots) = 0$ is homogeneous and linear if f is a homogeneous function of y and all its derivatives. If the right-hand side equals a function of x, the independent variable, it is still linear but now inhomogeneous.

An integral equation, a boundary condition, or a system of simultaneous linear algebraic equations can also be homogeneous or inhomogeneous in a similar way.

HOMOGENIZING. A process for reducing the size of particles in a liquid and useful in the preparation of numerous food substances, including milk, ice cream, salad dressings, various fruit juices, flavor concentrates, infant foods, among others.

A reduction of particle or globule size in a mixture of two immiscible liquids makes an emulsion possible. If an emulsifying agent is present, a more stable emulsion can be produced and coalescence of the dispersed phase is prevented. The homogenizer is also used to produce dispersions by reducing the particle size in solid-in-liquid mixtures. As in the preparation of an emulsion, a dispersing agent is needed to maintain a homogeneous mixture.

Typically, a homogenizer consists of a high-pressure, positive-displacement pump and an adjustable orifice. The pump is a piston or plunger type, usually consisting of three plungers, although some homogenizers are made with five or even seven plungers. The cylinder for each plunger has an inlet and discharge valve. The plunger pump must push the product through the homogenizing valve (adjustable orifice). For two-stage homogenization, two valves are arranged in series.

A typical homogenizing valve consists of a seat and plug of very hard abrasion-resistant materials (alloys such as Stellite are used). The seating surfaces must be lapped smooth and be parallel. In operation, the plug is spring-loaded against the seat. Spring compression is adjusted so that when the product flows, energy in the form of pressure is required to lift the plug. Although many products can be homogenized at pressures below 3000 pounds per square inch (204 atmospheres), machines are made to develop pressures in excess of 8000 pounds per square inch (544 atmospheres). In another design, a valve uses a compressed cone of stainless-steel wire inserted into a socket, the product being homogenized by flowing between the wires.

A number of theories have been proposed as to what actually breaks up the particles in the homogenizer: (1) As the product enters the area between the lapped surfaces, it is suddenly accelerated to velocities as high as 30,000 feet per minute (9,144 meters per minute) at a pressure of 5,000 pounds per square inch (340 atmospheres). When acceleration is this sudden, the particle (especially the liquid particle) is stretched or elongated to the point of breaking. (2) At this high velocity, there are shear forces between layers of liquids under flow that break up particles. (3) Cavitation may be the major cause of homogenization. When the pressure energy is converted into velocity energy, the vapor pressure of the product exceeds product pressure, resulting in the formation of vapor cavities which collapse upon leaving the valve at higher pressures. This collapsing, or implosion, of cavitation exerts tremendous force, breaking up the particles. Most homogenizers are designed to incorporate one or more of the foregoing principles.

HOMOIOTHERMY. Warm-bloodedness. The maintenance of a body temperature above that of the environment is common among animals; hence the usual terms warm-blooded and cold-blooded are inaccurate. Cold-blooded forms are those whose body temperature fluctuates with that of the surrounding air or water, so that the animal's activity is directly conditioned by external temperatures. They are more accurately described as poikilothermal. In contrast, homoiothermal animals tend to maintain a constant body temperature in spite of external fluctuations. Fluctuations are normal, although the human body usually maintains a constant temperature.

Only birds and mammals are homoiothermal. Both regulate the body temperature by producing excess heat and by regulating its radiation from the surface. Regulation is accomplished by nervous control of the blood vessels near the surface, by insulating vestiture, and by the evaporation of water from the body. When the surrounding air is warm, the blood flows more freely near the surface of the body and more heat is radiated, but when the air is cold, less blood reaches the surface and the heat is conserved. In air too warm to permit adequate radiation, the animal reduces its activity, exposes as much surface as possible, and either sweats or pants. The evaporation of water either from the mouth or from the sweat glands absorbs heat from the underlying tissues. Vestiture plays a passive role as an insulating coat, but it is capable of some regulation, especially in the birds. The erection of the feathers provides a thicker and looser covering of high insulating value, and their depression results in less interference with radiation.

Homoiothermy is one of the highest adaptations of living things, since it provides for the maintenance of optimum conditions for the vital processes of the body. Through it the animal becomes virtually independent of one of the most important of the fluctuating environmental conditions.

HOMOLOGOUS SERIES. Two organic compounds are said to be homologous if their molecular formulas differ by CH_2, or a multiple of CH_2. For example, the alkane series has the general formula, C_nH_{2n+2}, its first three members being methane, CH_4, ethane, C_2H_6, and propane, C_3H_8.

HOMOLOGY. Fundamental structural relationship, based on similarity of embryological development and evolutionary history. The antithesis of analogy, which is superficial likeness based on adaptation for similar uses.

The anterior appendages of terrestrial vertebrates, for example, are regarded as fundamentally similar structures, derived from the pentadactyl appendage; yet they include the wings of birds, flippers of aquatic mammals, and a great variety of less extreme adaptations, including the legs of animals and the arms of man. In contrast, the wings of birds and of insects are broad thin structures used for flight, but in structure and origin they show no resemblance beyond this point and so are analogous.

HOMOPTERA. The cicadas, leaf hoppers, plant lice, scale insects, and numerous other forms, constituting a large order of insects. They have sucking mouths which differ from those of most bugs in that the slender proboscis arises from the hind margin of the head and extends back between the legs. The wings, when present, are membraneous. The order includes about 16,000 species.

Many members of this order, particularly the plant lice, scale insects, and phylloxerans, are economically important.

The main families of Homoptera include:

Aleyrodidae	White flies
Aphidae	Aphids or plant lice
Cercopidae	Spittle bugs
Chermidae (also called *Psyllidae*)	Jumping plant lice
Cicadellidae (also called *Jassidae*)	Leafhoppers
Cicadidae	Cicadas
Coccidae	Scale insects and mealybugs
Fulgoridae	Plant hoppers
Membracidae	Tree hoppers

HONEY. Raw, unprocessed honey is a thick, viscous, high-density, very sweet, hygroscopic liquid that is formed by honeybees from the nectar of flowers and, to a limited extent, from the juices of fruits and honeydew. Honey is available commercially as a liquid, as crystallized honey, as comb honey, as chunk honey, and as powdered honey. Honey contains a large percentage of simple sugars, as well as essential oils of the flowers from which it is derived, plus about 20% water. The flavor of honey depends upon the flowers from which the nectar is derived, upon manufacturing conditions if it is processed, upon the season and climate during which it is gathered and stored by the honeybees, and upon its age. Under appropriate conditions, honey is one of the most storable of foods and can be kept for many years, particularly in a frozen state.

In food processing, honey is frequently used because of its qualities as a humectant and a source of reducing sugars. Bakers and candy makers prefer honey for these reasons plus the fact that it promotes caramelization and aids in obtaining uniform browning of baked goods, as well as providing clarity to glazes. In addition to bakery products and confections, honey is used in the manufacture of breakfast foods, snacks, sauces, and syrups, as well as a sweetener and bodying agent in some canned fruits, jams, jellies, and spreads. Honey is a common ingredient of graham crackers, where it blends with the dark whole-wheat flours, as it also does with whole-wheat breads.

Physical and Chemical Properties of Honey

Although a seemingly simple substance, honey is relatively complex and requires some rather exacting conditions when it is processed. Important factors include: (1) moisture content which, if excessive, causes the honey to ferment over a period of time; (2) the tendency of the glucose to crystallize out of the liquid phase, a process known in the trade as *granulation*; and (3) the presence of nitrogeneous substances, which even in very small amounts cause the honey to darken with age. Because water content is important to ultimate quality of the honey, including the possibility of fermentation, the water content is strictly regulated by most countries. In the United States, U.S. Grade A (Fancy) and Grade B (Choice) cannot contain over 18.6% water. Grade C (Standard) for reprocessing may contain up to 20% water. Honey with greater amounts of water is Grade D (Substandard).

Beeswax. This is a commercial byproduct of honey production. The wax represents approximately 1.9% of the weight of honey produced. In the United States, beeswax production ranges between 3 and 5 million pounds (1.4 and 2.3 million kilograms) per year. Freshly made wax is of a light yellow color, but becomes brown with age. However, the wax may be bleached by sunlight or with acids. Beeswax is made up mainly of a complex long-chain ester, myricil palmitate, $C_{15}H_{31}COOC_{30}H_{61}$, and cerotic acid, $C_{25}H_{51}COOH$; specific gravity, 0.965–0.969; mp 63°C. The wax is easily colored with dyes and finds numerous uses, as in polishes, candles, leather dressings, adhesives, cosmetics, and molded articles.

Mead (*Honey Wine*). This is one of the most ancient of fermented products and wines. Although regarded by some people as a curiosity in modern times, there is some demand for mead in various parts of the world. It is usually prepared and consumed on a regional basis.

Honey and Infant Botulism. Honey has been implicated in about one-third of the cases of infant botulism. Consequently, some physicians recommend that honey not be fed to infants under one year of age.

Additional Reading

See list of references at end of article on **Honeybees.**

HONEYBEES. There was little scientific knowledge of honeybees until the 1850s. Even though European scientists of the 1700s and early 1800s developed an understanding of the biological aspects of honey and wax production by bees, this knowledge did not contribute in a major way to the practical aspects of bee-keeping. It was not until 1852 that an American minister (L. L. Langstroth) discovered what became known as "bee space." This concept led to development of the first practical, movable-frame hives. This breakthrough, coupled with other important equipment developments, such as the centrifugal honey extractor (commercialized in 1870), the bee smoker, bee escape, and queen excluder, transformed beekeeping during the latter half of the 1800s from a minor activity on the part of many farmers to a serious business capable of centralization and full-time management. Prior to that time, there was little if any organization in terms of producing and marketing honey.

It is estimated that 60–70% of the honey produced in the United States comes from 1200 to 1500 fulltime beekeepers who operate 35–40% of the nation's nearly 5 million bee colonies. Production of honey in the world is estimated at 683,000 metric tons, of which the United States accounts for about 114,000 metric tons per year. Other large honey producing countries include Russia, China, Mexico, Uruguay, Canada, and Turkey. The principal honey producing states in the United States are Florida, North Dakota, Minnesota, California, and Wisconsin. However, honey production is found in practically all of the states to some degree.

A fulltime beekeeper maintains a minimum of 400 hives, but the average is about 1,200 hives. Large operators will maintain 20,000 or more hives. The optimal hive density for good honey yield usually is insufficient for efficient crop pollination. Thus crop growers compensate beekeepers for reduced honey crops resulting from maintaining relatively high hive density and also for the costs of moving colonies from one farm to the next during pollinating season. One of the specialty sectors of beekeeping is that of raising bees and queens for sale to beekeepers and growers of crops. This activity is concentrated in the Gulf states and southeastern states.

Strains and Hybrids

The most common strain of bees kept in North America are of Italian origin. These bees, yellow to brown in color, are industrious and relatively unexcitable. The Caucasian strain, even more gentle than the Italian, is also kept. This honeybee is gray-to-black in coloration. An import from Japan in the late 1970s, the species *Osmia cornifrons*, is under intensive study by the U.S. Department of Agriculture. It is believed that the species may be of particular value for solving pollination problems for small-farm fruit tree growers. The usual honeybee suffers from diseases and predator pests; they swarm; they sometimes abscond; they require considerable attention; and they sting. The Japanese honeybee, about two-thirds the size of the Italian or Caucasian bees, has few of the foregoing disadvantages. They produce no honey or beeswax, they are more active at cooler temperatures, they rarely range more than 300 feet (90 meters) from their nests, they are docile, and their sting is about like that of a mosquito bite. Although they produce no honey, they appear to be quite ideal for pollinating.

Biological Aspects. Bees are of the subgroup *Aculeata* within the order of *Hymenoptera* (*Insecta*). *Aculeata* also includes wasps and ants, these insects all incorporating the sting, which was developed in the course of evolution from the ovipositor apparatus. This modification involved a change of function, for it no longer serves for egg-laying; rather it is employed as an effective weapon of defense or as an injection cannula for the paralysis of prey. In view of the origin of the device, it is obvious that only female *Hymenoptera* can sting. The sting consists of several reciprocally movable chitinous elements. Into it there open the ducts of two glands, one of which produces poison. In the stinging *Hymenoptera* the eggs are ejected from the opening of the genital chamber at the base of the ovipositor.

Bees are members of the superfamily *Apoidea* and include, in addition to the universally known honeybee, more than 20,000 other species. They are distributed almost worldwide, and in the north their range extends well beyond the Arctic Circle. The smallest bees measure barely 2 millimeters, while the largest approach 4 centimeters in body length. They are among the most economically important animals, particularly since they play a critical role in the reproduction of numerous cultivated plants. Moreover, the study of their life history has charms all its own, for they have evolved fascinating forms of social interaction.

Bees are classified into six families, of which two, the *Halictidae* and the *Apidae*, have evolved social species. In addition to the honeybee, there are plasterer bees, mining bees or burrowing bees, mason bees, leaf-cutter bees, carpenter bees, and bumblebees.

At present, four species of *true honeybees* (tribe *Apini*) are known, all of them native to tropical southeastern Asia. But the so-called domesticated honeybee (*Apis mellifera*) has been distributed by humans all over the world. Before cane and beet sugar came into use, honey was the primary sweetener for foods. For that reason, the bees were most highly prized for their honey and wax. But, in more recent years, they have become even more prized for their beneficial activity in pollinating flowering crops.

Social Structure of Honeybee Colonies

All honeybees (see Fig. 1) build their combs of pure wax, which can be produced only by the workers. The combs are hung up vertically either in the open or in enclosed areas, and on both sides they have geometrically perfect hexagonal cells, a shape which minimizes the construction material required. The cells for the worker brood and those for the storage of honey and pollen are similar, while those for the males (drones) are larger. The queens are raised in special vertically hanging chambers, as shown by Fig. 2. Only the workers have the organs and instincts necessary for the activities of construction and foraging. The queen, somewhat larger than the workers, has as her only task the laying of eggs. Fertilized eggs give rise to workers or queens, depending upon the food given to the larvae, while the drones come from unfertilized eggs. The time at which reproductive individuals are to be produced is determined by the workers, who then prepare the appropriate cells and food. All the larvae are fed and cared for continually during their development. A queen can lay as many as 2000 eggs in a single day. In her 4 to 5 years of life, she produces about 2 million eggs. More than 80,000 bees can live in a colony.

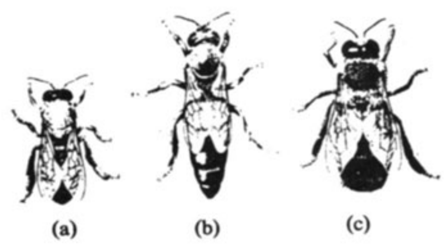


Fig. 1. Relative sizes of honeybees: (a) worker, (b) queen, (c) drone. (*USDA*.)

Fig. 2. Section of a honeybee comb. Drone cells are above and two queen cells are shown at the lower edge.

The tasks of the workers are manifold. They supply the colony with food, guard the nest, and build the combs. They also keep the combs clean, for these are used several times for the brood. By fanning with their wings, the workers cool the nest, and by their muscular activity they warm it. Thus, they ensure that the temperature in the brood area stays close to 35°C (95°F). Although every worker bee is, should special circumstances demand it, capable of performing any of these tasks, ordinarily there is an orderly division of labor, corresponding to the age of the workers. In their first days as workers, they act as janitors, keeping the combs clean. Next, after the pharyngeal gland in the head (See Fig. 3) has matured, they devote themselves to the larvae, feeding them first with the secretion of this gland and later with pollen and honey as well.

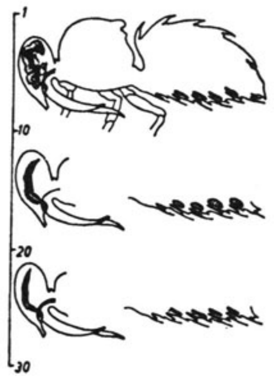


Fig. 3. Physiological changes having behavioral effects are correlated here with the age of the honeybee worker, the numbers at the left indicating the age of the worker. To the right of the ordinate is shown the stage of development of both the pharnygeal gland in the head and the wax glands in the abdomen.

The food given by worker honeybees to the young larvae during the first 3 days of their existence and to the larvae of queens until they are fully developed is known as *royal jelly*. This is a thick, white liquid formed in the stomach of the worker by partial digestion of honey and pollen and is apparently a highly-concentrated food. Queen cells are supplied with the material in excess of the needs of the larvae. If conditions within the colony deprive queen larvae of this abundance, they fail to become large and, in some cases, they may revert to development as intermediate forms between queen and worker. Such individuals may, however, have the instincts of queens and so may mate and lay fertile eggs. The change from royal jelly to a less concentrated food in the case of worker larvae apparently is responsible for the development of worker bees, since both queens and workers may develop from identical eggs.

At about the tenth day, the workers fly out briefly for the first time and become acquainted with the surroundings of the hive. In the following days, the pharyngeal gland becomes reduced and the wax glands begin to function. Now the worker bee becomes a construction worker, in addition to having responsibility for the food stores and carrying away refuse. Finally, when the worker approaches the age of 20 days, it goes out into the open more and more often. At first, it takes over guard duty at the flight hole; then it forages with great industry for pollen and nectar until the end of its life, in summer, after 4 to 5 weeks of labor.

Bee colonies propagate by swarms. In the early summer, shortly before one or more queens emerge from the queen cells, the old queen leaves the hive with about half its population. She first gathers her court into a cluster near the hive and then follows the advance guard, which has found a new nest site. The young queens which have matured in the old hive fly out repeatedly until they have mated, each with several drones. The supply of sperm thus accumulated must last the queen a lifetime. They mate in flight, at assembly places where the drones often congregate in great numbers. There the drones fling themselves on every female which flies past and has the appropriate scent signal. The inseminated queen returns to the nest and stings to death any rivals which may still be present. The drones too meet their fate in late summer. The workers drive them out of the nest with bites and stings, after which they are left to starve.

Although, traditionally, honeybees have been admired for their altruism, researchers at the University of California have shown that they project an aspect of allegiance to their queen and, in reality, the social structure is more that of a police state. What appears as selfless cooperation is enforced by a special platoon of enforcer worker bees. One researcher has observed that coercion may underlie the cooperation in honeybees as well as with other highly organized insects, such as ants, and that this is consistent with modern evolutionary theory.

Chemical signals are involved not only in mating. They also play a role in many aspects of communications by bees. Each queen secretes a substance by which the workers are continually reassured that the colony is not without a mother. Scent signals spread the alarm when the bees are in danger and also serve to identify food sources and the entrance to the hive.

Foraging for Nectar

Bees store up honey collected from various sources. Blossom honey is the thickened nectar of millions of flowers, which has passed through the stomachs of many bees and has been altered by glandular secretions. Bees also collect the sweet secretions of aphids which stick to leaves. With this, they form the leaf honey or pine honey, which is considered a delicacy. Several hundred kinds of plants produce nectar, which the bees also use for honey, but only a few kinds are common enough, or produce enough nectar, to be considered as major sources. The best sources of nectar for producing surplus honey vary from place to place. Some plants that are major nectar sources in the United States include: Alfalfa, aster, buckwheat, catclaw, citrus fruit, clover, cotton, fireweed, goldenrod, holly, horsemint, locust, mesquite, palmetto, tulip tree, tupelo, sage, sourwood, star thistle, sweet clover, sumac, and willow. The varying qualities of these sources is reflected in the color and flavor of the raw honey taken from the hive.

Since a good colony can store as much as 1 kilogram (2.2 pounds) of honey per day, the number of foraging flights undertaken in a day is astronomical. This efficiency depends upon two special achievements: (1) The bees can orient themselves; and (2) they can communicate with one another. Considering orientation, each bee must cover the distance between hive and collecting place as quickly and accurately as possible. This phenomenon was studied by Karl von Frisch and his coworkers for over 50 years. It was found that the bee uses all its senses—color leads it to the bright flowers; sense of smell enables it to distinguish different species of flowers. On the flight out and back, it not only observes conspicuous landmarks, but also uses the sun as a compass by keeping the flight path at the proper angles to the direction of the sun. Even the fact that the sun seemingly moves across the sky does not confuse the bee. Its sense of time permits it to take the time of day into account in its flights, correctly altering the setting of the sun compass. This sense of time also makes it possible for the bee to go at any time of the day to the sort of flower that is producing nectar then. Even if the sun is covered by clouds, the forager is not helpless. The direction of polarization of the light waves which penetrate from a patch of blue sky is dependent upon the position of the sun. Since the eyes of the honeybee, unlike human eyes, can measure this direction of polarization, the bee can orient to that just as well as to the sun itself.

Honeybee Communications

Because of their highly organized social structure, it is no surprise that the systematics of procedure should extend from the hive to their activities in the field when the bees are seeking nectar. Once a good source of nectar is found, this is effectively conveyed to other bees of the colony. Over many decades, naturalists have postulated the existence of such communications, but the mechanism escaped understanding. It was established that scout bees communicate to worker bees, not only in the hive, but also while in flight.

For a number of years, it was widely believed that the bees used aerial "dances" to convey their findings, but in recent years that theory has been challenged vigorously by another school of researchers.

Aerial Dance Theory. For what is considered one of the most exciting revelations of biology, Karl von Frisch received a 1973 Nobel Prize for describing how various kinds of aerial dances used by bees are a primary means of bee-to-bee communications. For example, he described waggle dances and round dances, as indicated very schematically in Fig. 4. Details are beyond the scope of this article. Although these performances of honeybees appear to border on the miraculous, they have been observed and documented by Dr. von Frisch and the supporters of his concept.

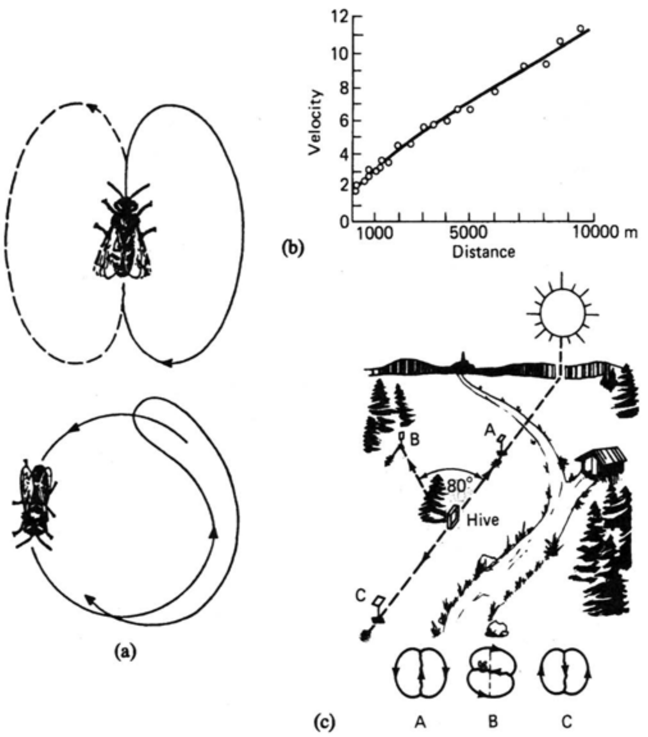


Fig. 4(a). Dances with which honeybees communicate. The "waggle" dance (top) is performed by long-distance foragers to communicate location of food sources. Both direction and distance are communicated. Direction is illustrated in (c). Distance is related to the speed or frequency with which individual dances are repeated. The round dance, indicated by the lower diagram, announces to the bees in the hive that a rich harvest has been located. Foragers repeat this dance each time they return to the hive until a more important new source is located. It is important to realize that the dance is normally done in the pitch darkness of a closed beehive, so that the dance is invisible to its comrades, who depend upon their senses of touch (vibrations) and smell to follow the tortuous course of the dance.

(b) Velocity of the waggle dance (ordinate) as a function of the distance (abscissa) between feeding place and hive.

(c) Relationship between flight angle with respect to sun and dance angle in the hive with respect to gravity. The latter is diagrammed for the three feeding sites shown (A, B, and C).

For example, Moffett (see reference listed) reported in 1990 findings based upon the use of an electronic robot bee in Germany. Construction of the electronic bee research robot is elegantly diagrammed in the Moffett reference. Inasmuch as the beehive is completely dark, the bees cannot depend upon visual perception. It was believed for many years that honeybees were incapable of detecting sounds. It now appears that vibrations from the bee robot are picked up by the honeybee's antennae. Hence, in essence, the researchers can "talk" to the bees by simulating the varying wing vibrations, which carry the information that the insect requires to locate a foraging target. Although many details remain unclear, investigators observe, "The angle between the dance direction and the vertical is known to signal the direction from the hive to food in relation to the sun. The bee waggles her abdomen and quivers her wings, indicating the distance. The intensity of the dance plus samples offered and the lingering odors on the bee's body suggest the type of food and its quality."

Also, having built another bee robot, researchers at Occidental College and the University of California (Santa Barbara), plus additional investigators at Princeton University, among others, refute the "aerial dance" theory. To date, they have been unable to gather evidence in support of the theory and conclude that honeybees depend largely on their sense of odor when searching for nectar. Thus, in the community of biologists, as of 1993, there remains a dichotomy of opinion.

Cognitive Maps. Research into honeybee behavior continues apace with major emphasis on bee navigation and flower recognition. Particular concern involves the so-called invertebrate-vertebrate dichotomy of navigation. As reported by J. L. Gould (Princeton University), the higher vertebrates use landmarks as a part of an overall map of an area so that a selection of routes (in contrast with a series of route-specific steps) may be used. In other words, the higher vertebrates have considerable flexibility in their navigation. Traditionally, it has been assumed that invertebrates were limited to navigating step by step from one specific landmark to the next. Recent extensive experiments by Gould tend to invalidate former assumptions and indicate that bees can store (in some fashion not fully understood) broad, maplike information, thus enhancing their navigational capabilities. In his research, Gould used individually marked honeybees (*Apis mellifera ligustica*).

Flower Recognition. In another series of studies, Gould reported that bees are able to learn to distinguish between flowers with different shapes or patterns. Earlier studies suggested that bees remember only isolated features, such as spatial frequency and line angles, rather than the photographic search images that are characteristic of vertebrates. New information indicates that bees can store flower patterns as a low-resolution eidetic image or photograph. Several decades ago, M. Hertz had suggested that bees spontaneously prefer highly dissected patterns (shapes with a high ratio of edge to area, or high spatial frequency) and that only the crudest sort of learned discrimination was possible. It was later found, however, that *shape learning* could be considerably more subtle, i.e., the memory process could be more sophisticated and complex than previously thought, but that nevertheless no photograph-like (eidetic) images were involved. Thus, two distinct hypotheses developed—the *isolated-feature or parameter hypothesis* and the *eidetic image or picture hypothesis*. Although not fully resolved, recent research tends to favor the latter hypothesis. Gould has suggested that the limiting factor is the resolution of the eidetic storage in the brain of the bee.

Other factors enter into how the bee discriminates pollen sources. For example, bees prefer to land on and learn to recognize most quickly violet-colored food sources. This, apparently, is an innate bias which must be accommodated by the insect's information storage system.

Interesting investigations also have been conducted among other foraging insects. A.C. Lewis (University of Colorado) has studied the memory constraints and flower choice in the cabbage butterfly (*Pieris rapae*). Decades ago, Darwin hypothesized that *flower constancy* in insects that feed on nectar results from the need to learn how to extract nectar from a flower of a given species. In experiments conducted by Lewis, it was found that the cabbage butterfly shows a flower constancy by continuing to visit flower species with which it had experience. The time required by individuals to find the source of nectar in flowers decreased with successive attempts, the performance following a learning curve. Learning to extract nectar from a second species interfered with the ability to extract nectar from the first. Insects that switch species thus experience a cost in time to learn.

Honeybee studies generally have involved the domestic honeybee (*Apis mellifera*), whose many subspecies have been distributed by human intervention throughout the world.

Other Honeybee Subspecies. The Indian subspecies (*A. mellifera derana*) closely parallels the domestic honeybee in structure and habits. Subspecies *A. mellifera dorsata* and *A. mellifera florea*, both of which live in southeastern Asia, represent a somewhat lower level of performance. Each of these species builds only one comb and hangs it in the open on a branch. These subspecies are not important commercially in terms of honey production.

"Africanized" Bees

The common honeybee of Europe and North and South America commonly is considered of Italian (European) origin. Although they are comparatively poor honey producers, they are known for their good temperament. In 1956, in an effort to breed a better honey producer, Brazilian entomologists imported 46 South African queen bees, which now are known to be of a much different temperament in terms of their interaction with society in general and thus developed the term "killer bee." Unfortunately, 26 queens escaped from the Brazilian laboratory and since that time have been spreading at an asymptotic pace northward into Central and North America.

Vigorous preventive measures were taken by Mexican officials and the U.S. Department of Agriculture. Migration of the "Africanized" bees was slowed, but not stopped. The first arrival of the unwanted bees was reported from a location near Hidalgo, Texas, in October 1990.

Entomologists predict that the bees ultimately will spread throughout the southern states of the United States and most likely will penetrate as far north as Pennsylvania in the east, Iowa in the midwest, and northern California in the west. It is believed that this rate of progression will depend largely on on the pattern of winter temperatures experienced as the bees travel northward. Responsibility for monitoring the progress and for taking preventive and planning measures to protect the American honey-producing and crop-pollinating industries rests with the U.S. Dept. of Agriculture (USDA) Laboratory located in Baton Rouge, Louisiana.

Hybridization of the African with the established European honeybees is unclear and difficult to forecast. Today, 80 percent of honey production in Argentina is credited to the "Africanized" bees, including hybridized bees.

Even though the Africanized honeybee has caused more deaths to humans than have the traditional European types, it is reported that the sting of the Africanized bee is no more serious than that from the European bee. What is different is the aggressiveness during swarming and the attendant "attacks" on people and animals in areas and situations where precautions normally would not be required, as, for example, in urban areas. A major concern is the probability that African bees may take over hives in the South and will affect traveling beekeepers in controlling their bees. Traveling beekeepers transport their hives around the countryside to pollinate an estimated $10 billion worth of crops each year.

Because of hybridization that is occurring in some of the transition zones as the African bees progress northward, the threat may be lessened. But, as one agricultural expert observes, this may require 25 years to develop.

Studies in honeybee genetics have been spurred in an effort to develop more productive countermeasures than those currently known.

Additional Reading

Cowen, R.: "Bumblebee Energy," *Science News*, 215 (October 6, 1990).
Darwin, C.: "The Effects of Cross- and Self-Fertilization in the Animal Kingdom," classic reference, Murray, London, 1876.
Doherty, J.: "The Hobby that Challenges You to Think Like a Bee," *Smithsonian*, 62 (July 1987).
Dyer, F. C., and J. L. Gould: "Honey Bee Orientation: A Backup System for Cloudy Days," *Science*, **214**, 1041–1042 (1981).
Gould, J. L.: *The Biology of Learning* (P. Marler and H. Terrace, Eds.), Springer-Verlag, Berlin, 1984.
Gould, J. L.: "How Bees Remember Flower Shapes," *Science*, **227**, 1492–1494 (1985).
Gould, J. L.: "The Locale Map of Honey Bees: Do Insects Have Cognitive Maps?" *Science*, **232**, 861–863 (1986).
Heinrich, B.: "The Regulation of Temperature in the Honeybee Swarm," *Sci. Amer.*, **244**(6), 146–160 (June 1981).
Horgan, J.: "Do Bees Think?" *Sci. Amer.*, 36 (May 1989).
Horgan, J.: "Disco-Bee (Robot)," *Sci. Amer.*, 31 (June 1989).
Horgan, J.: "Bee Police," *Sci. Amer.*, 28 (March 1990).
Horgan, J.: "Stinging Criticism," *Sci. Amer.*, 29 (November 1990).
Lewis, A. C.: "Memory Constraints and Flower Choice in *Pieris rapae*," *Science*, **232**, 863–865 (1986).
Moffett, M. W.: "Dance of the Electronic Bee," *Nat'l Geographic*, 134 (January 1990).
Needham, G. R., et al., Editors: "Africanized Honey Bees and Bee Mites," Wiley, New York, 1988.
Real, L. A.: "Animal Choice Behavior and the Evolution of Cognitive Architecture," *Science*, 980 (August 30, 1991).
Rinderer, T. E., et al.: "Hybridization Between European and Africanized Honey Bees in the Neotropical Yucatan Peninsula," *Science*, 3039 (July 19, 1991).
Robinson, G. E., et al.: "Hormonal and Genetic Control of Behavior Integration in Honey Bee Colonies," *Science*, 109 (October 6, 1989).
Rouibik, D. W.: "Ecology and Natural History of Tropical Bees," Cambridge Univ. Press, New York, 1989.
Spivak, M., Gletcher, D. J. C., and D. Michael: "The African Honey Bee," Westview, Boulder, Colorado, 1991.
Staff: "Bee Mites Buzz Off," *Nat'l. Food Review*, 43, U.S. Dept. of Agriculture, Washington, D.C. (October–December 1989).
Staff: "Myth of the Killer Bees," *Technology Review (MIT)*, 80 (February/March 1991).
Staff: "Africanized Bees Reach U.S.; Prepare to Settle," *Nat'l. Geographic, Geographica* section (April 1991).
Weiss, R.: "The African Advantage," *Science News*, 328 (May 26, 1990).

HONEY BUZZARD. See **Eagle.**

HONEYDEW. A sweet secretion produced by aphids, which, when abundant on trees, sometimes cause spotting on the leaves and anything below the tree like a heavy dew. Honeydew is freely sought by ants and is sometimes gathered by bees, but it makes a very inferior honey.

HOODOO. In geology, a columnar or pillar-like erosional remnant which has been carved and sculpted from relatively horizontal formations by differential erosion. The form and subsidiary features of hoodoos may be partly governed by joint planes and the differential hardness of the stratified sediments. The term applies particularly to eccentric and peculiar forms which are especially noticeable because of their fancied resemblance to animals and artifacts.

HOOKWORM. See **Dermatitis and Dermatosis.**

HOOPOE. See **Kingfishers and Other Coraciiformes.**

HOPHORNBEAM TREES. See **Hornbeam Trees.**

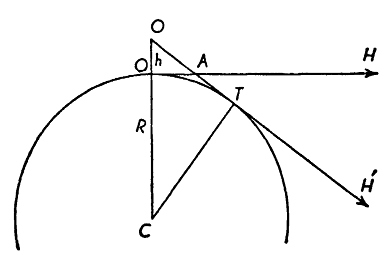
HOPS. See **Mulberry Family.**

HORIZON (Astronomical). Also called sensible horizon; real horizon. The plane that passes through the observer's eye and is perpendicular to the zenith at that point; or, the intersection of that plane with the celestial sphere (i.e., a great circle on the celestial sphere equidistant from the observer's zenith and nadir). It is the projection of a horizontal plane in every direction from the point of orientation. See also **Celestial Sphere** and **Astronomical Triangle.**

HORIZON (Celestial). Also called rational horizon; geometrical horizon; true horizon. The plane, through the center of the earth, perpendicular to a radius of the earth that passes through the point of observation on the earth's surface; or, the intersection of that plane with the celestial sphere.

In astronomy, the term horizon is used to describe the great circle cut out on the celestial sphere by a plane perpendicular to the direction of gravity. If this plane is tangent to the surface of the earth, the horizon so described is the *apparent horizon*; if the plane passes through the center of the earth, we have the *geocentric horizon*.

The difference in direction between the visible horizon (as the line where earth and sky meet) and the apparent astronomical horizon is known as the dip of the horizon. In the figure, $O'H'$ represents the direction of the visible horizon from an observer at a station O' elevated above the surface of the earth by an amount h. OH represents the direction of the horizon for the observer on the surface of the earth at O. The angle HAH′ is the dip of the horizon, and may be given, very approxi-



Visible and astronomical horizons.

mately, by this relation: the dip of the horizon (expressed in minutes of arc) is equal to the square root of the height of the observer above the surface of the earth (expressed in feet). The distance $O'T$ from the observer to the visible horizon is approximately as follows: the distance of the visible horizon (expressed in miles) is given by the square root of $\frac{3}{2}$ the height of the observer above the surface of the earth (expressed in feet). This distance is frequently very much increased by an effect known as looming of the horizon, produced by refraction of light in heated (or cooled) layers of the air near the surface. Both the expressions for the dip and distance of the horizon are applicable only when the point of observation of the visible horizon (the point T) is actually on the surface of the earth. See also **Celestial Sphere** and **Astronomical Triangle.**

HORIZON (Geographic). Also called apparent horizon, local horizon, and visible horizon, the distant line along which earth and sky appear to meet. In both popular usage and weather observing, this is the usual conception of horizon. Nearby prominences are said to obscure the horizon and are not considered to be a part of it. For observational reference, the minimum desirable horizon distance should be of the order of three miles.

Sea-level horizon, also called ideal horizon, sensible horizon, sea horizon, visible horizon, and apparent horizon, is the apparent junction of the sky and the sea-level surface of the earth; the horizon that is actually observed at sea. This type of horizon is used as the reference for establishing times of sunrise and sunset.

In these definitions of horizon, the zenith is considered to be at right angles to the horizon.

HORIZONTAL COORDINATE SYSTEM (Astronomy). A system of spherical coordinates on the celestial sphere, which uses the horizon as a fundamental plane. Planes perpendicular to the horizon cut out great circles on the celestial sphere known as vertical circles. The fundamental direction selected in the fundamental plane is true south. The azimuth of a point on the celestial sphere is the angular distance, measured in the plane of the horizon, from the true south direction to the point of intersection of the vertical circle through the object with the horizon. There are several different methods for expressing azimuth, but the astronomical method is to measure azimuth from the south through the west through 360°. The altitude of a point on the celestial sphere is the angular distance, measured along the vertical circle through the point, from the plane of the horizon to the point.

The horizontal system of spherical coordinates is frequently referred to as the altazimuth system.

See also **Celestial Sphere** and **Astronomical Triangle.**

HORMONES. During the past few years, the classical concept of hormone has been undergoing revision and expansion. A half-century ago, when only a few hormones were reasonably well understood, it was generally believed that there were a comparatively few very important complex biological substances generated by a few glands that stimulate and inhibit principal body functions. Investigators have learned, particularly since the early 1980s, that the creation of hormones is not an exclusive process for the endocrine glands, that receptors for any given hormone are not usually simply concentrated in relatively limited locations of the body, but are found sometimes where they were least suspected. For example, a few years ago when J. Roth and J. Kova (National Institutes of Health) confirmed insulin receptors in the human brain, they sought and found receptors in the testes and liver. In the case of insulin, this had led investigators to explore the possibility of insulin synthesis occurring, not just in the pancreas, but elsewhere in the body.

Researchers have also commenced studies of other life forms, insects and annelida for example, in a search for hormone receptors and, indeed, have found material similar to insulin and receptors that are reminiscent of human insulin receptors. That hormone production is not limited to the endocrine glands has been suspected for a number of years, but such cases were considered the exception rather than the rule. Modern investigators no longer accept this hypothesis. Suspected earlier, but not well confirmed until the mid-1980s, the human heart, once simply considered as a pump, is now known to create at least one hormone, *atrial natriuretic factor* (ANF). A few years earlier, researchers found that the human brain synthesizes important substances, such as endorphins and enkephalins.

Thus, the study and indeed the definition of hormones is departing rapidly from the former exclusive association with endocrinology. The field is becoming broader and more complex and the number of previously unidentified substances playing some form of hormonal role is increasing. More hormones are being found and in more locations in the living process, both human and other life forms. The field no longer is bracketed by a comparatively exclusive few sophisticated substances.

In connection with a new hypothesis for hormones, Roth has suggested that cell hormones and neurotransmitters began as what cell biologists term *tissue factor*—substances that stimulate cells to grow or come together or otherwise react biochemically. Only when animals evolved to have extreme cell differentiation and cellular organization did glands evolve to overproduce these hormones so the animals could use them in more clever and sophisticated ways. This theory would explain why many mammalian hormones are also tissue factors. As examples, insulin and glucagon, in addition to playing roles as hormones, also act locally as tissue factors on cells within the pancreas. Exocrine and endocrine functions overlap—there is no difference between exocrine and endocrine functions at the level of unicellular organisms. Roth points out that such a hypothesis would explain the finding that many classical messenger molecules, such as prostaglandins, nerve growth factor, and the hormonal substances are found in exocrine fluids such as saliva, intestinal secretions, milk, and semen.

Investigations are also being conducted apace in what might be called hormone genetics. It has been found, for example, that guinea pigs produce two different insulins, one type made in the pancreas, the other type synthesized in the brain and other organs. Thus, there are differences in gene expression. New findings may explain why cancer cells sometimes secrete hormones that cause severe metabolic disturbances. Lung cancers, for example, are prone to produce vasopressin, a cause of water retention. Perhaps the tumor-generated vasopressin is not normal vasopressin, but a slightly different hormone that has escaped detection by radioimmunoassays. There may be numerous other instances where investigators are seeking one of the better understood hormones, when a somewhat different, uncataloged hormone should be the target. Admittedly, the thoughtful speculation involved in the reevaluation of hormone science will require considerably more research and proof.

In studies of the immune system, investigators have reported on the finding of a heretofore unknown immunoregulatory hormone, 1,25-dihydroxyvitamin D_3. The substance was found to be effective in suppressing interleukin-2.

By applying recombinant DNA techniques, a group of researchers has produced two human fertility hormones, human chorionic gonadotropin (hCG) and human luteinizing hormone (hLH). This is one of the first examples in which recombinant DNA techniques have been used to produce molecules that are a combination of proteins and carbohydrates in mammalian cells. The two hormones are similarly structured, consisting of two polypeptide chains that are put together inside cells and processed. It has been suggested that the hormones will be useful in the treatment of infertility because they can induce both ovulation and sperm production. Past hormone treatments have involved extracts from pituitaries, urine, or placentas which do not yield a pure product.

Hormone Science in the Traditional Sense

In animals, hormones are organic compounds, usually of considerable complexity and even after years of research not fully understood, that are secreted by endocrine (ductless) glands, such as the adrenal gland, the thyroid and parathyroid glands, the pituitary gland, and the gonads, among others. See **Endocrine System.** Hormones are sometimes commonly called by the names of the gland which secrete them. Thus, there are adrenal cortical hormones, thyroid and parathyroid hormones, etc. Hormones are regulators of physiological processes within the body, exerting control over such processes as metabolism, growth, reproduction, molting, pigmentation, and electrolytic and osmotic balance, among other processes. Apparently, hormones achieve these objectives chemically and electrically, although the mechanisms are not fully understood and, in fact, the mechanisms may vary from one situation to the next. At one time, hormones were loosely called "chemical messengers" because they are transported from point to point within the organism and thus effect actions at distances from the region where they are made. If one visualizes secreting glands as sensors of a type detecting need for correction of some physiological process, then the hormones might be visualized as both the transmitters or carriers of this information and the initiators of actions as well. The conventional concept is that cells have receptors on their surface which sense the presence of specific hormones. At one time, it was firmly believed that hormones, particularly polypeptide hormones, such as insulin, prolactin, and growth hormone, all of which are large charged molecules, could not penetrate through the cell's membrane and actually enter the cell. This belief has since been altered because researchers have shown that insulin, for example, can enter into the cell. Referring to this process as "internalization," one investigator in 1978 suggested that the internalization of polypeptide hormones will be one of the most active topics in cell biology for a number of years.

Research has shown that hormones and/or their receptors may be degraded. As gross examples of this type of situation, it is known that many obese people with high concentrations of insulin in their blood also have normal concentrations of blood sugar. Why doesn't the insulin decrease the blood sugar concentration in these cases? It is well known that pregnant women produce much angiotensin II, which normally increases blood pressure, but these women usually do not have hypertension. What alterations in the hormone-cell mechanism provide this result? There are also instances where males have tumors which secrete large quantities of a hormone that stimulates the production of testosterone, and yet there is no evidence of abnormal amounts of testosterone. At least two questions can be posed: Do certain hormones lose their effectiveness with time? Or, are there changes in target receptor cells? Research has indicated that there may be a relationship between concentration of hormones and the surface receptors which bind them, such receptors being inactive for a time or possibly disappearing from the cell surface altogether. A number of investigators have observed that a better understanding of the manner in which hormones affect their own and other receptors possibly may result in new ways to treat certain diseases, including insulin-resistant diabetes.

In recent years, it has been shown that a wide variety of receptors are regulated by hormones. Some receptors are sensitive to only one hormone; this appears to be the case with insulin. Others appear to be regulated not only by one hormone, but others as well. For example, it has been shown that the receptors for TRH (thyrotropin-releasing hormone) are not exclusively regulated by TRH, but also by other hormones. Receptors for gonadotropins (pituitary hormones that act on the gonads) appear to be regulated by hormones in addition to gonadotropin. There are numerous other instances of this kind.

Hormones display not only great variations in function, but also in their chemical nature, of which there is a great diversity. Some are steroids, such as estrogen, progesterone, cortisone, etc., while others are amino acids (thyroxine), polypeptides (vasopressin), low-molecular-weight proteins, and conjugated proteins. Amino acid and steroid hormones have been isolated and many, including insulin, have been synthesized. Other types are prepared directly from the endocrine organs of animals.

Hormones produced by one species usually show similar activity in other species. The hormones showing greatest species specificity are proteins or conjugated proteins.

Hormones are markedly affected by deficiencies or excesses of the various vitamins and other dietary essentials.

Because of the great complexity of a number of the natural hormones, conventional approaches of organic synthesis which have been used so successfully over the years in connection with many drugs have not proved viable to date with some of the hormones. Insulin is an example. Presently, millions of diabetics still depend upon animal insulin as extracted from the pancreatic glands of slaughtered pigs. If diabetes mellitus continues to become more prevalent, as it has over the past several years, natural sources may not be sufficient.

Recombinant DNA technology was applied to the problem the first time a few decades ago. One group has been successful in inducing the bacterium *Escherichia coli* to manufacture and secrete rat proinsulin, an immediate precursor of rat insulin that incorporates insulin itself. Research like this is an important step toward the objective of developing bacterium-based industrial systems that can replace animal and human tissues as the source of medically useful proteins, such as insulin, growth hormone, and clotting factor.

Classes of Hormones

Hormones may be grouped into two distinct types: (1) *direct-acting*; and (2) *stimulating*-substances that stimulate other organs to produce their own characteristic hormones. The latter group is sometimes called the *tropic hormones*. See accompanying tabular summary of hormones.

Thyroid Hormones. These are compounds of the amino acid *thyronine*. They are present in the free form only to a slight extent, existing chiefly as constituents of the protein thyroglobulin. The most important of these acids in terms of hormone action are the 3,5,3′-tri, and the 3,5,3′,5′-tetraiodocompounds, *triiodothyronine* and *thyroxin*, the structures of which are given in the accompanying table. See also **Thyroid Gland.** The action of thyroid hormones is to accelerate cellular reactions and to increase the metabolic rate and oxygen consumption of tissues. They effect this action by stimulating many of the enzyme systems, not only the glucose oxidation system and the cytochrome chain for dehydrogenating the coenzyme NADPH, but other processes, such as the synthesis of proteins from amino acids. Their effects are clearly apparent in the pathological changes in the organism caused by their excess or deficiency. The thyrotrophic hormone and other biochemical interactions with the thyroid gland are discussed later in this entry.

Parathyroid Hormones. The influence of the parathyroid glands on the regulation of calcium concentrations in the blood of mammals was first recognized by MacCullum and Voegtlin in 1909.

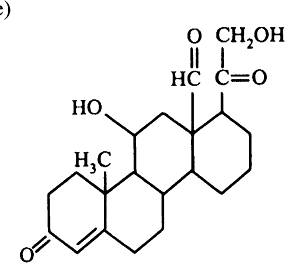
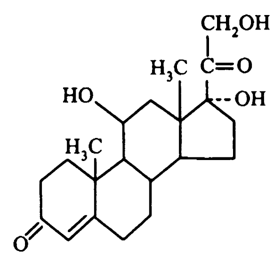
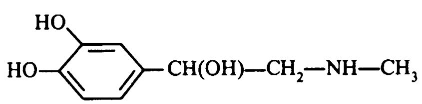
More recently, several groups of investigators have succeeded in purifying and partially identifying the structure of the hormone, variously called *parathormone* and *parathyroid hormone*. This is a single chain peptide hormone with a molecular weight of about 8,000. A second parathyroid hormone, *calcitonin*, was postulated by Copp (1961). Subsequent research has indicated that this hormone is actually the hormone which is now known to be produced by the thyroid gland. However, a parathyroid calcitonin may exist in certain species.

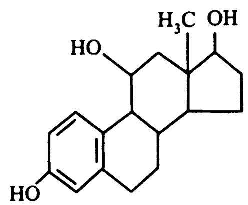
The more classical function of parathyroid hormone is concerned with its control of the maintenance of constant circulating calcium levels. Its action is on (1) the kidney, where it increases the phosphate in the urine, (2) the skeletal system, where it causes calcium resorption from bone, and (3) the digestive system, where it accelerates (stimulates) calcium absorption into the blood. The hormone and gland exhibit characteristics of feedback control; when the concentration of calcium ions in the blood falls, the secretion of the hormone increases, and when their concentration rises, the secretion of hormone decreases. See **Parathyroid Glands.**

Adrenal Cortical Hormones. The adrenal gland is made up of two parts, the medulla and the cortex, each of which secretes characteristic hormones. The hormones of the adrenal medulla are the catecholamines, epinephrine (adrenalin) and norepinephrine (noradrenalin), which are closely related chemically, differing only in that epinephrine has an added methyl group. See accompanying table. In fact, animal experiments have established a metabolic pathway for the biosynthesis of both compounds from the amino acid phenylalanine, which involves enzymatic oxidation and decarboxylation reactions. It is also to be noted that the isomeric form of norepinephrine is most important; the natural D-form (which incidentally, is levorotatory) has many times the activity of the synthetic isomer. Epinephrine has a pronounced action upon the circulatory system, increasing both blood pressure and pulse rate, and hence the cardiac output by its direct action upon the heart muscle, and especially because it causes constriction of the arterioles. However, its effects upon smooth muscles vary; it relaxes the muscles of the digestive system, but contracts the pyloric sphincter.

Norepinephrine does not affect the cardiac output, although it does raise the blood pressure by constricting the arterioles. Its muscular effects are less pronounced. Both epinephrine and norepinephrine release free fatty acids from adipose tissue, so raising its level in the blood. This effect is due to the action of the hormones in accelerating enzymatic reactions whereby the esters of the fatty acids are hydrolyzed. The third type of action of epinephrine is its effect upon the carbohydrate metabolism, notably the acceleration of the hydrolysis of glycogen in muscular tissue and the liver, and so raising the glucose level in the blood, and the rate of glucose oxidation, with resulting increase in oxygen

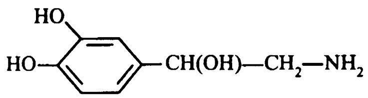
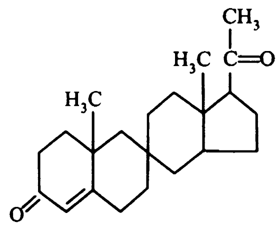
REPRESENTATIVE HUMAN HORMONES

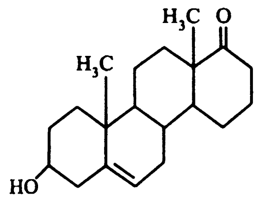
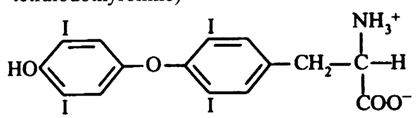
Hormone Common Names, (Synonyms), Structure and Production Site	Principal Physiologic Functions	Interrelationships with Vitamins
Adrenocorticotropic Hormone (ACTH) (Adrenocorticotropin; corticotropic hormone) Straight-chain, simple polypeptide, 39 amino acids, no S—S bridges. (See text, Fig. 1.) Molecular Weight ~4500 Production Site: Anterior pituitary	Maintenance of adrenal cortex Promotes secretion of steroids, oxidative phosphorylation in adrenal cortex Mobilizes and increases oxidation of free fatty acids in adipose tissue Increases gluconeogenesis in liver; increases cyclic adenosine monophosphate (AMP) in adrenal cortex Decreases urea formation in liver	Ascorbic acid: depleted in adrenal cortex on stimulation by ACTH Biotin and vitamin A: adrenocortical insufficiency noted in biotin and vitamin A deficiency Niacin: production of reduced nicotinamide adenine dinucleotide (phosphate) (NADPH) by ACTH via cyclic adenosine monophosphate (AMP) Niacin and pantothenic acid: synergistic with ACTH in steroid hormone synthesis Vitamin D: antagonized directly by ACTH via cortisol action
Aldosterone (Aldocortin; electrocortin; mineralocorticoid; 18-oxo-corticosterone)  Molecular Weight 360.4 Production Site: Adrenal cortex	Maintenance of normal electrolyte blood balances Prolongs survival of adrenalectomized animals Accelerates gluconeogenesis Regulates kidney function	Ascorbic acid: adrenal cortex depleted of ascorbic acid on production of aldosterone Biotin: prolongs life in adrenalectomized rats Niacin: nicotinamide adenine dinucleotide (phosphate) (NADPH) involved in synthesis of aldosterone
Cortisol (Hydrocortisone, 17-hydroxycorticosterone)  Molecular Weight 362.5 Production Site: Adrenal cortex	Increases (1) protein catabolism (excepting liver) gluconeogenesis; (2) carbohydrate anabolism (liver); (3) blood sugar; (4) glucose absorption; (5) brain excitation; (6) spread of infections; (7) urinary glucose and nitrogen; (8) stress tolerance; (9) lactation; (10) water diuresis Decreases (1) fat anabolism; (2) growth rate; (3) inflammation; (4) eosinophils; (5) lymphocytes; (6) antigen sensitivity; (7) respiratory quotient; (8) ketosis; (9) wound healing; (10) skin pigmentation; (11) RBC hemolysis. Regulates general adaptation syndrome, water balance, blood pressure, and hormone release.	Ascorbic acid: may be required for steroid hormone biosynthesis; depleted from adrenal cortex on cortical secretion Biotin: adrenocortical insufficiency noted in biotin deficiency Folic acid and pantothenic acids maintain secretions of steroids by adrenal cortex Niacin: nicotinamide adenine dinucleotide (phosphate) (NADPH) required for steroid hormone biosynthesis Vitamin A: deficiency causes cortical necrosis Vitamin D: action antagonized by cortisol by reducing calcium absorption in intestine
Epinephrine (Adrenaline, adrenin, suprarenin, vasotonin, vasoconstrictine, adrenamine, levorenine)  Molecular Weight 183.2 Production Site: Adrenal medulla and chromaffin cells in gut	Blood circulation: increases blood pressure; peripheral vasodilator; increases heart output and rate; flow increased in brain, liver, and skeletal muscle Central nervous system: causes restlessness, anxiety Kidney: reduces glomerular filtration rate Lung, intestine, genital system: inhibited motility Metabolic effects: increases oxygen consumption, temperature, basal metabolic rate, gluconeogenesis Pituitary effects: stimulates production and release of ACTH and corticoids	Ascorbic acid: maintains reduced state of epinephrine Ascorbic acid, folic acid, and vitamins B_6, and B_{12} are cofactors in synthesis of epinephrine from phenylalanine

Estradiol (Female hormone; dihydrotheelin; dihydrofollicular hormone dihydrofolliculin)  Molecular Weight 272.4 Production Sites: Ovarian follicles; testes; corpus luteum; adrenal cortex; placenta	Regulates menstrual cycle, female sex behavior Maintains secondary sex characteristics Affects antibody properties Induces estrus, uterine hypertrophy, vaginal cornification; potentiate sand stimulates calcitonin secretion	Folic acid: involved in mitotic effect of estradiol Niacin, diphosphopyridine nucleotide (DPN), triphosphopyridine nucleotide (TPN): involved in increased respiration and in cholesterol precursor synthesis Pyridoxine: competes as cofactor with estrogen sulfate in kynurenine aminotransferase activity Vitamin D: synergistic in calcium metabolism with estradiol Vitamin E: involved in follotropin production or release
Follicle-Stimulating Hormone (FSH) (Follotropin, luteoantine, thylakentrin, Prolan A, gonadotropin1, gametogenic hormone, follicle ripening hormone, gametokinetic hormone) Structure: Not fully definitized. Production Site: Anterior pituitary.	Female: stimulates ovarian follicles to grow and to develop, forming multiple layers and antra Male: stimulates seminferous tubules; stimulates spermatogenesis	Ascorbic acid: depletion in ovary due to follicle-stimulating hormone and luteinizing hormone action Vitamin E: required to maintenance of membranes in sex organs
Glucagon (HGF) (Hyperglycemic-glycogenolytic factor; glukagon; HG-factor) Structure: Polypeptide, 29 amino acids (structure determined). No S—S bridges. Molecular Weight ~3500 Production Site: Alpha cells in pancreas.	Increases: blood sugar; blood K^+, oxygen consumption, liver glycogenolysis, gluconeogenesis, nitrogen and salt excretion Decreases: liver glycogen, protein formation, gastric juice, fatty acid synthesis	Ascorbic acid: depletion of adrenal ascorbic acid by glucagon
Insulin (no synonyms) Structure: 51 amino acids. Known and synthesized. 3 S—S bridges. (See text, Fig. 4) Molecular Weight 5,734 (monomer); 12,000–48,000 (polymer), depending upon pH. Production Site: Beta cells of islets of pancreas.	Regulates carbohydrate and fat metabolism, especially glucose and fat oxidations Stimulates amino acid and glucose transport into cells and protein synthesis	Ascorbic acid: acts similarly to alloxan (i.e., antagonist)
Luteinizing Hormone (LH) (Luteotropin, ISCH) Structure: Globular glycoprotein with S—S bridges. Molecular Weight 26,000 Production Site: Anterior pituitary.	Female: promotes estrogen and progesterone secretion, ovulation; maintains ovarian tissues Male: stimulates Leydig cells to secrete testosterone; gametogenic with follotropin (FSH)	Ascrobic acid: ovarian depletion on LH stimulation Vitamin E: involved in spermatogenesis
Melanocyte-stimulating Hormone (MSH) (Melanotropin, chromatophorotropic hormone; pigmentation hormone) Structure: Polypeptide; purified, synthesized; alpha and beta forms; straight chains. Molecular Weight: 1500 (alpha) 2100–2600 (beta) Production Site: Intermediate lobe of pituitary.	Mammals: exerts small effect on skin pigmentation (protection from sunlight not fully proved) Expands or contracts pigments in various chromatophores Expands melanophore pigments with color changes in amphibia (adaptation to environment) Lower vertebrates: increases sensitivity to light; decreases dark adaptation time	Ascorbic acid: adrenal cortex depleted on ACTH and MSH activity Vitamin A: MSH decreases dark adaptation time.

(continued)

REPRESENTATIVE HUMAN HORMONES *(continued)*

Hormone Common Names, (Synonyms), Structure and Production Site	Principal Physiologic Functions	Interrelationships with Vitamins
Norepinephrine (Arterenol; noradrenaline; levarterenol)  Molecular Weight 169.2 Production Site: Adrenal medulla; adrenergic nerve endings; chromaffin cells.	Blood circulation: increases blood pressure; peripheral vasoconstrictor without change or slight decrease in output and heart rate. No flow increase in brain, liver, or muscle Central nervous system effects: adrenergic transmitter agent at synapses; no brain excitation Kidney: decreases glomerular filtration rate Lung, intestine, genital system: inhibited Metabolic effects: weak epinephrine effect	Ascorbic acid: protects against oxidation of norepinephrine Ascorbic acid, folic acid, and vitamin B_6 are cofactors in synthesis of norepinephrine from phenylalanine
Oxytocin (Oxytocic hormone; pitocin; uteracon; α-hypophamine) Cys—Tyr—Ile—Gln—Asn—Cys—Pro—Leu—Gly NH_2 (Cys–Cys bridge) Molecular Weight 1007 Production Site: Hypothalamus.	Uterine contraction, milk ejection, facilitates sperm ascent in female tract Decreases membrane potential of myometrium, basic metabolic rate, and liver glycogen Stimulates oviposition in hen, releases luteinizing hormone (LH) Increases blood sugar and urinary sodium and potassium	Findings on interrelationships with vitamins are not extensive
Parathyroid Hormone (PTH) (Parathormone) Structure: Simple polypeptide (83 amino acids), sequence determined; straight chain; No S—S bridges. Production Site: Parathyroid glands.	Increases blood calcium, kidney calcium reabsorption, phosphate excretion, and blood citrate level Mobilizes calcium and phosphate from bone Activates calcium and phosphate absorption from the gastrointestinal tract (for which vitamin D is required) Increases osteoclast formation	Vitamin D: synergistic with PTH in maintenance of serum calcium
Progesterone (Progestin, luteosterone)  Molecular Weight 314.5 Production Sites: Ovary (follicles, corpus luteum); testicles; adrenal cortex; placenta	In low concentrations: prepares uterus for blastocyst implantation; promotes ovulation and mammary gland development; regulates female sex accessory organs; weak corticosteroid properties; precursor to sex hormones In high concentrations: maintains pregnancy; represses ovulation and sex activity; inhibits vaginal cornification and parturition; decreases myometrial excitation	Ascorbic acid: depleted from adrenal cortex or ovary on progesterone formation Niacin: diphosphopyridine nucleotide (DPN) involved in progesterone synthesis
Prolactin LTH (Lactogenic hormone; lactogen; galactin; mammotropin) Structure: Single-chain protein, 205 amino acids Molecular Weight 23,000–25,000 Production Site: Anterior pituitary	Initiates lactation Develops mammary glands in female Increases weight and growth (similar to somatotrophin in some species) Participates in nidation of zygote Protein anabolism (some species) Growth and secretion of crop gland (birds) Luteotropic (only in mouse, rat) Promotes maternal behavior	Not fully determined. Generally participates with other substances having growth action
Relaxin (Releasin, cervilaxin) Structure: Polypeptide (4 peptides with activity have been isolated); about 30–40 amino acids in each peptide	Enlarges birth canal in preparation for parturition Separation of symphysis pubis, loss of rigidity in pelvic bones Decreases uterine motility Maintains pregnancy	Ascorbic acid: maintains mucoprotein ground substance in connective tissue, affected by relaxin

Molecular Weight 4000–5000 Production Site: Corpus luteum in pregnancy	Increases sensitivity to oxytocin; releases oxytocin Stimulates mammary gland Stimulates inhibition of water in uterus Inhibits uterine contraction	
Somatotropin (STH) (Growth hormone, GH; somatotrophic hormone; hypophyseal growth hormone) Structure: Known and synthesized; coiled, unbranched protein; 188 amino acid residues; 2 S—S bridges Molecular Weight 21,500 Production Site: Anterior pituitary	Promotes general growth of organism Promotes skeletal growth, protein anabolism, fat metabolism, carbohydrate metabolism, water, and salt metabolism	Relates with all vitamins in connection with growth actions
Testosterone (17 beta-hydroxy-4-androsten-3-one)  Molecular Weight 288.4 Production Sites: Interstitial cells of ovary and testis; adrenal cortex; embryonic placenta	Controls secondary male sex characteristics Maintains functional competence of male reproductive ducts and glands Increases protein anabolism; maintains spermatogenesis; inhibits follotropin Increases male sex behavior; increases closure of epiphyseal plates	Ascorbic acid, folic acid, vitamins A and E are synergists with testosterone for maturation of germ cells and increased anabolic activity
Thyroid-stimulating Hormone (TSH) (Thyrotrophic hormone, thyrotropin) Structure: Glycoprotein (300 amino acids) Molecular Weight 26,000–30,000 Production Site: S^2 type cell, anterior pituitary	Regulates body temperature via thyroxine Maintains thyroid gland and its secretory activity (colloid discharge) Maintains iodine uptake by thyroid gland Promotes differentiation in embryo during development via thyroxine Stimulates coupling of diodotyrosine to form thyroxine	Ascorbic acid, thiamine, riboflavin, and vitamin B_{12}: requirements increase in hyperthyroidism; tissue concentrations reduced Vitamin A: massive doses of vitamin A inhibit secretion of TSH; thyroid hormones required for carotene and retimene conversions Vitamins A, D, E, and K: requirements increased in hyperthyroidism; tissue concentrations reduced in Vitamin B_6, niacin: conversion to phosphorylated reactive forms impaired in hyperthyroidism
Thyroxine (T_4) (3,5,3',5' tetraiodothyronine)  Molecular Weight 776.9 Production Site: Thyroid gland	Regulates growth, differentiation, oxidative metabolism, electrolytic balance Increases carbohydrate metabolism, calorigenesis, protein anabolism, basal metabolic rate, oxygen consumption, fat catabolism, fertility Sensitizes nervous system	Ascorbic acid: synergist in cold survival Niacin: synergist in mitochondrial metabolism Vitamin A: T_4 is required for vitamin A synthesis in liver Vitamin B_{12}: T_4 aids in B_{12} absorption B complex vitamins: deficiencies develop in hyperthyroidism
Vasopressin (Arginine vasopressin; antidiuretic hormone; ADH; pitressin; tonephin; vasophysin) Cys—Tyr—Phe—Glu NH_2—Asp NH_2—Cys—Pro—Leu—Gly NH_2 **Vasopressin** Molecular Weight 1084 (arginine-vasopressin) Production Site: Hypothalamus	Elevates blood pressure (mammals) (reverse effect in birds) Decreases kidney blood flow Antidiuretic, releases ACTH Increases sodium chloride and urea excretion Regulates water balance Stimulates contraction of smooth muscles Increases renal tubular water reabsorption Releases anterior pituitary hormones	Not fully determined

SOURCE: Adapted from R. J. Kutsky.

utilization, carbon dioxide production, and body temperature. See **Adrenal Glands.**

The hormones of the adrenal cortex are steroids. See also **Steroid.** Among them there are a number of hormones with androgenic activity, such as adrenosterone and 17α-hydroxyprogesterone, which are discussed under the sex hormones later in this entry. In all, over ten steroids have been identified in the adrenal cortex, including seven of characteristic cortical activity. These are corticosterone, from which the others are named, 17α-hydroxyl-11-dehydrocorticosterone (cortisone), 17α-hydroxycorticosterone (cortisol or hydrocortisone), and 18-oxo-corticosterone (aldosterone). Only two hormones, cortisol and corticosterone, are normally released in fairly large quantities, and another, aldosterone, deserves mention because of its somewhat different effects, even though it is released to a far lesser extent.

All of these hormones are synthesized from cholesterol in the adrenal cortex, by an extended series of reactions which include many related compounds. Although these hormones have widespread effects throughout the organism, their primary mechanism is not known, so that many of the effects may be indirect. Much of the knowledge of their action arises from studies of insufficiency or hyperactivity of the adrenal cortex, which produces a wide variety of pathological conditions. See accompanying table.

It is generally considered that aldosterone, and to some extent the other hormones, have a regulatory effect upon the metabolism of electrolytes and water, particularly upon the concentration of the ions of the alkali metals in intracellular fluids. Administration of steroids also increases the concentration of calcium ions in those fluids. However, all three of these hormones have a number of other effects, roughly in the order of potency—cortisol, corticosterone, aldosterone. They produce changes in the metabolism of carbohydrates, proteins, and fats.

For the carbohydrates alone, three major effects are evident—increase in the rate of formation of glucose, increase in the rate of release of glucose from the liver, and increase in the rate of utilization of glucose. These hormones affect the digestive system, increasing the secretion of hydrochloric acid, pepsinogen, and trypsinogen. They prevent inflammatory responses to bacterial or even chemical stimuli; they counteract anaphylactic shock, and other effects of hypersensitivity. Obviously, these properties have led to their widespread therapeutic use.

There are relationships between the adrenal cortical hormones and the thyroid and pituitary glands. Depression of the function of the adrenals produces thyroid deficiency, whereas administration of thyroxine stimulates the ACTH-adrenal cortical mechanism.

Pituitary Hormones. The hormones of the hypophysis (pituitary gland) are quite numerous, being secreted variously in three parts of the gland—the neurohypophysis (posterior lobe), the adenohypophysis (anterior lobe), and the *pars intermedia*, which connects the other two.

The chief hormones of the neurohypophysis are the polypeptides oxytocin and vasopressin. The hormone characteristic of the *pars intermedia* is the melanocyte-stimulating hormone. It is usually spoken of in the plural, since in most mammals both alpha and beta forms are known. The structures of the first two are shown in the accompanying table. See also **Diabetes Insipidus.**

The most prominent effect of oxytocin is the contraction of smooth muscle, especially of the uterus. It also has a major effect upon the muscles about the breast, and so stimulates the ejection of milk in lactating animals. It has a definite stimulating effect upon the muscles of the ureter, urinary bladder, intestine, and gall bladder.

The most prominent effect of vasopressin is upon the kidneys, where it stimulates the resorption of water in the tubules (which by repeated release and absorption concentrate the urine). It also constricts the coronary arteries, raises the blood pressure, and exhibits the effect of oxytocin upon smooth muscles, but generally to a lesser degree.

The action of the melanocyte-stimulating hormones has been established by studies of animals, in which they cause dispersal of certain black pigments from the cells that contain them, with resulting darkening of the skin.

The adenohypophysis is the part of the gland in which the tropic hormones are secreted. They include the adrenocorticotropic hormone (ACTH), the thyrotropic hormone (TSH), and somatotropin, as well as three hormones with pronounced effects upon the gonads: the hormone prolactin, the follicle-stimulating hormone (FSH) and the luteinizing or interstitial cell stimulating hormone (LH or ISCH).

ACTH. Adrenocorticotropin (ACTH) in humans is a polypeptide containing a sequence of 39 amino acids, although work with animal forms of it and with degradation products of the human form have shown that not all of them are essential to the activity of the hormone. This sequence for the human ACTH is shown in Fig. 1.

Ser—Tyr—Ser—Met—Glu—His—Phe—Arg—Tyr—Gly—Lys—Pro—
Val—Gly—Lys—Lys—Arg—Arg—Pro—Val—Lys—Val—Tyr—Pro—
NH_2
|
Asp—Ala—Gly—Glu—Asp—Glu—Ser—Ala—Glu—Ala—Phe—Pro—
Leu—Glu—Phe

Fig. 1. Amino acid sequence of human adrenocorticotropin (ACTH). Abbreviations of amino acids will be found in entry on **Amino Acids**.

The primary function of ACTH is the stimulation of the adrenal cortex to produce its hormones, which have already been discussed. This is evident from the therapeutic effect of administration of ACTH, which is closely similar to that of these hormones, so that if the action of only one of them is sought, its administration is preferable. Moreover, ACTH stimulates secretion of the androgenic substances mentioned as produced by the adrenal cortex.

Thyrotropic Hormone. This hormone (TSH) stimulates the development of the thyroid and controls its secretion. Although purified preparations of it have been obtained, they consist of a mixture of proteins of high mean molecular weight (about 30,000). Some of their amino acids have been determined, as well as their carbohydrates, but the structures have not been elucidated.

Growth Hormone. Somatotropin is the growth hormone. Purified preparations of extracts of it from the human adenohypophysis have been crystalized. They are known to be proteins, of mean molecular weight 21,000, and containing a single polypeptide chain. This hormone differs from the others of its group in not acting primarily upon the other endocrine glands, but in controlling the gain in body weight and the rate of skeletal growth. The growth abnormalities, such as dwarfism and giantism, have been shown to result from its hypo- and hypersecretion. In addition to its effect upon growth and anabolism generally, it has been found to affect the kidneys and pancreas, and to influence glucose, galactose, and lipid metabolism. See also **Pituitary Gland.**

Gonadotropic Hormones. These include follicle stimulating hormones (FSH), luteinizing or interstitial cell stimulating hormone (LH or ISCH), and prolactin. Their structures are not known; the molecular weight of human LH is about 26,000, that of human FSH is about 30,000, and that of human prolactin is uncertain. They are proteins, with variable amounts of carbohydrates. FSH induces the growth of Graafian follicles in the ovary and the production of spermatozoa in the testis. LH stimulates the final development of the ovarian follicles, the appearance of estrus, and the change of the follicles to corpora lutea. In the male, it stimulates the secretion of testosterone. Since these effects are due to the effect of this hormone upon interstitial cells, it is also called ISCH. Prolactin stimulates lactation after birth, acts with estrogen to promote the growth of the mammary gland, and influences the activity of the corpora lutea.

Male Hormones. The androgenic hormones produced in the testes (and adrenal gland) have a widespread effect upon the development of secondary sexual characteristics (musculature, facial hair, larynx, etc.), as well as upon the sexual organs and responses themselves. They also promote anabolism to a marked degree by their effect upon nitrogen and calcium metabolism. The structure of testosterone is shown in the accompanying table. See **Gonads.**

Female Hormones. Closely related to the male androgenic hormones, and probably synthesized from them in the female organism, are the estrogenic hormones which are produced principally in the ovary. Although β-estradiol is the normally secreted ovarian hormone, a number of other estrogenic substances have been isolated from urine and from animal studies. They include α-estradiol, estriol, and estrone. The structures of these hormones are given in Fig. 2.

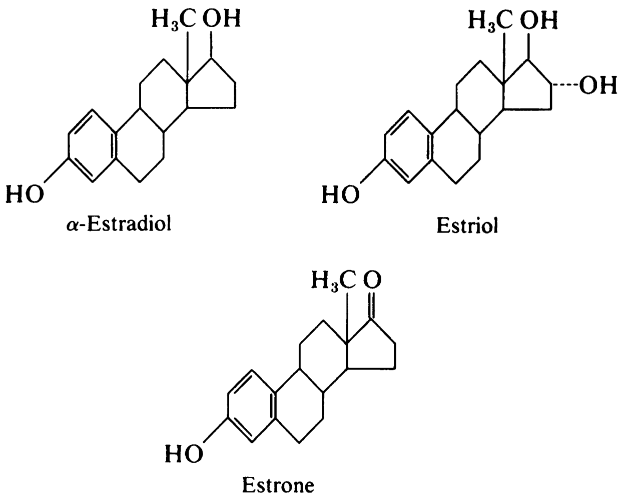


Fig. 2. Major ovarian hormones.

These hormones are important in both the menstrual cycle and the reproductive cycle, and of course play an important role in oral contraceptives (the "pill"). They induce growth of the vaginal epithelium, secretion of mucus by the glands of the cervix, and initiate the growth of the endometrium, which is taken over by progesterone (from the corpus luteum) later in the cycle. They activate the proliferation of the mammary gland during pregnancy. As the androgens do for the male, the estrogens bring about the secondary sexual characteristics of the female. They have a number of effects upon metabolism, notably that of calcium and phosphorus, and of lipids and proteins. A number of other estrogens, some made synthetically and others obtained from animals, are known.

The corpus luteum produces two hormones, progesterone and relaxin. The structures of these hormones are shown in the accompanying table.

Progesterone acts to complete the proliferation of the endometrium, which was initiated by the estrogenic hormones, and to prepare it for the ovum. In pregnancy the continued action of progesterone is necessary. It aids the growth of the breasts and has a definite effect against ovulation. It is also the biosynthetic precursor of some of the estrogenic hormones. Relaxin has been shown to have a relaxing effect on the cartilaginous junction of the public bones in preparation for parturition. See also **Embryo.**

Feedback in Hormone Control Systems

Not only do the hormones initiate or stimulate biological processes, both directly and by bringing about production of other hormones in other glands, but they also act to maintain the organism in a steady state, or *homeostasis.* Thus the gonadotropic hormones from the hypophysis stimulate the testes, but the resulting production there of androgens like testosterone, inhibits the action of the hypophysis in producing the gonadotropic hormones. The complicated cycle of adjustment in the human female is shown in the cycle illustrated in Fig. 3.

As shown in the figure, the regulation of the ovarian hormones in the human female involves both positive and negative feedback. The follicle-stimulating hormone (FSH) from the adenohypophysis stimulates the Graafian follicles, which thus produce estrogens. These not only inhibit FSH production through negative feedback, but also stimulate

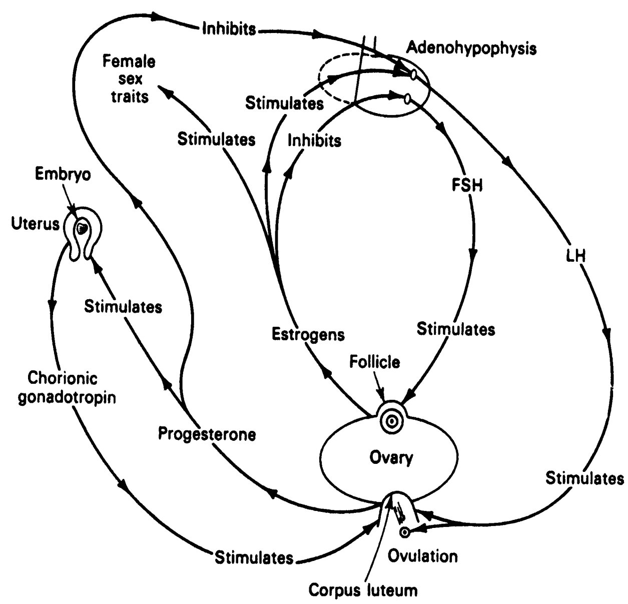


Fig. 3. Cycle of hormone adjustment in human female.

the adenohypophysis to increase its production of luteinizing hormone (LH) through positive feedback. This hormone in turn brings about ovulation from the Graafian follicle. After the ova are discharged, the LH stimulates the empty follicle, now the corpus luteum, to produce progesterone.

This hormone brings about the changes in the reproductive organs required for the development of the embryo. Then the progesterone partly inhibits the adenohypophysis from producing further LH, an example of negative feedback; as a result, there is no further ovulation. The progesterone also acts as a positive feedback and stimulates the production of FSH.

When pregnancy intervenes, a new feedback mechanism must be introduced, or the embryo would be expelled by the shedding of the lining of the uterus in menstruation. Here the placenta (chorion) of the embryo itself produces hormones, as already noted. Its LH stimulates continuing production of progesterone from the corpus luteum, thus preventing menstruation and stimulating the continuing development of the uterus as needed by the growing embryo. The extra progesterone also inhibits further ovulation in spite of the presence of the gonadotropin from the placenta (chorion).

Pancreas and Nonendocrine Hormone Sources. In addition to producing hormones, the pancreas also generates digestive fluids (*pancreatic juice*). It is the hormone function which makes the pancreas a part of the endocrine system. See **Endocrine System; Pancreas.** The pancreas secretes *insulin* and *glucagon,* both hormones. The structure of glucagon consists of a single chain of amino acids. See Figs. 4 and 5.

These two hormones have two opposing effects. That of insulin is *hypoglycemic,* i.e., it increases the rate of utilization of glucose, the probable process being an effect of insulin to increase the penetration of glucose through the cell walls as well as increased phosphorylation. The overall result of action of insulin in its relation to glucose is to increase the rate of the reactions by which glucose is oxidized, but also its transformation to glycogen. The enzyme glucagon raises blood glucose levels by increasing the rate of hydrolysis of glycogen (in the liver)

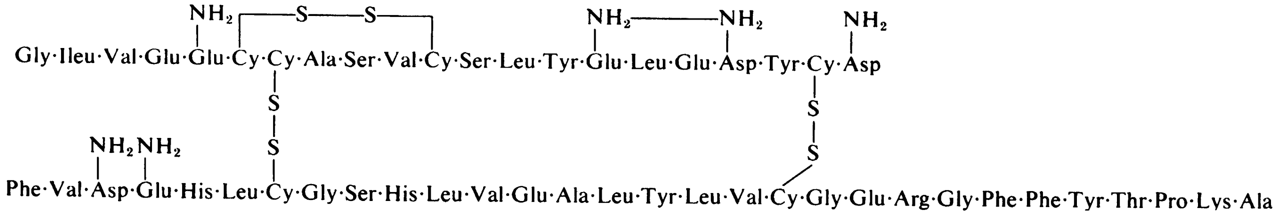


Fig. 4. Primary structure of bovine insulin.

His·Ser·Glu(NH_2)·Gly·Thr·Phe·Thr·Ser·Asp·Tyr·Ser·Lys·Tyr·Leu·Asp·Ser·Arg·Arg·Ala·Glu(NH_2)·Asp·Phe·Val·Glu(NH_2)·Tyr·Leu·Met·Asp(NH_2)·Thr

Fig. 5. Glucagon.

to increase the formation ultimately of glucose. Insulin increases the rate of entry of amino acids into cells and their rate of protein biosynthesis. Insulin also accelerates the formation of lipids from carbohydrates, whereas glucagon stimulates the formulation of keto compounds from lipids, inhibits the synthesis of fatty acids, and accelerates the breakdown of various phosphorus and nitrogen compounds. The primary result of insulin deficiency is diabetes mellitus. See **Diabetes Mellitus.**

Hormones may be produced by organs other than the endocrine glands. Conspicuous among such organs is the placenta, the organ on the wall of the uterus to which the umbilical cord is attached. It has been found to produce the same estrogenic hormones as the ovary, the same hormones (progesterone and relaxin) as does the corpus luteum, and gonadotropic hormones (and luteinizing hormones) similar to, but not identical with, those produced by the adenohypophysis.

Other hormones which do not originate in endocrine glands are the cholecystokinin of the intestine, and the enterogastrone and gastrin of the stomach. The first is produced by the upper intestinal mucosa and causes the gall bladder to contract; the enterogastrone is produced in the same tissue and inhibits gastric motility and secretion; it also excites secretion of digestive fluids, principally hydrochloric acid.

In addition to the entries covering specific endocrine glands, see also **Endocrine System; Nervous System and the Brain;** and **Steroids.**

In plants, a *plant hormone* or "phytohormone" is an organic compound produced by the plant, controlling growth and other functions at sites remote from where the hormone is produced. Plant hormones also act in very minute amounts. Plant hormones include the auxins, gibberellins, and kinetins. These are described in the entries on **Gibberellic Acid and Gibberellin Plant Growth Hormones;** and **Plant Growth Modification and Regulation.** Plant hormones are also mentioned in a number of specific plant-related entries.

Additional Reading

Conn, P. M., and W. F. Crowley, Jr.: "Gonadotropin-Releasing Hormone and Its Analogues," *N. Eng. J. Med.*, 93 (January 10, 1991).

Erickson, D.: "Human Growth Hormone," *Sci. Amer.*, 164 (September 1990).

Erickson, D.: "Hormone Derivatives May Combat PMS and Epilepsy," *Sci. Amer.*, 124 (May 1991).

Evans, R. M.: "The Steroid and Thyroid Hormone Receptor Superfamily," *Science*, 889 (May 13, 1988).

Fackelmann, K. A.: "High-Pressure Hormone," *Sci. News*, 344 (December 1, 1990).

Golde, D. W., and J. C. Gasson: "Hormones That Stimulate the Growth of Blood Cells," *Sci. Amer.*, 62 (July 1988).

Kravitz, E. A.: "Hormonal Control of Behavior: Amines and the Biasing of Behavioral Output in Lobsters," *Science*, 1775 (September 30, 1988).

Lerner, R. A., and A. Tramontano: "Catalytic Antibodies," *Sci. Amer.*, 58 (March 1988).

Marx, J.: "How Peptide Hormones Get Ready for Work," *Science*, 779 (May 10, 1991).

HORNBEAM TREES. Of the family *Carpinaceae* (hornbeam family), these trees are of the genus *Carpinus*, except the hophornbeams, which are of the genus *Ostrya*. These trees once were classified with the birches (family *Betulaceae*). There are both shrubs and trees within *Carpinus*, all deciduous and hardy. The common hornbeam of Europe is the *C. betulus*, which can attain a height up to 75 feet (22.5 meters). The tree is generally of a pyramidal contour. The flowers are unisexual. In terms of landscaping, one authority attributes much to what may be termed the interesting texture of the trunk and base, where there are innumerable muscle and tendon-like configurations. It is also of interest to note that the common European hornbeam has been used in formal gardens as a form of "hedge-on-stilts." There are several variations of the common European hornbeam, including the 'Columnaris,' the 'Fastigiata,' the 'Incisa,' and the 'Intertexta,' thus providing a range of size, coloration, and density.

The common hornbeam native to and common in North America is the *C. caroliniana*, also sometimes called the American hornbeam, the blue beech, or the water beech. This plant may be described as a tall shrub or small tree and ranges up to 30 or 40 feet (9 to 12 meters) in height. One excellent specimen selected by The American Forestry Association in 1966 is located in Canton, Ohio. The tree has a circumference at $4\frac{1}{2}$ feet (1.4 meters) of 118 in (300 cm), a height of 42 feet (12.8 meters), and a spread of 65 feet (19.8 meters). The current champion tree (1975) is located in Ulster County, New York. The tree has a circumference at $4\frac{1}{2}$ feet (1.4 meters) of 95 inches (241 cm), a height of 69 feet (21 meters), and a spread of 56 feet (17 meters).

The branches of the American hornbeam are slender and extended nearly horizontally from the trunk. The leaf is narrow, ovate, sharp-pointed and a dull light green, lighter color underneath. The staminate catkins are about $1\frac{1}{2}$ inches (3.8 centimeters) long. The bark is scaly and gray-brown. This tree ranges from Nova Scotia and Quebec west to the Great Lakes and south through Nebraska, Kansas, and Oklahoma, to Texas in the southwest and Florida in the southeast. The tree is quite common in New England.

The Japanese hornbeam (*C. japonica*) reaches a height of about 50 feet (15 meters) and is wide-spreading. It has male catkins from 1 to 2 inches (2.5 to 5 centimeters) in length. A more ornamental Japanese hornbeam is the *C. laxiflora*. Other species include the *C. orientalis*, a bushy shrub of southeastern Europe and Asia Minor; and the *C. turczaninowil*, a thin, spindly hornbeam.

Closely related to the hornbeams are the hophornbeams. See accompanying photo. Generally, these trees are medium-to-large in size and are deciduous. One excellent specimen selected by The American For-

Eastern hophorn beam tree (*Ostrya virginiana*.)

estry Association is the Eastern Hophornbeam (*Ostrya virginiana*) located in Grand Travis County, Michigan in 1976. The tree has a circumference at $4\frac{1}{2}$ feet (1.4 meters) of 115 inches (292 centimeters), a height of 74 feet (22.2 meters), and a spread of 111 feet (33.8 meters). They are also characterized by drooping male catkins and upright female catkins. Their fruit is in the form of a nutlet contained in a husk. The hophornbeam of southern Europe and Asia Minor is the *Ostrya carpinifolia*, which can attain a height up to 65 feet (19.5 meters). The common hophornbeam in America is the American hornbeam (*O. virginiana*), sometimes also called the ironwood or leverwood. Generally, this is a fairly small tree, ranging from 30 to 45 feet (9 to 13.5 meters) in height, but under favorable conditions the tree can do much better. The bark is gray-brown, scaly, and has perpendicular scoring. The leaves are from 3 to 4 inches (7.6 to 10 centimeters) in length, narrow, ovate, double-toothed, sharp-pointed, and of a dull light-green color. The staminate flowers normally occur in three drooping catkins. The tree prefers a dry soil and open woods. It ranges from Nova Scotia and New Brunswick southward along the Saint Lawrence and Lower Ottawa Rivers, westward to Lake Huron, and northwest to Minnesota and the Dakotas, thence southward as far as Kansas and Nebraska. In the east, it is found in the Alleghany Mountains and south into Florida. The tree also is found in eastern Texas. Thus, the tree has a broad climatic range.

The *Ostrya knowltoni* is found on the southern slopes of the Colorado River canyon in Arizona and northward to Flagstaff. It is abundant to the 6,000 to 7,000-foot (1800 to 2100-meter) level. The height of the tree ranges from 25 to 40 feet (7.5 to 12 meters). The trunk is about 15 inches (38 centimeters) in diameter. The bark is light brown and scaly, with bright orange underneath. The leaf is small, $1\frac{1}{2}$ to 2 inches (3.8 to 5 centimeters) long, ovate, soft, hairy above and smooth underneath.

The Catalina ironwood (*Lyonothamnus Gray*) or Lyon tree is of interest because it is so rare and because of its unusual location. The tree grows on the canyon slopes of the steep shores of Santa Catalina Island, just off the coast of southern California. The tree was discovered by William Lyon, a young forester, in 1884 and thus so named. It is postulated that the tree was on the island at the time when the sea level was many hundreds of feet lower than it is today. There are a number of species of the tree, particularly on the eastern side of the island. They are also found on Santa Rosa and Santa Cruz islands nearby. Geologists postulate that at one time these islands were a connected land mass. However, the tree leaf differs from one location to the next. The tree has compound foliage on Santa Rosa and Santa Cruz, whereas this is not the case on Santa Catalina.

If the trees were numerous, the wood would be of considerable economic value because it is very strong and tough, weighing about 50 pounds per cubic foot (801 kilograms per cubic meter). It is believed that the Canalino Indians once used the wood for handles and shaft wood. The trees grow straight and tall, varying greatly in size, some towering up to 60 feet (18 meters) in height, with trunks of about $1\frac{1}{2}$ feet (0.5 meter) in diameter. They are found at an elevation between 500 and 2,000 feet (150 and 600 meters). The flowers are small and lacy, in groups on branches that are somewhat flat. They bloom in June and July.

The leaf has a simple-bladed fern-like foliage, projecting an aura of antiquity. It is blade-shaped with teeth coarsely cut. The seed is oblong and light-brown. The bark is dark brown, but appears to weather to a lighter color. It is often tattered, with the underbark showing through. The tree is difficult to propagate from cuttings or seeds. Root sprouts thus far have proved most effective.

HORNBILL. See **Kingfishers and Other Coraciiformes.**

HORNBLENDE. The mineral hornblende is a complex silicate which is probably an isomorphous mixture of three molecules, a calcium-iron-magnesium silicate, an aluminum-iron-magnesium silicate and an iron-magnesium silicate. A general formula is

$$(Ca,Na,K)_{2-3}(Mg,Fe^{2+},Fe^{3+},Al)_5(Al,Si)_8O_{22}(OH)_2.$$

Manganese and alkalies are sometimes present as is also titanium. It is monoclinic, with prismatic crystals, often pseudo-hexagonal. Bladed, fibrous, columnar, granular and compact massive varieties also are common. It has a perfect prismatic cleavage; hardness, 5–6; specific gravity, 3.02–3.27; color, green, greenish-brown, brown and black; luster, vitreous to silky; transparent to opaque.

Hornblende is a common constituent of many of the igneous rocks such as granite, syenite, diorite, or gabbro, of gneisses and schists and is the principal mineral of the amphibolites. Hornblende alters easily to chlorite and epidote. A variety of hornblende that contains little (less than 5%) of iron oxides is gray to white in color and named edenite, from its locality in Edenville, New York. Very dark brown to black hornblendes, which contain titanium, ordinarily are called basaltic hornblende from the fact that they are usually a constituent of basalts and similar rocks.

Well-known localities for hornblende are in The Czech Republic and Slovakia, Mount Vesuvius, Italy, Norway, Sweden, and, in the United States, in Massachusetts, New Hampshire, and New York. Black hornblende is found in Renfrew County, Canada. The word hornblende is derived from the German *horn*, and *blende*, to blind or dazzle. The term blende was often used to refer to a brilliant nonmetallic luster, i.e., zinc blende.

See also terms listed under **Mineralogy.**

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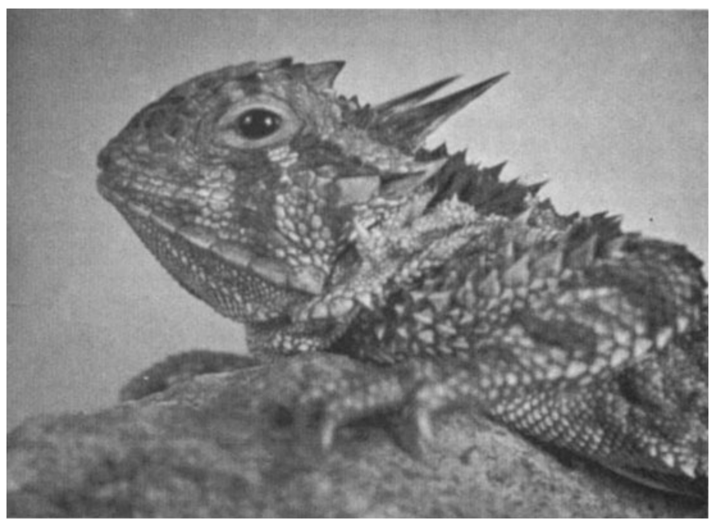
HORNBLENDITE. A coarse-grained rock related to gabbro which consists almost wholly of hornblende. Olivine being present, this rock may grade into a hornblende-peridotite (cortlandtite). Hornblendite is a rare rock type and of relatively little importance.

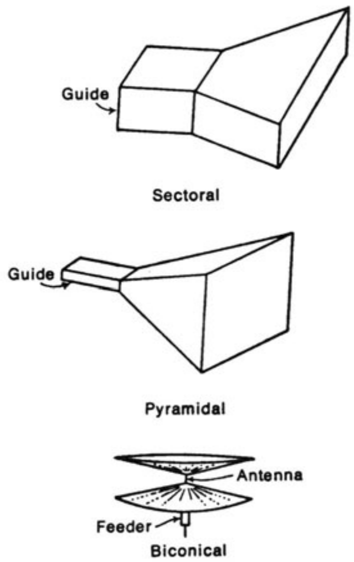
HORNED TOAD (*Reptilia, Sauria; Phrynosoma*). Small spiny lizards of the southwestern states and Mexico. They have short broad bodies and short tails, hence the confusion of terms in the common name. Horned lizard is a better term.



Horned toad. (*A. M. Winchester.*)

Horned toads are desert animals and are capable of living for incredibly long periods without food or water. They cannot, however, survive for the long periods of years as has sometimes been claimed.

HORN (Electromagnetic). Horn radiators are used to obtain directional radiation characteristics which cannot be obtained as conveniently with simple antennae. As such directors, they are used both with conventional antennae and with waveguides, but in either case they serve to direct the radiation in a pattern from the open end of the horn in a manner determined by the dimensions of the horn. The important dimensions are the horn opening (in terms of wavelength of the radiation) and the flare angle. While theoretically an infinitely long horn will give a radiation pattern whose angle conforms to that of the horn, those of practical length do not confine the beam to quite this degree. For



Types of electromagnetic horns.

example, a horn with an angle of 15° may give a radiation pattern which spreads 23°. The common horns may be divided into three classes, sectoral, pyramidal and biconical. The sectoral horn has two sides which are parallel and the other two flared. The pyramidal horn has all sides flared. The conical horn is really a pyramidal horn with a circular cross section. The biconical horn consists of two cones with their vertices coinciding or adjacent to one another. The first two types are used where a singly directed beam of radiation is desired, the exact pattern in both vertical and horizontal planes being determined by the dimensions. The biconical horn gives a uniform pattern in a plane perpendicular to the axis and highly directional in any plane containing the axis.

Sectoral and pyramidal horns may be excited by more or less conventional antennae or by waveguides. In the former case, a short section of waveguide is attached to the end of the horn and this is excited by the antennae. In the latter case, the horn is really a flared extension of the guide and may be looked upon as an impedance-transforming section for matching the impedance of the guide to that of free space. Biconical horns are excited by a variety of antenna arrangements in the space between vertices of the horns. Because of space considerations, horns are not feasible except at ultra-high frequencies, but in the microwave region they are widely used as radiators.

HORNET (*Insecta, Hymenoptera*). A name loosely applied to many of the larger wasps, particularly to the species which build paper nests and have a severe sting. See Fig. 1.

Fig. 1. Hornet, a member of the wasp family (*Vespidae*). (*USDA diagram.*)

The true hornet is a European wasp (*Vespa crabro*). In America, the term may refer to any form of large stinging wasp that makes paper nests. Hornets are all social. In the southern United States, a smaller species (*V. carolina*) goes by the name of hornet. The white-faced hornet is the common American hornet (*V. maculata*).

The hornet is usually yellow and black in coloration and has a pugnacious spirit. Its sting is severe. The nest is usually pear-shaped, but sometimes round, and suspends from a branch of a tree or roof of a building. The nest consists of horizontal cones all facing downward. A small hole is left in the side of the nest as an entrance. The nest may accommodate from a few hundred to over 5,000 hornets.

These insects live much as honeybees, with similar work habits. Their food is the nectar of flowers. Adult hornets prefer carbohydrate food sources; the young prefer protein foods from caterpillars. Hornets are susceptible to bacterial and fungus disease and have numerous insect enemies, factors which keep their numbers under control. Among their worst enemies are other hornets which pillage the food they store.

HORNFELS. A more or less general term applied to fine-grained, massive, and frequently speckled rock, the result of contact metamorphism developed in slates by granitic intrusions.

HORN FLY (*Insecta, Diptera*). This insect is most irritating and injurious to cattle, but also will attack goat, horse, and sheep. On occasions, the fly will also be a pest on dogs. The species *Haemotobia* or *Siphona irritans* (Linne) pierces the animal's skin and sucks blood. The associated pain and irritation trouble the animals during resting and feeding periods and, as a result, the cattle lose weight, milk cows have a lower milk yield, and the general health of the animals deteriorates, particularly when the irritation continues over a long period. The horn fly is about half the size of the common house fly, but appears very much like it. Although not fully proven, some scientists believe that the horn fly may carry anthrax disease.

The insect was first noted in the United States in Philadelphia in 1887, but has since spread throughout the continental United States and also to Hawaii.

Control is usually by spraying the animals along their back and flanks with one of several formulations, including methoxychlor, ronnel, or toxaphene. Treatment of dairy cows should be confined to hand-rubbing a suitable formulation around the neck area, thus avoiding possible contamination of milk. Pyrethrins and allethrin also have proved effective. Where practical, housing the animals in darkened structures equipped with entrance curtains that help to remove the flies can be helpful.

The fly maggots depend largely on animal dung for their food. Thus, cleanliness about shelters frequented by the animals cannot be overstressed.

HORN (Substance). A hard translucent material formed by the development of epidermal cells containing a substance known as keratin. The outer layers of the skin are keratinized and the nails, claws and hoofs of mammals are formed of similar material. Horn is also developed in large amounts in the appendages of the head which go by the same name. Horns may be bony cores sheathed in horn or solid bony growths. The former occur in cattle and the latter in deer. The median horn or horns borne on the head of rhinoceroses are quite unlike true horns. They are formed of aggregated hair-like components firmly based on roughened areas of the underlying bones.

HORN-TAIL (*Insecta, Hymenoptera*). Large sawflies (woodwasps) whose larvae bore in the trunks of trees. The adults have a cylindrical body and in the female sex a short strong ovipositor which is the source of the name horn-tail. With this organ holes are drilled into the wood of the tree for the deposition of the eggs.

HORNWORM (*Insecta, Lepidoptera*). Closely related species of this worm are known as tobacco worms and tomato hornworms. The adult moths do not injure plants, but their larvae are quite damaging.

Southern or tobacco hornworm (*Protoparce sexta*, Johanssen). See Fig. 1. This insect is found in most of the United States and ranges southward into South America.

Northern or tomato hornworm (*Protoparce quinquemaculata*, Haworth). See Fig. 2. This insect also is found throughout the United States and ranges northward into Canada.

Fig. 1. Tobacco hornworm moth. (*USDA*.)

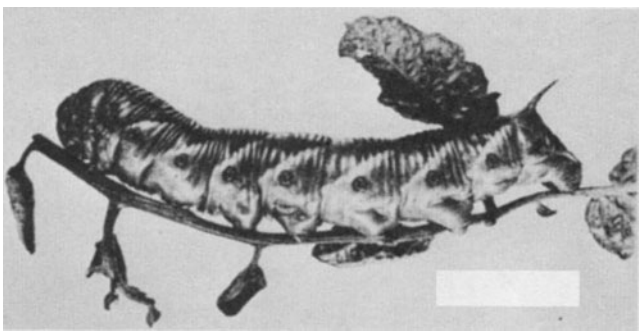


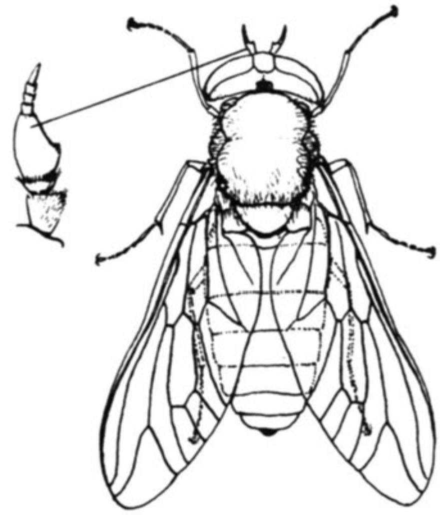
Fig. 2. Tomato Hornworm larva. (*USDA*.)

In habit and damage, the two species are strikingly similar and, in fact, both species may be found attacking the same crop. In addition to tobacco and tomato, the insect is injurious to eggplant, pepper, and tomato. The worms also feed on a number of weeds often associated with these crops.

The worms are green with diagonal lines on their sides. They have a prominent horn (red horn on tobacco hornworm; black horn on tomato hornworm) on the rear end. They range up to 4 inches (10 centimeters) in length. Damage is caused by their eating of foliage and fruit. While widely distributed, infestations are usually localized. Because the worms are so large, handpicking is comparatively easy. Control chemicals used are carbaryl, endosulfan, and toxaphene. A natural enemy is the braconid wasp *Apanteles congregatus* (Say).

HOROLOGIUM. A southern constellation situated near Eridanus.

HORSE FLY (*Insecta, Diptera*). These insects are of special species (*Tabanua* and *Chrysops* spp.) and are irritating and injurious to horses, mules, cattle, hogs, deer, and other wild animals. On occasion, they are pests to humans. The flies look something like bees and are of a tan to brown coloration. The wings are faintly spotted. During spring and summer, the adult flies severely irritate the aforementioned animals and hover closely about the head, neck, and forequarters of the host, waiting for an opportunity to strike and lay their eggs. The bite is painful because the mouth parts of the insect are very sharp. The fly will suck blood from the animal's neck or back for several minutes. The animal may twitch and run about in an unmanageable fashion. Some scientists have estimated that in regions where the horse fly is abundant, an animal in the field may lose as much as 3 ounces (about 90 grams) of blood per day to horse fly inflictions. Some species of horse fly carry such diseases as tularaemia, Calabar swellings (filariasis), and el debab, a disease of camels and horses that occurs in Algeria. Evidence also indicates that these flies are carriers of swamp fever, surra, and anthrax.



Horse fly showing detail of biting structure. (*USDA*.)

Control measures usually are directed against the maggots. The insect usually winters as a fully grown larva in some wet area near a stream. The maggot is pointed at both ends and is about 2 inches (5 centimeters) in length. They are fully grown by late spring, after which they pupate until early summer in dried mud. Adult flies commence to appear in early summer. Depending upon species, there are one or two generations per year.

Preventive measures, such as draining wet places and taking precautions similar to those for mosquito control, are effective. Spraying of the animals with allethrin, incorporating piperonyl butoxide, can be effective. Broadcasting a granular insecticide over areas frequented by animals also is effective.

HORSEPOWER. See **Units and Standards.**

HORSES, ASSES, AND ZEBRAS (*Mammalia, Perissodactyla*). With exception of Grevy's Zebra, all horses, asses, and zebras are placed in the genus *Equus* (Equines).

The horse is a hoofed animal with a single toe on each foot, encased in a massive hoof. The teeth are very high-crowned grinding structures. Two wild species of central Asia, the Tarpan and Prezewalski's horse (*Equus prezewalski*), are most closely related to the domestic horse and probably represent the original stock from which the domestic horse (*Equus caballus*) was derived. Although much of the developmental history of the horses is known from North American fossils, there is no evidence to indicate that horses were on the continent when it was explored for the first time by Europeans. The wild horses of the American West are feral descendents of the domesticated horses imported originally from Spain.

Prezewalski's horse is believed to be the only true ancestor of the domestic horses that are found in the wild today. This is a small, quite heavyset animal, reddish-brown in color. See Fig. 1. Many years ago, it commonly roamed the plains of Eurasia. As nomads settled on the steppes, their domesticated stock won out in competition for water and pasture. The last Prezewalski's horse to be captured occurred in 1947 and was moved to Askanya-Nova, a reserve in the Ukraine. The horse was named for a Russian colonel who first reported its existence in the 1870s. About a thousand of this species remain, all born in captivity in various zoos and wildlife parks. As of 1992, plans were underway to reintroduce the horse into the wild on its former Mongolian native ranges. It is interesting to note that Prezewalski's horse has a few characteristics that are unique to the species. The horse has a distinctively short mane and no forelock. Genetically, the horse has 66 chromosomes instead of 64, as found in domestic horses and feral horses, such as mustangs.

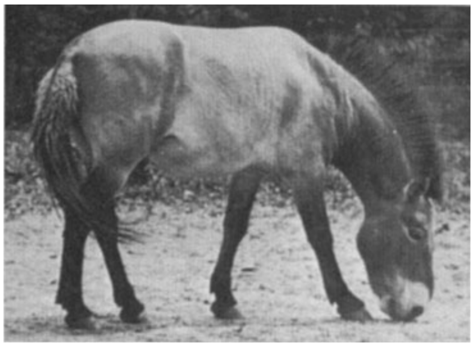


Fig. 1. Prezewalski's horse. (*A. M. Winchester.*)

The tarpan is a wild horse of the steppes of central Asia. This species has been regarded as feral rather than a natural species, but this interpretation has been disputed. It is in any case closely related to the domestic horse and may be an ancestral form. The western American mustang, a spirited, agile horse, is exemplary of interbreeding in the wild and is believed to be descended from stock introduced by Spanish explorers. The mustang also is known as the Indian pony or bronco. A worldwide classification of common horse and pony breeds is given in the accompanying table.

Classification of Horses. Under domestication, many varieties of horses have been developed for riding, driving, draft animals, and other uses. They also have been crossed with the domestic ass to produce mules for various uses, and have been hybridized experimentally with other species. Some authorities believe that the horse has no equal in its capacity and adaptability to withstand extreme climatic conditions and the various uses to which it has been placed by humans.

There are several ways to classify horses, as, for example, the draft horse (for working and pulling heavy loads); the light horse (light loading and riding); and the pony (a small horse, normally not over 14 hands high). A hand is considered four inches (10 centimeters) in breadth; thus $4 \times 14 = 56 = 4$ feet, 8 inches, or (1.42 meters). Horses also are commonly designated by color. Bay is considered the more or less standard color and this is reddish-brown. Other colors of horses include brown, gray, chestnut (a particular reddish-brown), and roan (various—reddish-brown with a significant sprinkling of gray or white). Pinto signifies marked with spots of two or more colors; Palomino designates a golden or brownish-gray horse with an ivory or silvery-white mane and tail.

There are numerous breeds of horses, often classified generally as light, coach, and heavy breeds. The coach breeds originated in England and include the Cleveland Bay, the Yorkshire Coach, and the Hackney. The latter is possibly the most popular and is $14\frac{1}{2}$ to $15\frac{1}{2}$ hands high, quite strong, with high carriage of the tail, and usually of a dark color. Heavy breeds include the Clydesdale, developed in Scotland and named for the Clyde River water. The height is 16 to $16\frac{1}{2}$ hands high; color is dark brown or bay; noted for its heavy fetlock (tuft of hair on back of leg just above hoof), and high action. The Belgian is another heavy breed, originated in Belgium, with a height of from 16 to 17 hands, weight from 1,800 to 2,200 pounds (816 to 997 kilograms); color chestnut or sorrel (particular reddish-brown shade); flaxen mane and tail. The Percheron is a famous draft breed that originated in Normandy. The head is small; the contour is Arabian; height is from 16 to 17 hands; weight from 1,900 to 2,100 pounds (861 to 953 kilograms); heavy, but active and supple: excellent for agricultural jobs.

Thoroughbred race horses were originally developed by crossing English with Turkish and Arabic horses. Many special racing and riding breeds have been developed, much too numerous for description here. Commonly known breeds include the Tennessee walking hoıse which has a distinctive gait—the running walk; the American saddle horse; the American quarter horse, particularly adapted for running short distances; and numerous others.

One breed of domesticated horse found in Germany and Austria is shown in Fig. 2. These horses are born totally raven black, but within a comparatively short time turn nearly pure white.

COMMON HORSE AND PONY BREEDS WORLDWIDE

Horse Breeds

AFRICA
Egyptian Arabian, Libyan Berber, Berber, Fulani horse, Nigerian horse, Basuto pony
ARGENTINA
Criollo
AUSTRALIA AND NEW ZEALAND
Brumby, Wales horse, New Zealand pony
AUSTRIA
Lipizzan
BELGIUM
Belgian warm-blooded horse, Brabant, Ardenner
BRAZIL
Crioulo, campolino, mangalarga
CANADA
Royal Canadian mounted police horse, Stable Island pony, Canadian cutting horse
CHINA
China pony
CZECH REPUBLIC AND SLOVAKIA
Kladrub
DENMARK
Fjord horse, Frederiksborger, Knabstniper, Jutländer
FINLAND
Finnish universal, Finnish draft horse
FRANCE
French thoroughbred, half-bred trotting horse, Norman trotter, Norman horse, Anglo-Arabian Camargue horse, Ardenne horse, trait du nord, Bretonne, Percheron, Boullonnais, Poitevine
GERMANY
East Prussian horse, Hannoveraner, Oldenburger, Holsteiner, Dülmener, Württemberger, Rhinish, Schleswiger, Pinzgauer
GREAT BRITAIN
Exmoor pony, new forest pony, Dartmoor pony, fell pony, dales pony, Welsh mountain pony, Welsh pony, highland pony, Shetland pony, Welsh cob, English thoroughbred, Shire, Clydesdale, Suffolk punch
GREECE
Peneia pony, Pindos pony, Skyros pony
HAITI
Haiti pony
HUNGARY
Nonius, furioso, lipizzan, Arabian (shagya)
ICELAND
Icelandic pony
INDIA
Kathiawar pony, Marwar pony
INDONESIA
Sumba pony, sandalwood pony, Sumbawa pony, Java pony, Timor pony, Batak pony, Bali pony
IRAN
Persian Arabian, darashoori (schiras horse), jaf, tchenarani, Turkomane, Polo ponies, Turkmen pony, British pony
IRELAND
Irish Clydesdale
ITALY
Salemer, Kalabrier, Aveligneser
MEXICO
Mexican horse
NETHERLANDS
Frisian horse, gelderse, Groninger
NORWAY
Fjord horse, Döle horse, Gudbrandsdaler, Döle trotting horse
PERU
Criollo (costeno), morochuco
POLAND
Huzuler, komik, Sokólsker, Mazure, Poznan horse Arabian, Anglo-Arabian
PORTUGAL
Lusitanian
RUSSIA
Viatka, zemaitnika (petschora), tori, Bukhyonii horse, Kirghiz horse, karabagh, lokai, yomud, akhal-tekkiner, Arabian, Métis, Orlov trotting horse, Russian-American trotting horse, Vladimir
SPAIN
Sorreia, Andalusian, Arabian
SWEDEN
Gotlander (Skogruss), Swedish warm-blooded horse
SWITZERLAND
Freiburger, Einsiedler horse, Anglo-Norman, Holsteiner

COMMON HORSE AND PONY BREEDS WORLDWIDE *(continued)*

TIBET
Tibetan horse
TURKEY
Anatolian pony, Karakabeyer
UNITED STATES
American thoroughbred, quarter horse, Kentucky saddle horse, Tennessee walking horse, Missouri fox-trotting horse, Morgan, standardbred trotting horse, Spanish mustang (the wild horse of Wyoming), galiceno, Chincoteague pony, Assateague pony, American Cleveland-bay, American hackney, Welsh and Shetland ponies, American pony (P.O.A. = Pony of America), pinto, appaloosa, palomino, albino, American Percheron, American Belgian
VENEZUELA
Llanero (prairie) horse

Donkey Breeds

Poitou ass
Puli ass
Spanish giant ass
Gascogne ass
Savoy ass
Sicilian ass
Macedonian ass
Maskat ass

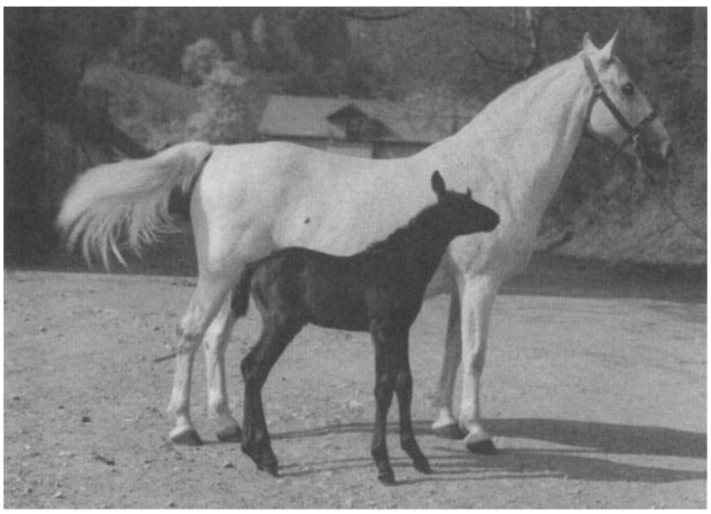


Fig. 2. White mare and black colt of a domesticated horse found mainly in Germany and Austria. At an early age, the solid-black colt changes its pigmentation and becomes a white adult.

Asses

One breed of wild asses occurs in Asia and two breeds occur in Africa. However, all are believed to have been derived from the Nubian Wild Ass. Possibly the best known is the Onager (*Equus onager*), of a rust color, and very horse-like in appearance. The animal is found in the more arid regions of India and throughout Iran and Afghanistan. Traveling in small herds, the animals appear to prefer a semi-desert habitat. Possibly, the Kiang (*E. hemionus*) accounts for the greatest population of wild asses today. This animal is found in Tibet and Mongolia and is considerably larger than the Onager. In winter, the animal has a coat of dark, shaggy hair and a white and red coloration in summer. Another breed or two, small in stature and in number, can be found in the semiarid parts of Mongolia. The asses found in Africa are usually of a gray tone, with white muzzles and white under sides. The Nubian is characterized by distinct black shoulder stripes. The African asses prefer mountain and desert terrain. A domesticated ass is referred to as a donkey.

The Quagga (couagga) is now an extinct species. It was a South African animal (*E. quagga*) related to the zebras and asses. It was reddish-brown above, blending to white on the legs, and marked with dark brown stripes on the head, neck, and fore part of the body. The last known specimen died in 1875 at the Berlin Zoo, although hopefully some wild specimens still may be roaming in the secluded areas of South Africa. The animal was overharvested for use as native labor and food.

A mule is a hybrid between the domestic horse and ass, produced by mating a mare and a jack. The reciprocal cross of stallion and she ass is called a hinny. Mules have large ears, small hoofs, and tufted tail characteristic of the ass and the stature of the horse. They are strong and hardy, resistant to disease and generally adverse conditions. Since they are infertile, they are always bred by crossing the two species. Some authorities claim that hinnys are smaller and lacking in the qualities desired in mules; while others claim that both forms fall within the range of variation to be expected in the hybrid.

Zebras

The Zebra (*E. burchelli*, . . .) is related to the asses and the quagga and is distinguished by the complete or nearly complete transverse striping of the body and legs. The stripes vary in the several species from white to yellow brown, alternating with dark brown to black. They range throughout much of Africa south of the Sahara Desert, but are concentrated in the south. Sub-species include: The Common Zebra (characterized by a V-shaped junction-pattern occurring in the middle of the sides); the Damaraland Zebra (Zaire, Zambia, Botswana regions); the East African Zebra (Rhodesia, Abyssinia, the Sudan); and Selou's Zebra (central and southeast Africa). General characteristics of zebras include: Grazing like horses; traveling in big herds; mixing with other animals; a main dietary item for lions; shy and nervous; can be quite pugnacious in self-defense, kicking vigorously with either front or hind feet and inflict severe bites; coloration and striping puts them at disadvantage when viewed against a green backdrop, but is an advantage in tall grass and open plains; frequently harbor intestinal parasites which are believed to aid in their digestion. See Fig. 3.

Fig. 3. Chapman's zebra (*Equus antiquorium*.)

Grevy's Zebra (*Dolichohippus*) is not classified with *Equus*. However, it appears much as a horse, with large head and big ears. The animal prefers desert scrub foliage and requires a minimum of water. It is found in Somalia. Certain anatomical distinctions set it in its own genus.

Other members of the order of *Perissodactyla* are listed under **Perissodactyla.**

HORST. See **Fault.**

HOST. An animal which is used as a source of food by a parasite. The parasite may live on the surface of the body or within it and may be harmless or harmful, but in all cases the host is the source of its food.

HOT SHORTNESS. Brittleness of metals in the hot working temperature range.

HOTWELL. A hotwell is a tank or container in which heated liquid collects. An example is the hotwell attached to and made part of a steam condenser of the surface type. As the steam is condensed, the condensate drops to the bottom of the condenser shell and flows into the hotwell, from which it is pumped.

HOT WORKING. Plastic deformation of metals at temperatures sufficiently elevated so that the effects of the working are nullified by concurrent softening processes. Thus, when steel is hot worked by rolling at a white heat, the metal recrystallizes and softens almost immediately after it is deformed. Similarly, the deformation of lead at room temperature is also hot working and accounts for the fact that it is not possible to work-harden this material at this temperature. An empirical rule states that the lower limit of the hot-working temperature range is the recrystallization temperature.

Forging, rolling, pressing, extruding, swaging, drawing, or forming of metals at temperatures above their recrystallization temperatures are examples of hot working.

HOUR ANGLE. In reference to a celestial object, the spherical coordinate, in the equatorial system of coordinates, measured in the plane of the celestial equator from the local meridian, in the direction of apparent rotation of the celestial sphere, to the intersection of the hour circle through the object with the equator. Since time and hour angle are practically synonymous (e.g., the hour angle of the mean sun is local mean time), the determination of hour angle is vitally necessary for the determination of local time and, hence, longitude.

At sea, hour angle is determined by measuring the altitude of the object by means of the sextant, reducing the observed altitude to true geocentric, and solving the astronomical triangle. For the solution of the triangle, both the declination of the object and the latitude of the observer must be known. The declination may be immediately obtained from the tabulated coordinates of the object, but the latitude can be obtained only by some previous observation. If the ship is in motion, the latitude must be obtained by dead reckoning from the previously determined position.

See also **Celestial Sphere** and **Astronomical Triangle.**

HOUR CIRCLE. See **Celestial Sphere.**

HOUSEFLY (*Insecta, Diptera*). A true fly, *Musca domestica*, well known for its habit of frequenting houses and alighting on all kinds of food. Since it also visits filth of any kind, it is an important carrier of disease, especially typhoid fever, and has been the object of public health crusades for many years. With the improvement of sanitation, the danger has been lessened, although it has not been entirely eliminated.

The housefly breeds in horse manure and in various kinds of decaying organic matter. Proper disposal of such wastes is an important measure in the control of the insect.

The housefly is found worldwide except at high altitudes. Hair covers most of the head of the housefly. Its jaws work horizontally and the stubby snout has a piercing stylet. It does not bite, but pierces and sucks its food. The eyes are large, covering about three-fourths of the facial area. The eyes of a fly are made up of approximately 4000 six-sided facets. All facets together frame the total object of view. However, the fly does not focus for a sharp image, nor does it have the ability to close its eyes. The vision is believed to be reasonably sharp for distances of 2 to 3 feet (up to 1 meter) or less.

Several generations of flies may be born in one season. They live for only a few weeks, but their eggs live on in fertile debris until spring.

The housefly is a principal agent for transmitting many diseases in areas where it is not carefully controlled. The legs are hairy and well designed to carry filth. Infection may be transferred by the piercing proboscis of the fly, or the fly may vomit its own food, leaving a trail of infection. As many as several million microorganisms may be found in the intestines of a housefly. The flying speed of the housefly is approximately 5 miles (8 kilometers) per hour.

HUBBLE'S LAW. See **Cosmology.**

HUE. The attribute of color perception that determines whether it is red, yellow, green, blue, purple, or the like. White, black, and gray are not considered hues.

HUMBOLDT CURRENT (also called Peru Current). The cold ocean current flowing north along the coasts of Chile and Peru. It is one of the swiftest of ocean currents. The Peru current originates where part of the water that flows toward the east across the subantarctic Pacific Ocean is deflected toward the north as it approaches South America. The northern limit of the current can be placed a little south of the equator, where the flow turns toward the west, joining the south equatorial current.

The southern portion of the Humboldt current is sometimes called the *Chile current*.

HUMECTANTS AND MOISTURE-RETAINING AGENTS. Substances that have affinity for water, with stabilizing action on the water content of a material, are called *humectants* or moisture-retaining agents. Ideally, a humectant maintains within a rather narrow range the moisture content caused by humidity fluctuations. These materials are widely used in certain food products, as well as tobacco, and in recent years have taken on increasing importance in the case of intermediate-moisture foods. Traditionally, humectants have been used to retain moisture in foods like coconut and marshmallows which otherwise would quickly dry and become tasteless. For example, flaked coconut is kept moist in the container by adding glycerine and glyceryl monostearate.

Among the most commonly used hemectants are glycerine, potassium polymetaphosphate, propylene glycol, sodium chloride, sorbitol, sucrose, and triacetin. Also, phosphates are added to the pickling solutions used to treat cured meats, such as ham, bacon, corned beef, etc., by soaking or injection. Their principal purpose is for moisture binding to reduce the loss of fluids during curing and cooking.

During the last few years, important research has gone into the addition of multiple humectants and water to food systems. Studies have shown that a hysteresis effect may occur with certain humectants, i.e., a different rate of moisture absorption than the rate for moisture desorption. Multiple humectants tend to compensate these hysteresis effects, giving uniform rates in both directions.

HUME-ROTHERY RULES. When alloy systems form distinct phases, it is found that the ratio of the number of valence electrons to the number of atoms is characteristic of the phase (e.g., β, γ-, ϵ-) whatever the actual elements making up the alloy. Thus, both $Na_{31}Pb_8$ and Ni_5Zn_{21} are γ-structures, with the electron-atom ratio 21:13. The rules are explained by the tendency to form a structure in which all the Brillouin zones are nearly full, or else entirely empty.

HUMIDITY. Generally, some measure of water-vapor content of air. *Absolute humidity* is the ratio of the mass of water vapor present to the volume occupied by the mixture; that is, the density of the water vapor component. The percentage of water vapor in the total composition of the air may be determined by passing a measured quantity of air through a tube containing an absorbing substance that removes all the vapor, and which can be weighed before and after the absorption.

Absolute humidity is usually expressed in grams of water vapor per cubic meter or, in engineering practice, in grains per cubic foot. Because this measure of atmospheric humidity is not conservative with respect to adiabatic expansion or compression, it is not commonly used by meteorologists. As occasionally used in air-conditioning practice, absolute humidity refers to the number of grains of water vapor per pound of moist air, which is dimensionally identical with the specific humidity (defined below).

Critical humidity is the point at which the partial pressure of water vapor in the atmosphere is equal to the saturation vapor pressure. Condensation on suitable nuclei will occur when the humidity reaches or exceeds this value.

Relative humidity is the ratio of the actual vapor pressure of the air, at any temperature, to the maximum of saturation vapor pressure at the same temperature. It expresses the vapor content as a fraction or percentage of the concentration necessary to render the vapor saturated at the given temperature. At the dew point, the relative humidity is 100%. A rise of temperature without the addition of more vapor reduces the relative humidity (but not the absolute humidity), while a fall of temperature increases it and may bring about saturation. Relative humidity is measured by the hygrometer.

Specific humidity is the (dimensionless) ratio of the mass of water vapor to the total mass of the system. It may be approximated by the mixing ratio for many purposes:

$$q = \frac{w}{1 + w}$$

where q is the specific humidity and w the mixing ratio.

See also **Psychrometric Chart.**

HUNTING. The tendency of a rotating mechanism which normally should operate at constant speed to pulsate in speed above and below the normal point, is known as hunting. It may occur in prime movers controlled by governors which are too isosynchronous, or in electric apparatus where rotating and stationary parts are electrically coupled. The nature of such coupling is essentially elastic, and may, under certain circumstances, lead to hunting action on the part of the rotor. Governors which hunt must be corrected by the use of dash pots or other damping devices, and the introduction to the governor characteristic of a slight amount of speed regulation.

See also **Governor.**

HYADES. An open, V-shaped, moving cluster of stars in the constellation of Taurus. References to the Hyades are to be found in all the ancient literatures, Virgil referring to them as the "rainy Hyades." The group is exceedingly rich in double stars, which, even with a small telescope and low magnifying power, present a beautiful appearance.

The Hyades form one of the best known of the so-called moving star clusters. The brightest star of the Hyades, Aldebaran, is not a member of the cluster, but has an independent motion through space and just happens to be in its present position at this time.

HYBRID (Biological). An organism produced by parents belonging to different species or to different strains of the same species. A hybrid combines characteristics derived from the two parent stocks and in some cases is more desirable than either. Beauty of flowers, productivity of various plants, and appearance and hardiness of animals have been enhanced by controlled hybridization.

When a hybrid is once secured its propagation is hampered by the fact that the diverse hereditary characters are reassorted in hereditary transmission by sexual reproduction. Hybrids are often infertile but even when they are capable of producing offspring they rarely breed true. The mule is the only animal hybrid of great value, and it is produced always by parents of the two species, horse and ass. Plant hybrids are not subject to this limitation, for they can usually be propagated by bulbs, cuttings, or grafts. Plants produced in this fashion are sometimes referred to as cultivars. See also **Plant Breeding.**

HYDANTOIN PROCESS. See **Amino Acids.**

HYDATID DISEASE. Also referred to as *echinococcosis*, this is an infection with *Echinococcus granulosus* or *E. multiocularis*, which are cestodes. These worms live in the intestines of dogs and wolves, whose feces include infective eggs. Such material may inadvertently find its way to a substance that is ingested by sheep, cattle, or humans. Infection with *E. granulosus* is most commonly found in regions where sheep and cattle are produced as, for example, in the western United States, parts of Canada, and Alaska. Upon ingestion of eggs, oncospheres are carried by the bloodstream to the liver, lungs, and other organs. These cause the development of cysts, often with neurologic symptoms. The cysts may grow to a diameter of 6 inches (15 centimeters), and contain many worms.

Mice and small animals are intermediate hosts to *E. multiocularis* which resides in the intestines of foxes and dogs. These cestodes are found mainly in the Northern Hemisphere—Europe, Canada, Alaska, and north-central United States. They produce extensive alveolar hydatid cysts, frequently resulting in jaundice.

Frequently, the therapy for hydatid disease involves surgical excision of cysts. Cryosurgery is frequently used. Some success has been reported with the drug mebendazole in the treatment of the disease. See also **Tapeworm.**

HYDATOGENESIS. A term used by petrologists to designate the process by which rocks are formed from highly aqueous solutions. Some petrologists limit the use of the term to rocks which have been deposited from water-rich magmatic solutions.

HYDRA (the serpent). A southern constellation that forms the outline of a serpent.

HYDRATE. Excluding the loose usages in which the term hydrate indicates merely the presence of water or of its elements in 2:1 ratio, as in carbohydrate, the term hydrate denotes the appearance of water in compounds. There are a number of ways in which water may appear in stoichiometric proportions in compounds. Moreover, these ways may be described from more than one point of view. A somewhat systematic approach is to view these compounds from the point of view of the extent of integration of the water, or its elements, into the compound.

The term "water of constitution" is a somewhat old usage, applied to compounds in which no H_2O groupings appear in the structure of the compound, but the compound may undergo reaction, usually reversible, in which water is one of the products. Magnesium hydroxide and sulfuric acid could thus be said to have "water of constitution," even though it appears in their structure as hydroxyl groups, or hydroxyl groups and hydrogen atoms (protons).

The term "cationic water" may be used to describe the situation in which water appears in coordination compounds apparently joined to cations by covalent bonds. However, the fact that a number of such compounds exhibit "hydrate isomerism" is evidence for cationic bonding, as well as it is for the existence of other forms of these compounds in which the presence of water is due to electrostatic attractions or crystal stability requirements.

The term "anionic water" describes the situation in which water is joined to anions through covalent bonds, or more frequently, through hydrogen bonds. The type case is copper(II) sulfate pentahydrate, where the cation has a coordination number of four and presumably the fifth molecule of H_2O is bound to the sulfate ion (as well as to other H_2O molecules) by hydrogen bonds.

The term "lattice water" is commonly applied to cases in which the water molecules are occupying definite positions in the crystal lattice but are apparently not coordinated with either cations or anions. Again, clear-cut cases are those in which the compound is so highly hydrated that both lattice water and "ion water" are present.

The water in crystals may, however, be present in other than definite lattice positions. For example, the water molecules may be found in holes in the lattices, or they may occupy random positions in the lattices. The latter situation is often found in ion exchange resins where loss of water, up to a certain point, does not materially change the lattice structure.

Finally, in essentially noncrystalline materials, such as hydrous precipitates and colloidal gels, the water present is at the limiting case of being a hydrate, in which virtually no bonding, in the chemical sense, exists.

HYDRAULIC CONTROLLER. A device that uses a liquid control medium to provide an output signal which is a function of an input error signal. Aside from the use of a liquid controlling medium, hydraulic controllers are similar in operating principle to electric, electronic, and pneumatic controllers. In fact, there are striking similarities between hydraulic control and pneumatic control. Because a liquid control medium is essentially incompressible, there is an excellent speed of response between controller and final actuating element. Hydraulic control systems also are characterized by high power gain inasmuch as liquids can be converted readily to high pressures or flows through the use of various types of pumps. The final actuators are comparatively simple; most outputs are two hydraulic lines that can be tied directly to a straight-type cylinder to provide a linear mechanical output. Inasmuch as the parts of a hydraulic system are essentially self-lubricating, they have a long life when properly designed.

Limitations of hydraulic control systems include special maintenance problems in connection with hydraulic fluids—fire hazard and leakage, and somewhat higher cost, dependent upon the size of the equipment.

Hydraulic controllers are extensively used as liquid pipeline-pressure controllers where a pipeline control valve can be operated against sudden pressure surges. Edge-guiding control systems are also common. For example, a hydraulic system can control the edge of a moving steel strip (typical strip velocity of 1,000 feet; 300 meters per minute) to plus or minus $\frac{1}{64}$-inch (0.4 millimeter) and accomplish this by shifting a coil of steel weighing up to 50,000 pounds (22,680 kilograms).

The hydraulic relay is the heart of a hydraulic control system. Commonly, a jet-pipe valve is used—as shown in Fig. 1. By pivoting a jet pipe, a fluid jet can be directed from one recovery port to another. The fluid energy is converted entirely into a velocity head as it leaves the jet-pipe tip and then is reconverted into a pressure head as it is recovered by the recovery ports. The relationship between jet-pipe motion and recovery pressure is shown in Fig. 2. Although the jet pipe can be used at higher pressures, most applications are less than 800 psi (~54 atmospheres). The proportional operation of the jet pipe makes it useful in proportional-speed floating systems (integral control) as indicated in Fig. 3(a). Position feedback can be provided by rebalancing the jet pipe from the work cylinder as shown in Fig. 3(b). A proportional-plus-reset arrangement is shown in Fig. 3(c). In this last instance, the proportional feedback is reduced to zero as the oil bleeds through the needle valve. The hydraulic flow obtainable from a jet pipe is a function of the pressure drop across the jet pipe.

Flapper valves of the type shown in Fig. 4 also are used. The spool valve, shown in Fig. 5, when used as a hydraulic relay usually is constructed in either a three-way or a four-way valve-porting arrangement. See Fig. 6. The mechanical displacement of the spools allows the hydraulic-pressure supply to be ported in a fashion that will dis-

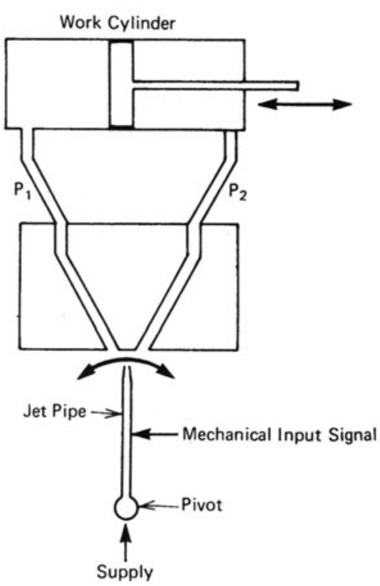


Fig. 1. Jet-pipe valve used in hydraulic control system.

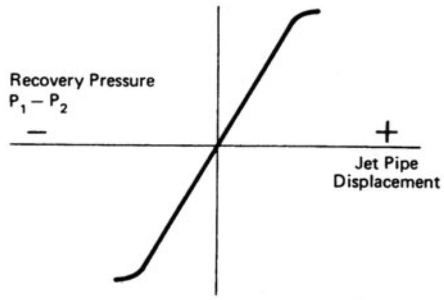


Fig. 2. Jet-pipe motion and recovery pressure relationship.

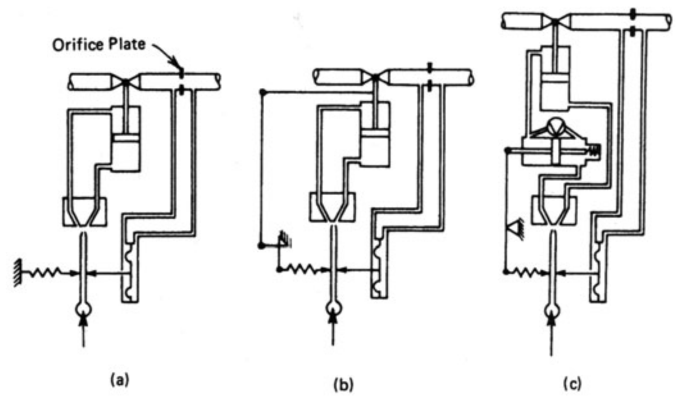


Fig. 3. Hydraulic controllers: (1) proportional speed floating control; (b) proportional position control; (c) proportional plus reset control.

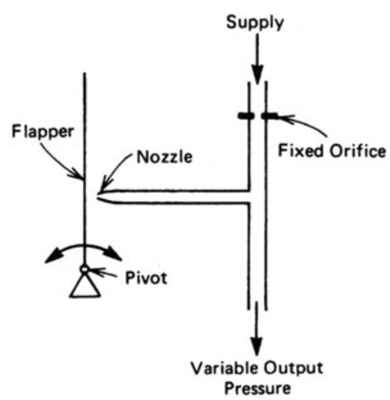


Fig. 4. Single flapper valve.

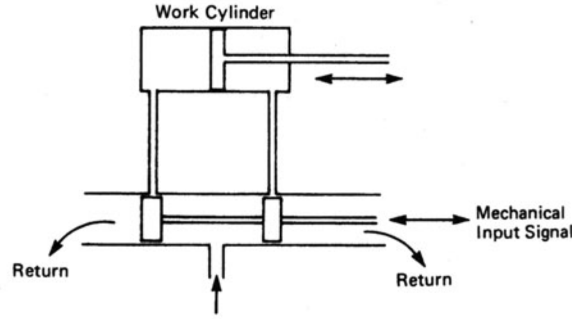


Fig. 5. Spool valve.

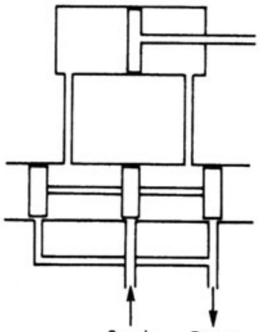


Fig. 6. Four-way spool valve.

place the work cylinder in either direction, depending upon the spool displacement.

HYDRAULIC RADIUS. The theory of hydraulics indicates that the ratio of the frictional area to the volume of the liquid stream is an important dimension governing the friction loss. The hydraulic radius, which expresses this fact, is the cross-sectional area of flow divided by the wetted perimeter of a cross section of the conduit. The hydraulic radius of a circular pipe flowing full of water is one-fourth of the diameter. The hydraulic radius of an open canal is the cross-sectional area of the stream divided by the wetted perimeter of the cross section.

HYDRAULICS. Hydraulics is the dynamics of liquids (hydrodynamics), especially applied to the practical problems of engineering. Although this general definition is entirely correct, in common usage hydraulics is the study of water at rest or in motion. This conception of hydraulics is used in this article. The mechanics of fluids (liquids and gases) in general is termed fluid mechanics. A basic proposition of hydraulics is that water is incompressible. While this condition is not completely met in fact, the compressibility of water is so small as to be negligible for practically all propositions of hydraulics. The viscosity of water varies with the temperature and is one reason for change of conditions of water flow in pipes with changing temperature. The unit weight of fresh water is usually taken as 62.4 pounds per cubic foot (~1000 kg/cu meter).

The science of hydraulics is divisible into hydrostatics and hydrokinetics. Hydrostatics is the hydrodynamics of liquids considered apart from their motion: hydrokinetics is the hydrodynamics of moving, especially flowing, liquids. Among the subjects included in any study of hydrostatics are the following: (1) the pressure on a submerged area of any shape or inclination, (2) the measurement of pressure on water at rest by manometers or pressure gauges, (3) buoyancy and flotation. Practical application of (1) is to be found in problems associated with water gates, large valves, pressure against dams, tanks, hydraulic presses, etc.

Hydrokinetics includes a great many different phases of hydraulics. Most of these will be found treated in specialized articles, references to which are given below. The flow of fluids supplies many cases of the application of hydraulic science. Flows of steady, uniform, unsteady and non-uniform types, and the friction losses occasioned thereby, in closed or open conduits; the measurement of flows and the discharges under given conditions, are part of this phase of hydraulics; also, there is to be considered the flow of water through openings, such as orifices, nozzles, and weirs. The flow of water in pipe lines offers a great many problems in addition to friction: the discharge through different sections of branching and looping pipes, siphons, fittings, valves, etc., is included. Measurement of discharge of large amounts of water, as in stream and river flow, offers problems different from those met in closed conduits. Furthermore, the forces occasioned by deviated flows of water, as met in hydraulic turbines, the pump, and other hydraulic machinery, are fit subjects to be included in any study of hydromechanics. See also **Fluid Flow.**

HYDRAZINE. $H_2N{\cdot}NH_2$, formula weight 32.04, colorless, fuming liquid, mp 1°C, bp 113°C, sp gr 1.011, decomposes when heated above 350°C at atmospheric pressure into N_2 and NH_2, also decomposes in presence of a catalyst (e.g., platinum) into N_2 and NH_3. Hydrazine burns when ignited in air with a violet-colored flame. The compound is soluble in all proportions with H_2O and is soluble in alcohol. Hydrazine forms a hydrate with one molecule of H_2O. Upon moderate heating or in a vacuum, the hydrate yields hydrazine and H_2O. Hydrazine is a base slightly weaker than NH_4OH.

Hydrazine is a tonnage chemical with numerous uses, including that of a propellant for rockets, yielding exhaust products at a high temperature and of a low molecular weight; use as a strong reducing agent in the manufacture of various chemicals; and as a blowing agent for foamed rubber. The compound reacts with citric acid to form *Continazin*, an antituberculan drug.

Although the earlier processes for the commercial production of hydrazine used urea as a raw material, modern processes employ direct ammonia oxidation. In one such process, reactions occur in two steps:

(1) $NH_3 + NaOCl \rightarrow NH_2Cl + NaOH$,
(2) $NH_3 + NH_2Cl + NaOH \rightarrow H_2N{\cdot}NH_2 + NaCl + H_2O$.

High-grade hypochlorite is required for Step 1. Special agents, such as gelatin, ethylenediamine tetracetic acid, glue, high alcohols, and formaldehyde, are required to inhibit undesirable side reactions that would reduce the hydrazine yield through formation of ammonium chloride and N_2. In another hydrazine process, chlorine, NH_3, and H_2SO_4, along with methylethyl ketone, are used as the charge. The products of this process include hydrazine hydrate, hydrazine sulfate, ketazine, and dialkyldiazacyclopropane. Hydrazine also is used as a start-up ingredient in the preparation of cooling water for nuclear reactors where it is desired to keep the oxygen content of the water to an absolute minimum and thus decrease corrosion. Oxygen reacts with hydrazine. $H_2N{\cdot}NH_2 + O_2 \rightarrow N_2 + 2H_2O$. When no oxygen is present in the water, the hydrazine acts as a sink for dissolved oxygen that may enter later, by maintaining metal oxides at their lower oxidation states.

Hydrazine forms two series of salts: (1) hydrazinium (1+) chloride, $H_2NNH_3^+Cl^-$, nitrate, $H_2NNH_3^+NO_3^-$, hemisulfate, $(H_2NNH_3^+)_2SO_4^{2-}$, (2) hydrazinium (2+) chloride, $H_3NNH_3^{2+}(Cl^-)_2$, dinitrate, $H_3NNH_3^{2+}(NO_3^-)_2$, hydrogen sulfate, $H_3NNH_3^{2+}(HSO_4^-)_2$, all soluble in H_2O. This last is produced when hydrogen azide reacts with concentrated H_2SO_4. It is very hygroscopic and decomposes in aqueous solution to give the slightly soluble monosulfate and H_2SO_4. The monosulfate and difluoride, which have been thought to have the structures $N_2H_5^+HSO_4^-$ and $N_2H_5^+HF_2^-$ in the solids, have been shown in fact to be $N_2H_6^{2+}SO_4^{2-}$ and $N_2H_6^{2+}(F^-)_2$. Hydrazinium azide, $N_2H_5^+N_3^-$, is a soluble solid.

In the laboratory, hydrazine can be prepared by converting one-half of a given amount of NH_3 into chloramine, NH_2Cl, by sodium hypochlorite solution in the presence of a colloid and heating. The remaining one-half of the NH_3 reacts with chloramine to form hydrazine. The product is then cooled to 0°C and H_2SO_4 added in amount to react with the hydrazine to form hydrazine sulfate, $N_2H_6SO_4$, insoluble solid. Hydrazine hemisulfate, $(N_2H_5)_2SO_4$, is soluble in H_2O. It can also be made by the reaction of NH_3 and hydroxylamine-O-sulfonic acid.

Phenylhydrazine is a colorless liquid, slightly soluble in H_2O, miscible in all proportions with alcohol or ether, forms salts with acids, e.g., phenylhydrazine hydrochloride or phenylhydrazinium chloride, $C_6H_5NHNH_3Cl$, is a powerful reducing agent, with alkaline copper(II) salt solution (Fehling's solution) yields copper(I) oxide precipitate, reacts with carbonyl group of aldehydes or ketones yielding phenylhydrazones, white solids, of definite melting point and utilized in identification of aldehydes and ketones, e.g., acetaldehyde phenylhydrazone, $CH_3CH{:}NNHC_6H_5$.

Phenylhydrazine, as hydrochloride solution plus sodium acetate, reacts with polyhydroxy aldehydes or ketones yielding *osazones* or diphenylhydrazones, yellow solids, of definite melting point and utilized in identification of sugars, e.g., phenyl-d-glucosazone, $CH_2OH(CHOH)_3C{:}(NNHC_6H_5)CH{:}(NNHC_6H_5)$ plus aniline $C_6H_5NH_2$ plus NH_3.

Attention should be given to the difference between osazones and osones. An *osone* is formed by reaction of an osazone with HCl, e.g., glucosone, $CH_2OH(CHOH)_3CO{\cdot}CHO$.

1,1-Diphenylhydrazine is made by reduction of diphenylnitrosamine, $(C_6H_5)_2N{\cdot}NO$, by zinc plus acetic acid, the nitrosamine being formed by reaction of diphenylamine, $(C_6H_5)_2NH$, and nitrous acid.

Tetraphenylhydrazine is a white solid, soluble in chloroform, acetone, benzene, or toluene, and upon standing is changed into triphenylamine plus azobenzene. In solution, tetraphenylhydrazine dissociates into nitrogen diphenyl, $(C_6H_5)_2N{\cdot}$, free radical, which in toluene at 90°C reacts with nitric oxide, NO. Tetraphenylhydrazine is formed by oxidation of diphenylamine, $(C_6H_5)_2NH$, by lead dioxide.

Hydrazine reacts with ketones to form *azines*.

HYDRAZOIC ACID. HN_3, formula weight 43.03, colorless, odorous, poisonous liquid, mp −80°C, bp 37°C, explodes with marked violence. Also known as azoimide and hydronitric acid, the compound is miscible in all proportions with H_2O, alcohol, and ether. Hydrazoic acid reacts (1) with metals, e.g., magnesium, aluminum, zinc, iron, to form azides or hydrazoates (or trinitrides), (2) with heavy metal salt solutions to form insoluble azides, e.g., silver azide AgN_3, mercury(I) azide HgN_3, lead azide PbN_6. Silver, mercury(I), and copper(I) azides decompose in the light to form nitrogen plus the metal. (3) It reacts with NH_4OH to form ammonium azide $NH_4{\cdot}N_3$, (4) with hydrazine to form hydrazine azide $N_2H_4{\cdot}HN_3$, (5) with sodium hypochlorite plus acetic acid to form chlorazide ClN_3, explosive, (6) with sodium amalgam to form NH_3 with some hydrazine, (7) with potassium permanganate to form nitrogen and H_2O.

Hydrazoic acid is formed (1) by reaction of sodium nitrate with molten sodamide, (2) by reaction of nitrous oxide with molten sodamide, (3) by reaction of nitrous acid and hydrazinium ion ($N_2H_5^+$), (4) by oxidation of hydrazinium salts, (5) by reaction of ethyl nitrite with NaOH solution and acidifying. See also **Azides.**

HYDRAZONES. The products of the reaction between an aldehyde or a ketone with phenylhydrazine are termed *hydrazones*. Sometimes the compounds are referred to as phenylhydrazones.

$CH_3{\cdot}CHO$ (acetaldehyde)	$+ C_6H_5{\cdot}NH{\cdot}NH_2 \rightarrow$ (phenylhydrazine)	$CH_3{\cdot}CH{:}N{\cdot}NH{\cdot}C_6H_5 + H_2O$ (acetaldehyde hydrazone)
C_6H_5CHO (benzaldehyde)	$+ C_6H_5{\cdot}NH{\cdot}NH_2 \rightarrow$	$C_6H_5{\cdot}CH{:}N{\cdot}NH{\cdot}C_6H_5 + H_2O$ (benzylidenehydrazone)
$(CH_3)_2CO$ (acetone)	$+ C_6H_5{\cdot}NH{\cdot}NH_2 \rightarrow$	$(CH_3)_2C{:}N{\cdot}NH{\cdot}C_6H_5 + H_2O$ (acetone hydrazone)
$C_6H_5{\cdot}CO{\cdot}CH_3$ (acetophenone)	$+ C_6H_5{\cdot}NH{\cdot}NH_2 \rightarrow$	$(C_6H_5)(CH_3)C{:}N{\cdot}NH{\cdot}C_6H_5 + H_2O$ (acetophenonehydrazone)

Several of the hydrazones may be decomposed by strong acids whereupon the original aldehyde or ketone is regenerated, along with the formation of a phenylhydrazine salt. When reduced, hydrazones yield primary amines.

HYDRIDE. A binary compound of hydrogen. Hydrides traditionally have been classified into three groups. In modern terminology, these are conveniently designated as covalent, electrovalent and metallic, although reference to the entries for hydrogen and the various hydrogen halides shows that a number of binary hydrogen compounds are partly ionic and partly covalent. See also **Hydrogen.**

Covalent hydrides are formed by the non-metals. In general, the elements of main groups III to VII form single compounds consisting of a single atom of the element combined with a number of hydrogen atoms equal to the number of electrons which the element needs to complete its octet. Exceptions are beryllium, aluminum, and indium, which have polymeric hydrides, and boron and gallium, which have dimeric hydrides. Then also the elements of lower atomic number in main groups IV, V and VI (carbon, silicon, germanium, nitrogen, phosphorus, oxygen, and sulfur) and boron form more than one hydride. The covalent hydrides are volatile with low melting points and low boiling points (except as those properties are modified, as in the case of hydrogen fluoride, water and ammonia, by hydrogen bonding). They are nonconductors of electricity in the liquid state or when dissolved in nonpolar solvents.

Complex hydrides are formed by some elements (particularly in main group III) having too few electrons to attain an octet in the neutral hydrides. These are structurally similar to the corresponding complex chlorides and are all excellent reducing agents. The most important are the tetrahydroborate, BH_4^-, -aluminate (frequently called alanate in the European literature), AlH_4^-, -gallate, GaH_4^-, and -indate, InH_4^-. There is evidence for polymeric ions, such as $B_2H_7^-$. Anions derived from higher hydrides are also known, e.g., $B_4H_{11}^-$.

Only the strongly electropositive elements, the alkali metals, the alkaline earth metals, and certain lanthanide and actinide metals, form electrovalent hydrides. The compounds are definitely crystalline, the alkali hydrides being cubic, but the structure increasing in complexity in going from main group 1 to main group 2 and to the lanthanides and actinides. In fact, hydrides of the last two groups, while approaching the alkali and alkaline earth hydrides in electropositive character, and while also giving evidence of the presence of H^- ions in their structures, are usually non-stoichiometric, compositions such as $CeH_{2.70}$, $PrH_{2.85}$ and $ThH_{3.07}$ being found. In this respect, those compounds approach in character the metallic hydrides.

This gradation in properties extends to the metallic hydrides themselves, some of which, such as copper hydride, approaches closely, but never quite reaches, a 1:1 atomic ratio of hydrogen to copper. In the case of palladium, the pressure-composition graph at temperatures below 200°C indicates a wide range of composition at little or no increase in pressure. At higher temperatures the flat portion shortens, and two breaks develop before and after it, indicating solid solutions, one of which approaches a 2:1 atomic ratio of H to Pd.

A group of complex metal hydrides have been used successfully for the preparation on an industrial scale of many organic and metallorganic compounds. Among these complex hydrides are highly reactive lithium aluminum hydride and the related sodium aluminum hydride and magnesium aluminum hydride. A more selectively reactive group of complex hydrides are the lithium, sodium, and potassium borohydrides. These compounds also have properties which make them useful as high energy fuels and rocket propellants.

Lithium aluminum hydride, $LiAlH_4$, also known as lithium aluminohydride and often abbreviated as LAH is prepared by the reaction of lithium hydride and aluminum chloride in ether solution

$$4LiH + AlCl_3 \rightarrow LiAlH_4 + 3LiCl$$

with some prior prepared complex hydride used as a seeding material to control the reaction rate. Lithium aluminum hydride forms a microcrystalline powder which is stable in dry air but decomposes above 125°C. It is soluble in many organic compounds like ether, dimethyl Cellosolve, tetrahydrofuran, but is only slightly soluble in dioxane. It reacts vigorously with water, yielding hydrogen in a manner similar to the reaction of the simple hydrides:

$$LiAlH_4 + 4H_2O \rightarrow 4H_2 + LiOH + Al(OH)_3$$

It reacts with carbon dioxide to form methyl alcohol, or formaldehyde, or formic acid. It is a powerful reducing agent and reduces aldehydes, ketones, quinones, acids, esters, anhydrides, lactones, epoxides, and acid chlorides to the corresponding alcohols; amides, lactams, imides, nitriles, isocyanides, oximes, hydroxylamines, and related compounds to amines; dithiols, disulfides, polysulfides, sulfoxides, sulfones, and related compounds to the mercaptan or sulfide (see **Sulfur**); and aromatic nitro compounds to azo compounds. Olefinic bonds are not attacked unless conjugated with a nitrile, phenyl, or carbonyl group.

Sodium aluminum hydride can be prepared like the lithium analogue but tetrahydrofuran is used as the solvent because the sodium complex is insoluble in ether. The sodium compound produces virtually the same reductions as the lithium compound.

Magnesium aluminum hydride may be prepared by treating an etherate of magnesium bromide with an ether solution of lithium aluminum hydride.

$$2LiAlH_4 + MgBr_2 \rightarrow Mg(AlH_4)_2 + 2LiBr$$

or by use of an excess of magnesium hydride in ether solution with an ether solution of aluminum chloride. While the reducing activity of magnesium aluminum hydride is similar to that of the lithium complex in that polar double and triple bonds such as carbonyl and nitrile groups are reduced whereas nonpolar groups are not attacked, the magnesium complex, however, does not reduce the triple bond of propargyl aldehyde nor the double bond of cinnamic acid in contrast to the lithium complex.

Lithium borohydride, $LiBH_4$, may be prepared by the reaction of aluminum borohydride on ethyllithium or by the action of diborane on ethyllithium. It forms orthorhombic crystals which decompose at 250 to 272°C and while it is stable under usual conditions it is decomposed by humid air. It reacts readily with water and is a strong reducing agent.

Aluminum borohydride, AlB_3H_{12}, is a liquid which boils at about 44.5°C. It ignites in air. It can be prepared by the reaction of diborane with trimethylaluminum. It reacts readily with hydrogen chloride and water to yield hydrogen. It can be used in organic syntheses.

Metals like palladium and platinum absorb hydrogen forming mixtures which may be considered as alloys. Such mixtures may be placed into two groups: those in which the absorption takes place with decrease in temperature like those mixtures of hydrogen with palladium and tantalum; and those which absorb hydrogen with an increase in temperature like those with calcium, iron, nickel, and platinum.

HYDROCEPHALUS. A condition characterized by abnormally large amounts of cerebrospinal fluid around or within the brain, usually associated with enlargement of the cerebral ventricles. See **Meningitis.**

HYDROCHLORIC ACID. HCl (hydrogen chloride gas) in aqueous solution, colorless when pure. Commercial grades of HCl (also known as muriatic acid) generally are marketed in three concentrations: (1) 18° Bé (sp gr 1.1417 at 15.6°C, 27.92% HCl); (2) 20° Bé (sp gr 1.160, 31.45% HCl); and (3) 22° Bé (sp gr 1.1789, 35.21% HCl). Frequently the commercial grades are slightly yellow because of impurities, notably dissolved iron. Fuming hydrochloric acid contains about 37% HCl, with a sp gr 1.194. Reagent grade hydrochloric acid usually is of this latter high strength, and is perfectly clear and colorless. The maximum limits set on impurities commonly are: NH_4 0.003%; arsenic 0.000001%; free chlorine 0.0001%; heavy metals, such as lead 0.001%; iron 0.00002%; sulfates 0.0001%; sulfites 0.0001%; and residue after ignition 0.0005%. A mixture of three parts HCl and one part HNO_3 is known as *aqua regia*, a powerful solvent and oxidizing agent which will dissolve materials that may be unaffected by either acid alone. Gold and platinum are soluble in aqua regia.

Hydrochloric acid is a very-high-tonnage chemical, finding major uses in (1) the cleaning and preparation of metals prior to application of coatings, (2) the recovery of zinc from galvanized iron scrap, (3) the production of numerous chlorides, and (4) production of chlorine. At one time, HCl was extensively used as a source of both hydrogen and chlorine by way of electrolysis. This process was made obsolete many years ago when the chlor-alkali process (electrolysis of sodium chloride brines) was introduced for the production of chlorine. In recent years, however, the production of by-product HCl, resulting from chlorination of numerous organic compounds, has increased. In some of these instances, the installation of a HCl electrolysis plant may be economically feasible. For industrial consumption anhydrous HCl gas also is available in steel cylinders under a pressure of 1,000 psi (68 atmospheres). Hydrochloric acid forms a constant-boiling solution with H_2O (20.22% HCl) which has a bp 108.58°C (760 mm Hg).

Dilute HCl reacts (1) with many hydroxides, e.g., NaOH, to yield the corresponding chloride, e.g., sodium chloride, solution, (2) with many ordinary oxides, e.g., magnesium oxide, to yield the corresponding chloride, e.g., magnesium chloride, solution, (3) with many carbonates, e.g., calcium carbonate, to yield the corresponding chloride, e.g., calcium chloride solution plus CO_2, (4) with many sulfides, e.g., ferrous sulfide, to yield the corresponding chloride, e.g., ferrous chloride, solution plus H_2S, (5) with many metals, e.g., zinc (but not copper) to yield the corresponding chloride, e.g., zinc chloride, solution plus hydrogen gas, (6) with some special oxides, e.g., lead or manganese dioxide, to yield lead or manganese chloride plus chlorine gas, (7) with solution of some salts, e.g., silver nitrate, to yield the corresponding chloride, silver chloride, precipitate. Higher strengths of hydrochloric acid usually react similarly to the dilute. Hydrochloric acid sometimes reacts as a reducing acid, e.g., (6) above.

All metallic chlorides, except silver chloride and mercurous chloride, are soluble in H_2O, but lead chloride, cuprous chloride and thallium chloride are only slightly soluble. Metallic chlorides when heated melt, and volatilize or decompose, e.g., sodium chloride, mp 804°C; calcium, strontium, barium chloride volatilize at red heat; magnesium chloride crystals yield magnesium oxide residue and hydrogen chloride; cupric chloride yields cuprous chloride and chlorine. See also **Chlorine; Chlorinated Organics; Halides; Hypochlorites;** and **Sodium Chloride.**

Hydrogen Chloride. This is a colorless gas, heavier than air, density 1.639 g/l at standard conditions. The gas is poisonous and quickly causes suffocation. Formula weight 36.47, mp −111°C, bp −85°C, critical pressure 83 atm, critical temperature 51.3°C. The gas is very soluble in H_2O, accounting for the high concentrations of hydrochloric acid obtainable. Although hydrogen chloride gas may be used directly in some industrial operations, normally it is generated for the purpose of dissolving in H_2O to form hydrochloric acid. The most common route to HCl is by reacting sodium chloride with H_2SO_4. This is a two-step, exothermic reaction: (1) $NaCl + H_2SO_4 \rightarrow NaHSO_4 + HCl$, and (2) $NaCl + NaHSO_4 \rightarrow Na_2SO_4 + HCl$. Preparation of hydrochloric acid from the gas involves an absorption tower where the gas meets a fine spray of H_2O. Ratio controllers are used to assure maximum yield of the acid of desired concentration. These controls are easily adjusted for obtaining different concentrations. In most chlorinations of organic compounds, only half of the chlorine is used to substitute for hydrogen atoms, the remaining chlorine forming HCl. Frequently, this by-product HCl is recycled or recovered.

HYDROELECTRIC POWER. In a hydroelectric power plant, advantage is taken of the gravitational energy available from water flowing from a higher level to a lower level. In seeking a lower level, the flowing water is directed to exit through a hydraulic turbine, which in turn drives an electric generator. A substantial portion of the world's electric power is derived from hydro facilities, particularly in some countries. In general, however, the hydro percentage of total power generation has been declining over the past several years, notably for economic reasons. The technology of hydro power has been developed to a very advanced state.

Classification of Hydroelectric Plants

Low-, Medium-, and High-Head Facilities. There is no definite line of demarcation between high, medium, and low hydraulic heads. Generally, a plant with a head of more than 500 feet (152 m) can be considered a high-head development; a plant with a head of 50 feet (15 m) is definitely in the low-head class. See Figs. 1 and 2.

Briefly, the characteristics of the low-head plant are: vertical, reaction type, runners using large volumes of water and requiring large water passages. Substructure is both extensive and expensive, and intake works are large and complicated. Large diameter generators are made necessary by the low rotational speeds. Characteristics of the high-head plant are: horizontal impulse turbines, small volumes of water at high pressures, plant at some distance from the dam. The advantage of smaller and simpler substructure is offset by the presence of a long water conduit, or penstock, between dam and plant. The turbines

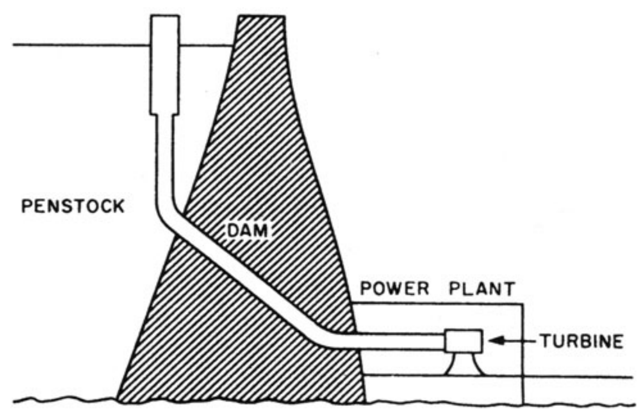


Fig. 1. River plant (high head). Typified by the Hoover Dam and Niagara River installations.

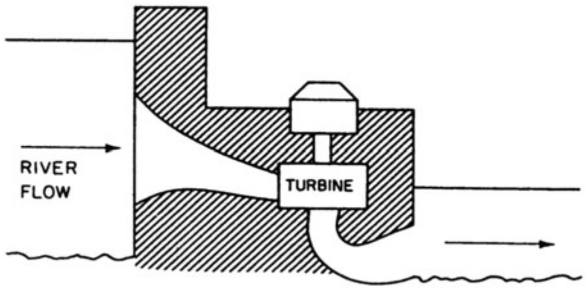


Fig. 2. River plant (low head). Typified by the Bonneville and St. Lawrence Waterway installations.

are high-speed and allow smaller generator diameter. The high-speed is accounted for by the high heads used. Inherently, the impulse turbine has a low characteristic speed.

Impounded Volume. The possible hydroelectric development sites along the flow of a stream are of two types, namely, those suitable for run-of-the-river plants (see Fig. 3) and those offering natural impounding basins for storage plants. In general, the run-of-the-river plant is cheaper than the storage plant of equal capacity, but it suffers seasonal variation of output more or less proportional to the variation of stream flow.

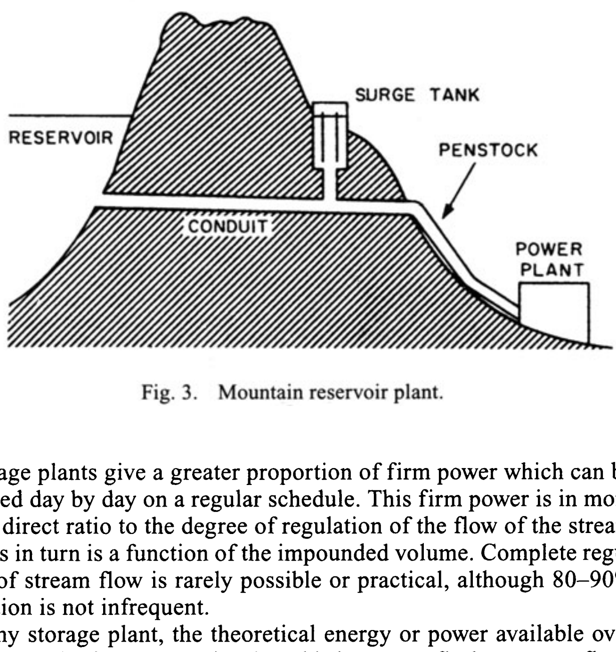


Fig. 3. Mountain reservoir plant.

Storage plants give a greater proportion of firm power which can be delivered day by day on a regular schedule. This firm power is in more or less direct ratio to the degree of regulation of the flow of the stream and this in turn is a function of the impounded volume. Complete regulation of stream flow is rarely possible or practical, although 80–90% regulation is not infrequent.

In any storage plant, the theoretical energy or power available over and beyond the firm power developed is known as flash power or flood peak power. Firm power commands, commercially, a considerably higher rate than flash power.

When integrated with steam generating plants, hydroplants are frequently used to give peak power outputs to take care of peak load conditions and thus avoid the expensive standby service of additional steam generating equipment. Such service, of course, may still permit the delivery of a certain amount of firm power.

If all the run-of-river plants were located upstream from the storage plants they would be operated continuously on a base load plan, because, were they idle, their small reservoirs would quickly overflow and water would be wasted over the crest gates. If, however, they are located between storage plants, the run of the river, as far as they are concerned, is just what the storage plants are passing on to them. Thus, located downstream from a storage plant, a run-of-river plant will produce an increase in output when the storage plant increases its output.

In the hydroelectric plant the turbines and generators are the main items of equipment. The hydroelectric superstructure, as usually laid out, has one large building housing the main units and an electrical bay, or wing, of one or more stories in which are located the switching equipment, offices, storerooms, and most of the auxiliary equipment.

Hydro sites that are developed to use but part of the normal stream flow are exceptions to the general rule. Only rarely is a development made where conservation of the water and its use in the most efficient manner are not paramount features of operation. Failure to give due cognizance to this feature may wipe out the net operating profit; hence a continuous, watchful scrutiny of all natural factors which can affect the station operation is a duty of the operating personnel.

A hydraulic turbine suffers loss of efficiency at heads above or below the designed value because of shock losses. At the correct head there will be one point of best efficiency, somewhere between 80–95% of full load. When a number of units are installed in a plant, and when steam reserve is available, it is generally possible to operate the units near the point of best efficiency. There are four faults of operation and maintenance which can reduce the maximum energy production of a plant:

1. Waste of water over spillways.
2. Improper distribution of the load between the station units.
3. Water leakage through valves, gates, dam or flow line.
4. Wear on moving parts, especially corrosion or erosion of the runner.

The relative simplicity of hydroelectric equipment makes hydraulic efficiency of the turbine the principal consideration.

Combined hydro and steam power plants, designed for use as pumped-storage plants are described later.

Hydraulic Turbines

The fundamentals of the turbine were incorporated into the wheels built before the turn of the nineteenth century, but its principal development has occurred since that time. Beginning with Fourenyon and his outward flow turbine, Jonval, Boyden, Swain, and Francis rapidly brought the reaction turbine to an advanced stage of development. By 1875 the inward flow turbine, as perfected by Francis, and which now bears his name, had established itself in the lead, a position which it maintained until about 1900, when the impulse, or Pelton, type of wheel had progressed to the point of dominating the high head field.

The inherent slow speed of the Francis-type runner on low heads was a fault that the propeller-type runner was designed to cure. During the decade 1910–1920 progress was made with this type of wheel, and by 1920 the propeller-type runner, often called the Nagler runner, was definitely established in the hydroelectric field. Later it was arranged so that the blades could be adjusted and set at different angles to accommodate changes in elevation of the forebay level without undue loss of efficiency. The success of the propeller-type turbine encouraged American adoption of the Kaplan turbine, on which the blade adjustment is performed automatically, being under the same control as the turbine gates.

As between impulse and reaction types, the action in the impulse turbine is easiest to understand. There is no difficulty in visualizing the transformation of pressure head into velocity head at the nozzle, nor of understanding the push, or impulse, that is given to the buckets by the stream of water. The jet is directed upon the rotor tangentially, and hence this type is also called the tangential turbine. The velocity of the jet of water is only slightly less than the free spouting velocity under the effective head h. Impulse buckets are divided into two halves by a "splitter" and the axial thrusts which would otherwise have to be borne by special bearings are equalized.

The essential difference between the impulse and reaction types is that in the former the entire energy received by the wheel is in the velocity form, while in the latter it may be partially in the velocity form, but is also, in a large measure, still in the pressure form. The reaction of conversion of residual pressure into velocity in the runner is the source of much of the torque delivered to the reaction turbine. If the turbine were blocked stationary and had its gates opened, the water would issue from the turbine as from a nozzle. Now, by removing the blocking, let these nozzles begin to rotate, and the absolute velocity of water leaving them is found to be diminishing, the energy having been absorbed by the runner. At the best speed the final velocity will be just sufficient to enable the water to clear the runner. At this time, the wheel may be absorbing from 90–95% of the energy that the water had in the pressure form just before reaching the turbine gates.

In a Francis turbine, the water flows inward, then downward and into the draft tube. See Fig. 4.

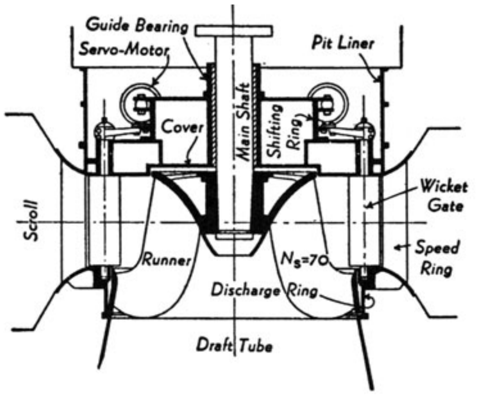


Fig. 4. Cross section showing component parts of a Francis turbine.

A convenient classification of hydraulic turbines is:

1. *Reaction Turbine* (Water under pressure is only partially converted into velocity before it enters the turbine runner.)
   (a) Francis Turbine
   (b) Propeller Turbine
   —1 Fixed-blade
   —2 Adjustable-blade (*Example*: Kaplan turbine)
   —3 Axial-flow (*Example*: Dariaz turbine)
   —4 Diagonal-flow
2. *Impulse Turbine* (*Example*: Pelton Wheel) (Water under pressure is entirely converted into velocity before it enters the turbine runner.)

The Francis turbine is rarely a horizontal shaft machine, except in small sizes and where it is desired to avoid the expense of excavation for a vertical setting. The standard runner consists of two crowns between which the buckets or blades are placed. It is best adapted to vertical setting. In order to pass the large discharges possible in a high specific speed wheel, the buckets are curved downward. Some axial flow action is present in runners of high specific speed. Water is admitted to the runner through guide vanes and gates.

Loss of efficiency at part load is sometimes a serious fault as, for instance, where only one or two units are installed in an isolated plant. The feature of the Kaplan turbine is that the blade angles and gates are adjusted simultaneously by the governor mechanism so that the blades are always in the position best suited for full utilization of the flow, through the reduction of eddying and shock losses. The result is that the efficiency at part load holds up remarkably well.

Conveying the water from the penstock and directing the proper amount of it correctly against the runner requires first, a scroll case; second, a speed ring; and third, turbine gates.

The scroll case for medium and high head development is circular in form. In plan, it leads from the penstock and wraps, in spiral form, around the speed ring. The cross section of the spiral at any point should be such that the water flows with uniform velocity. This leads to the spiral form, because the water is being delivered to the turbine uniformly around the entire circumference.

The speed ring is that part of the turbine which joins the discharge ring with the turbine cover and pit liner. The ribs between the top and bottom portions must be strong enough to support the dead weight above the casing, consisting of concrete, generator, and turbine rotative parts; hence the speed ring is a very important part of the turbine.

Inside the speed ring, and rigidly bolted to it, is the inlet gate mechanism. The mechanism is operated by the governor which, by opening or closing the gates, can maintain a control of speed under variable load. Gates are of the guide vane type and, while various types of gates have been used, the wicket gate is in general use at the present time. Its principal advantage is its efficiency. Shock losses at part gate opening are reduced to a minimum in the wicket gate. It is not particularly tight and has many wearing parts, most of which are bronze bushed and grease lubricated.

The Pelton wheel is either a solid or open disk, to the rim of which are attached buckets upon which a jet of water is played from a stationary nozzle. A horizontal shaft is the usual arrangement, but vertical shaft units have also been put into operation. The advantage of using the vertical arrangement is that more than one jet can be played on the buckets; this is obtained, however, at the expense of some loss of efficiency. The Pelton wheel is overhung on the bearing and often, for additional capacity, two wheels are overhung on the same generator. Variable power demand is met by decreasing the amount of water in the jet, by deflecting the jet from the buckets, or both. Some turbines of this type have a relief jet which opens as the main jet closes. Afterwards, a dash pot slowly closes the relief jet, slowly enough to prevent a large pressure rise in the penstock. The same is also accomplished by deflecting the jet from the wheel upon loss of load, then slowly closing the valve controlling the jet.

Draft Tube. Hydraulic turbines frequently discharge the water with considerably more velocity than would be economical from the efficiency viewpoint, were it not possible to recover a great deal of that energy by the proper use of a diffusing chamber at the outlet. The diffusing chamber or tube is known as the draft tube, and there are a variety of types. However, the main objective is to convert the velocity head residing in the water leaving the turbine into pressure head. If this can be done efficiently, the turbine can be set somewhat below normal tailwater level.

The greater the specific speed of a turbine runner the higher will be the velocity of the water discharged into the draft tube, and the more important the recovery of this velocity by draft tube design. The draft tube is to take the water from the turbine at a point where the pressure is considerably less than atmospheric, and, by efficiently reducing the velocity, convert it into pressure head so that it can emerge smoothly into the tailrace at atmospheric pressure. By "efficiently" is meant without shock or whirl loss. Not all the velocity head can be recovered, for the water must be given to the tailrace at normal tailrace velocity to prevent its backing up into the turbine. Also, whatever friction loss occurs in the draft tube adds to this reduction of useful head.

Typical turbine efficiency curves are shown in Fig. 5. A schematic diagram of a typical hydro governor system is shown in Fig. 6.

Pumped-Storage Plants

Growing emphasis over the past couple of decades has been placed upon the use of special hydro plants as a means of storing energy in the

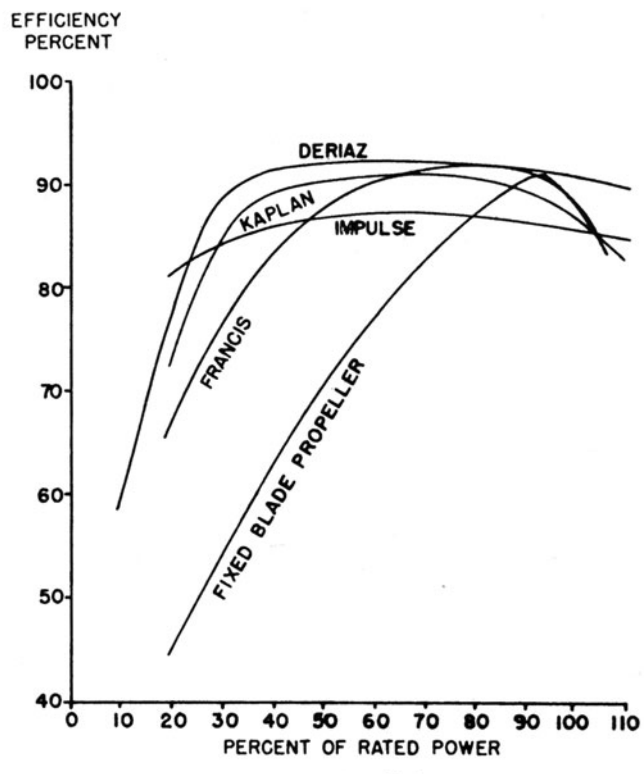


Fig. 5. Hydraulic turbine efficiency curves.

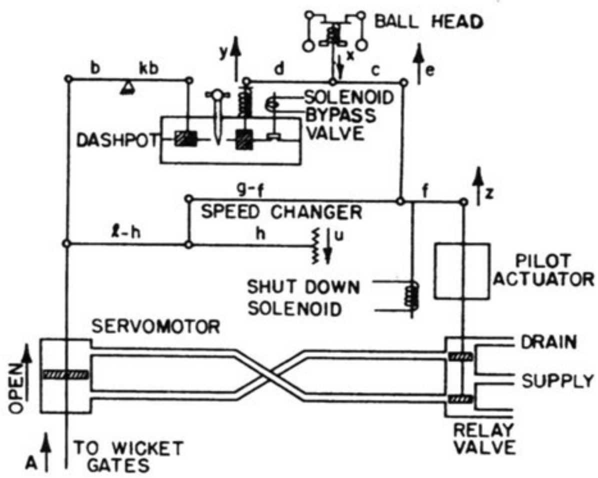


Fig. 6. Typical hydro governor system.

form of a head of water—pumped into an upper reservoir during off-peak hours. The history of pumped-storage plants dates back to the late 1920s when several plants were first installed in Europe. The Rocky River plant of Connecticut Light and Power Company was the first to be built in the United States during that early period. From the viewpoint of plant location, there are three categories of pumped-storage installations:

1. *Combined with conventional hydro plant*. Plants of this type are used in locations suitable for conventional hydro plants, but where rainfall or water availability and system demand are out of phase. For example, in Switzerland, demand is highest in winter when water is scarce and lowest in summer where there is an abundance of water. The available energy in summer can be used to fill the reservoirs and store the available water for later use in meeting the winter demand.

2. *Pure pumped storage*. The advantages of this type of plant are its flexibility of location, in that the upper reservoir need have no source of water other than what is pumped into it, and the possibility of developing large plants with a small reservoir and high head. This type of plant is commonly used in steam-based systems which lack the many advantages of available hydro generation. As well as providing fast and reliable peaking power, pumped-storage units have the added advantage of smoothing the weekly load curve and enabling more of the efficient base-loaded steam plants to be operated continuously.

3. *Pumped storage with diversion*. This situation arises when available water must be shared between power generation and irrigation use. Water which must be pumped to a higher reservoir to feed an irrigation canal can be used as a source of peaking power if allowed to run back down through a pump-turbine.

Design Configurations. The three fundamental configurations include: (1) separate pump and turbine on the same shaft; (2) reversible pump-turbines; and (3) axial-flow units.

The turbines used in the first class are Francis type for heads in the range 100 to 1,000 feet (30.5 to 305 meters) and Pelton wheels for heads up to 3,000 feet (914 meters). These units are usually mounted on a horizontal shaft with a clutch (hydraulic or friction) between the motor-generator and the pump. The turbine is usually rigidly connected to the shaft and is dewatered, using compressed air during pumping. Sometimes small impulse turbines are installed on the shaft for starting and braking.

Reversible pump-turbines are of radial or mixed flow type—Francis or Deriaz—and have been designed to operate at a wide range of heads.

Axial-flow units are designed to operate at low heads, around 20 feet (6 meters), and have adjustable blades similar to those of a Kaplan turbine. The bulb type is used in Europe and the tube type has been designed and built in the United States. Both can operate as turbines or as pumps in both directions of flow by reversing the pitch of the blades. A 9-megawatt bulb unit was installed at Saint Malo, France, and in the Rance tidal project near Saint Malo, there are twenty-four 10-megawatt bulb units. These are described in entry on **Tidal Energy.**

Hydroelectric Power in the United States

In the 1930s, hydroelectric power furnished almost 40% of the electric energy needs of the United States. Since the 1950s, hydropower has grown less rapidly than other forms of electricity generation. Total hydropower generated by hydro plants in the early 1990s represents 10–13% of the nation's needs. Total output is estimated at about 70 GW (gigawatts). Construction of new hydro facilities has decreased for several reasons, notable of which are very high construction costs and environmental factors. Although pure water is the fuel entering a hydro plant and essentially pure water exits the facility and there is no release of carbon dixoide, nitrogen oxides, sulfur dioxide, and so on into the atmosphere, objections have been raised pertaining to the amount of land (and aesthetics) that must be sacrificed for a hydro plant and the large dam and reservoir needed. Another objection is the manner in which some water species are threatened. Another important factor, of course, is that many of the excellent sites are already in the hydropower network.

Large hydro installations in the United States include Grand Coulee (Washington), 6.5 GW; John Day (Oregon), 2.2 GW; and Chief Joseph (Washington), 2.1 GW. Only Grand Coulee ranks among the top ten largest installed capacity hydro plants in the world. Throughout the United States, there are over 40 hydro plants. The U.S. Corps of Engineers has estimated that the ultimate potential for hydro power in the United States is over a half-million GW, with the possible development of over 10,000 different sites. Realistically, however, hydro power probably will not exceed 75 to 105 GW by the year 2020. Approximately 85% of the potential for new plant sites are located in the western states.

Hydroelectric Power in Canada

In contrast with the United States, 57 percent of Canada's electric power is furnished by hydro plants. Approximately one-half of this energy comes from hydro plants in Quebec. Less than one-third of Canada's power is generated by burning fossil fuels. British Colombia, Ontario, Newfoundland, and Manitoba also have extensive hydro facilities.

Historically, Canada had depended upon hydro plants for many years. See Fig. 7. Nevertheless, planners in Canada expect that hydropower as a percentage of total power generated will decline gradually over the next several years because of several factors:

1. Most of the better sites have been used.
2. The growth rates of real fossil fuel prices were negative between 1950 and 1973, which favored the construction of thermal, including nuclear, facilities during that period.

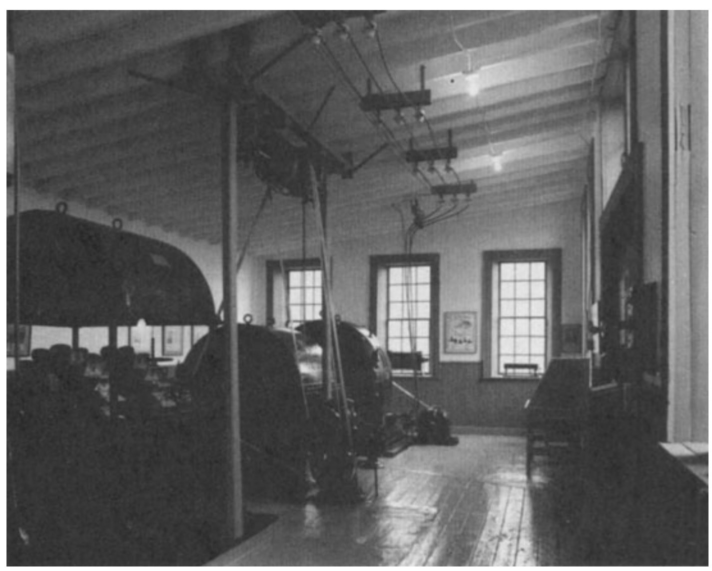


Fig. 7. The first hydroelectric development in Newfoundland was built at Petty Harbour, and electricity was transmitted to St. John's for the first time on April 19, 1900. The original equipment consisted of a Pelton wheel turbine, which was connected to a 250kVA General Electric generator. In 1907, a second Pelton wheel turbine and a generator similar in size to the original were installed. To meet increased demand and to provide for further expansion of the system, the building was extended in 1914 and a Voith turbine and a 500 kVA Westinghouse generator were installed. (*Courtesy, Newfoundland Light & Power Co. Limited.*)

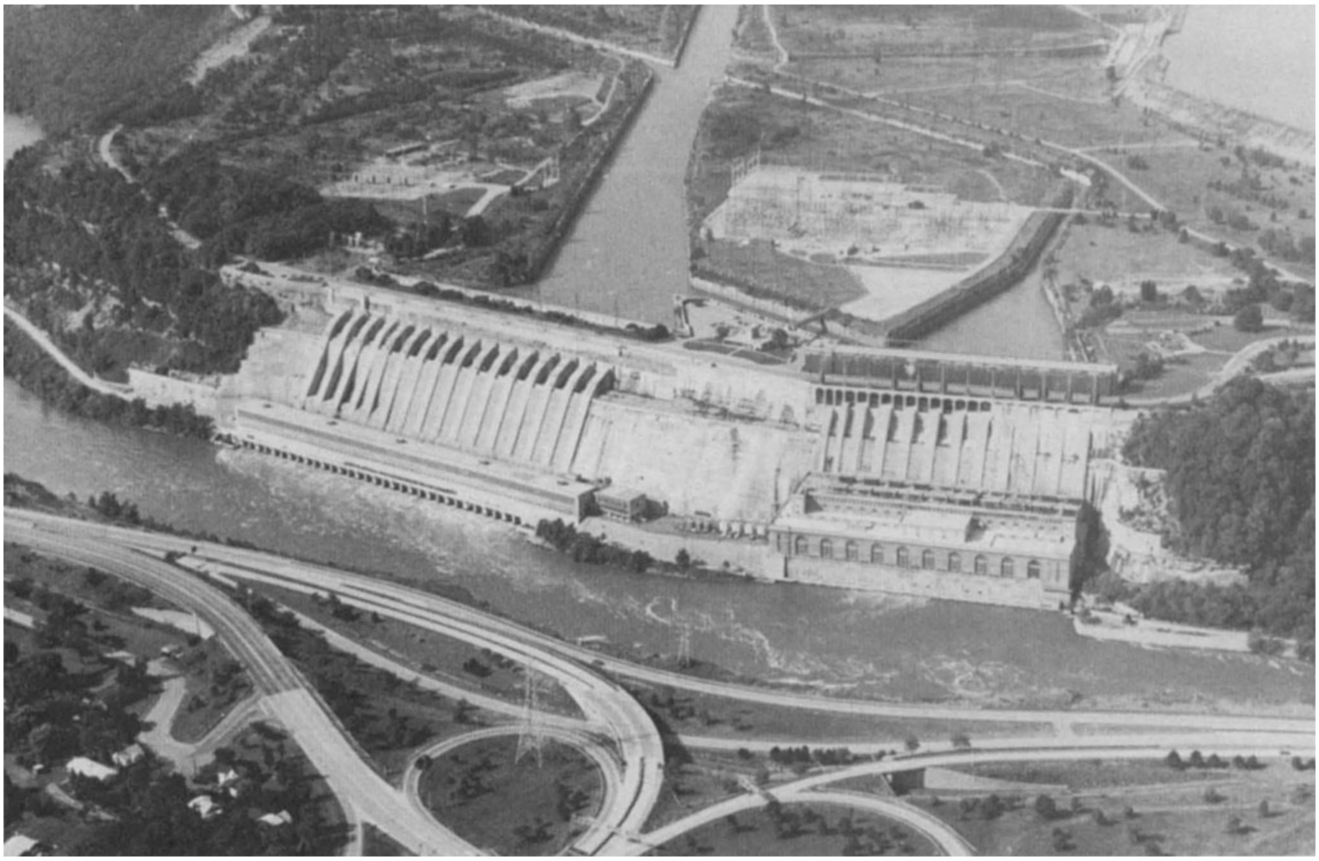


Fig. 8. Called the Queenston-Chippaway development when construction began in 1917, the Sir Adam Beck-Niagara Generating Station No. 1, shown at the right of the view, was for many years the largest hydroelectric plant in the world. To utilize the maximum fall of the river, water had to be diverted from an intake 2 miles (3.2 km) above the Horseshoe Falls to the plant at the base of the Niagara Gorge. To accomplish this, engineers built an open canal waterway some $12\frac{1}{2}$ miles (20 km) long from Chippaway across country to a triangular basin called a *forebay* on the escarpment more than 295 feet (90 m) above the Niagara River. From the forebay, giant *penstocks* (tubes) carry the water to the powerhouse below, located on the river's edge. The powerhouse, accommodating the generating, transforming, and controlling equipment, rises more than halfway up the cliff to a height of 180 feet (55 m) above the foundation. With a total of ten generating units, the plant has an installed capacity of 414,650 kW.

Construction of the Sir Adam Beck Generating Station No. 2, shown at the left of view, began in 1951 beside the first plant. For the second plant, it no longer was feasible to interrupt surface traffic to build another open canal. Instead, two underground tunnels were built to carry 18 million gallons (68 million liters) of water per minute from Chippaway to the forebay. With a finished diameter of 46 feet (14 m), the parallel tunnels are 5.6 miles (9 km) long and pass directly under the City of Niagara Falls, Ontario at a depth of 331 feet (101 m). Opened in 1954, the Beck No. 2 station houses 16 generating units and is almost twice as long as the Beck No. 1 station. The more recent plant has an installed capacity of 1,223,600 kW.

To accommodate for peak loads on the Beck No. 2 station, a pumped-storage reservoir, one corner of which is shown at the extreme upper righthand corner of view, was created. A separate pumping-generating station, containing 6 generators, adds 176,700 kW to the total capacity of the Beck No. 2 station. When the pumps are reversed, they act as turbogenerators. (*Ontario Hydro.*)

3. A planned development of nuclear energy as an alternative source for future energy demand.

Canada has two hydro plants that are among the ten largest hydro facilities in the world: La Grande 2, Quebec (5.3 GW); and Churchill Falls, Newfoundland (5.2 GW). The development of power, jointly by Canada and the U.S., along the St. Lawrence River is exemplary of international cooperation. The total installed capacity, involving 3 dams and 16 miles (26 km) of dykes, utilize the drop in water level between Lake Ontario and the powerhouses 125 miles (201 km) downstream. The main dam and powerhouses form a continuous structure some 3300 feet (1006 m) long. Generators, totaling 32 in number with a total capacity of 1.8 GW are not housed in conventional structures, but are protected by removable hatch covers. Generators on the Canadian side of the river feed into Ontario Hydro's grid system. Their capacity totals more than 0.9 GW, equal to the needs of about 600,000 homes. The Canadian power station (Robert H. Saunders) is located at Cornwall, Ontario. The main U.S. counterpart is located at the Robert Moses Power Dam. Flooding of the huge headpond area called for vast removal of property. Homes and even cemeteries were relocated. Ontario Hydro built new shopping centers, schools, churches, roads, sidewalks, waterworks, sewage treatment plants, and recreation areas for the 6500 persons displaced. Only farm families and cottage owners were involved on the sparsely populated American side of the project. Both Canadian and U.S. stations were opened in 1958. The headpond area is estimated at 100 sq miles (259,000 sq km) and the watershed area at nearly 300,000 sq miles (777,000 sq km). See Figs. 8, 9, and 10.

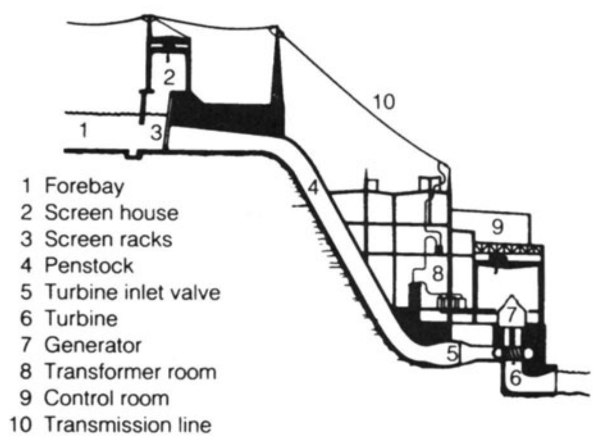


Fig. 9. Sectional view of Sir Adam Beck generating station. (*Ontario Hydro.*)

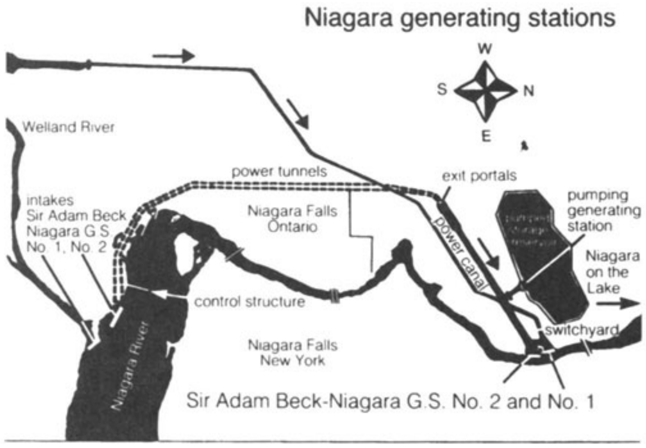


Fig. 10 Detail of site of the Sir Adam Beck Niagara generating stations Nos. 1 and 2. (*Ontario Hydro.*)

Hydropower Worldwide

Nations other than Canada and the United States having hydroelectric power plants with installed capacities in excess of 1 GW include: Brazil, Paraguay, Venezuela, Argentina, Russia, Mozambique, China, and Mexico. Spain, Italy, and France, the individual facilities of which are smaller, still generate an impressive percentage of their total electric production with hydro plants: Spain, 58.5%; Italy, 42.2%; and France, 32%. (The French hydro percent, however, is decreasing because of that country's emphasis on nuclear facilities in recent years.) Hydroelectric production in Russia is estimated between 20 and 22% of the total electric generating capacity. Worldwide, especially in some of the naturally favored underdeveloped nations, much potential for hydropower remains unexploited.

The relative advantages and limitations of hydroelectric power are summarized in the accompanying table.

RELATIVE ADVANTAGES AND LIMITATIONS
OF HYDROELECTRIC POWER INSTALLATIONS

ADVANTAGES

- Continuous low-cost power production except when droughts occur.
- Low maintenance costs.
- No consumption of irreplaceable fossil fuel
- No air pollution
- Reservoir lakes can be used for recreation in majority, but not in all cases.
- Reservoirs can provide considerable, but not complete flood protection to downstream areas
- Reservoirs are capable of storing large quantities of water for long periods of time, but not indefinitely.
- Downstream flow can be managed to aid in water-quality control and to level out the extremes of winter versus summer stream conditions.
- Ground-water reserves are increased by recharging from the reservoir.

LIMITATIONS

- High initial cost of construction.
- Recreational facilities can be adversely affected in reservoirs where draw down in the dry season lowers the water level.
- Flood protection can best be provided by an *empty* reservoir, while power production is best from a full reservoir. A *full* reservoir cannot retain a major flood, an empty reservoir generates no power. The compromise then, is to retain enough water in a reservoir to insure continuous power generation, but leave a margin of free board to take the major surges out of a sudden torrential rain storm
- Loss of land suitable for agriculture.
- Power production may be curtailed or even discontinued in time of drought.
- Original stream valley is inundated.
- Some water is lost by evaporation from the reservoir surface.
- In coastal areas, such as Oregon and Washington, the construction of dams prohibits the upstream migration of anadromous fish, such as the Pacific Salmon, unless some arrangements, such as a "fish ladder" is provided.

SOURCE: Battelle Memorial Institute, Columbus, Ohio.

Additional Reading

In addition to specific references indicated below, statistics on hydroelectric power are continuously updated and available from such organizations as: U.S. Federal Power Commission, U.S. Army Corps of Engineers, Bureau of Reclamation, U.S. Department of the Interior—all in Washington, D.C. Also, Edison Electric Institute, New York; and Electric Power Research Institute, Palo Alto, California.

In Canada, additional information can be obtained from Energy, Mines and Resources, Canada, Information Services, Ottawa, Ontario; and from Ontario Hydro, Corporate Communications, Toronto, Ontario.

HYDROFLUORIC ACID. HF (hydrogen fluoride gas) in aqueous solution, colorless when pure, fuming (dependent on concentration), highly corrosive, extremely reactive, available commercially in 30, 52, 60, and 80% HF concentrations. There is a maximum constant boiling point 111°C (750 torr) at 43% HF (distillate) for mixtures of HF and water. Because HF attacks glass and many other container materials, the laboratory HF reagent is packaged in polyethylene bottles or carboys. Larger containers for industrial use usually are steel drums or tanks with a polyethylene lining. Anhydrous HF is available in tank cars of 22- or 42-ton (20- or 38-metric ton) capacity, as well as in steel cylinders of 100- or 200-pound (45- or 90-kilogram) capacity.

The formula weight of HF is 20.01 (calculated). However, its apparent molecular weight ranges widely with temperature and pressure. The molecular weight of saturated HF vapor at 19.51°C is 78.24; at 100°C, the value is 49.08. Because of strong hydrogen bonding between molecules, significant polymerization occurs, thus resulting in marked departures from ideal behavior, both in the gaseous and liquid phases. The polymerization mechanism has not been fully determined, but both ring- and chain-type structures have been suggested. Of interest is the comparison of the high boiling point of HF (+19.5°C) with the boiling points of other acids in the halogen series: HCl, −85°C; HBr, −65°C; HI, −36°C. Because of this polymerization, the formula H_2F_2 often has been used for hydrogen fluoride, although the polymers in the gas appear to be chiefly $(HF)_6$. Evidence of the stability of these hydrogen bonds is furnished by the existence of the hydrogen fluoride ion (HF_2^-) in ionic crystals and acid fluoride solutions.

Liquid hydrogen fluoride is one of the three binary hydrides (the others are H_2O and NH_3) which are self-ionized and highly associated (Trouton constant 26.6, bp 19.54°C, mp −83.7°C). It has a dielectric constant of 83.6 at 0°C and is an excellent ionizing solvent. Because of its very high acidity, most oxygen-containing substances are protonated in solution in HF, forming substituted oxonium ions or oxonium ion itself by solvolysis. It has a very low viscosity—0.256 centipoises at 0°C. Its surface tension is also exceptionally low. Its density at the boiling point is 0.991 g/ml.

The effects of hydrogen fluoride on glass were observed by A. S. Marggraf in 1764. In 1771, Scheele established that a new acid had been discovered. In 1814, Davy showed that the acid contained a newly found element, fluorine. Fluorine was not isolated until 1886 by Moissan. The first anhydrous HF was not prepared until 1856 by Frémy. The first commercial shipment of HF (anhydrous acid) was not made until 1936.

Because of the strong affinity of HF for H_2O, there is no known chemical substance that can be used for drying it. HF immediately reacts on complexes with the drying agents. Thus, the compound can be dehydrated only by electrolysis. Even though HF is highly reactive, it has the characteristics of a weak acid, due to the extensive polymerization of the HF.

Anhydrous hydrogen fluoride is prepared commercially by reacting calcium fluoride (acid-grade fluorspar) with concentrated H_2SO_4 in a heated reactor. The presence of silica in the calcium fluoride is highly objectionable inasmuch as each pound of silica present will consume 2.6 lb of CaF_2 to form silicon tetrafluoride. When the latter compound is absorbed with HF in H_2O, fluosilicic acid is formed, representing a further loss of net HF produced: (1) $CaF_2 + H_2SO_4 \rightarrow CaSO_4 + 2HF$; (2) $4HF + SiO_2 \rightarrow SiF_4 + 2H_2O$; (3) $2HF + SiF_4 \rightarrow H_2SiF_6$. After reacting at a temperature of 200–250°C, the HF is treated to remove dust and H_2SO_4 fumes and then condensed as 99% HF.

Uses: Principal uses for HF include: (1) the production of aluminum fluoride and synthetic cryolite required for aluminum production, (2) the production of fluorinated organics of several types and for several applications, including aerosol propellants, special-purpose solvents, refrigerants, and plastics (polytetrafluoroethylene, polyvinylidene fluoride, polychlorotrifluoroethylene), (3) in the formulation of atomic-energy feed materials, (4) as an alkylation catalyst in petroleum processing, (5) as a pickling acid in stainless-steel and nonferrous metals manufacture, (6) as an agent for etching and polishing glass, (7) as a reactant in several organic syntheses, (8) in the manufacture of elemental fluorine, and (9) as a starting material for the preparation of fluorides and fluoborates.

Although the manufacture of atmosphere-damaging cholorofluorocarbons in the past has represented a high-tonnage requirement for hydrofluoric acid, the newer, less-polluting hydrochlorofluorocarbons will require an even larger supply of hydrofluoric acid. A new manufacturing facility was opened in Coahuila, Mexico, near Corpus Christi, Texas, in 1992.

Additional Reading

Meyers, R. A.: "Handbook of Chemicals Production Processes," McGraw-Hill, New York, 1986.

Sax, N. I.: "Dangerous Properties of Industrial Materials," Van Nostrand Reinhold, New York. (Updated periodically.)

Staff: "World-Wide hazardous Chemicals and Pollutants," Forum for Scientific Excellence, American Chemical Society, Washington, D.C., 1990.

Staff: "Handbook of Chemistry and Physics," 73rd Edition, CRC Press, Boco Raton, Florida, 1992–1993.

Staff: "Hydrofluoric Acid (Anhydrous and Aqueous)," Chemical Safety Data Sheet, Manufacturing Chemists Association, Inc., Washington, D.C. (Updated periodically.)

Welch, J. T., Editor: "Selective Fluorination in Organic and Bioorganic Chemistry," American Chemical Society, Washington, D.C., 1991.

HYDROFOIL. A watercraft equipped with wing-like transverse surfaces suspended below the hull. As the craft moves forward, hydrodynamic forces are produced, just as aerodynamic forces are produced in air, to give lift. At the start, the vessel operates as a normal displacement vessel, but as the speed increases the lift on the hydrofoils increases to raise the craft out of the water, thereby decreasing the water resistance to that of the hydrofoils alone. Higher speeds can be obtained, since the water resistance for the required lift is considerably less for the gear still submerged in the water. To be most effective, propulsive units, such as engine-propeller combinations, or jet engines operating well above the water, are necessary.

Hydrofoil craft are of a general class of designs collectively referred to as *interface vehicles*.

HYDROGEN. Chemical element symbol H, at. no. 1, at. wt. 1,008, periodic table group 1, mp −259.14°C, bp −252.87°C, density 0.089 (solid at 4.2K), 0.071 (liquid at 20.4K), sp gr 0.0696 (air = 1,0000). Solid hydrogen has a hexagonal crystal structure. Hydrogen at standard conditions is a colorless, odorless, tasteless gas, suffocating, but not toxic. Hydrogen occurs chiefly combined with oxygen in H_2O, with carbon in hydrocarbons, with carbon and oxygen, and with carbon and several other elements, including oxygen, nitrogen, sulfur, phosphorus, and most metals in a vast variety of hundreds of thousands of organic compounds. See also **Organic Chemistry.** Hydrogen is considered by some scientists as the primordial substance from which all other elements in the universe were developed. In terms of cosmic abundance, with a rating of silicon = 10,000, it has been estimated that the figure for hydrogen is about 3.5×10^8, this figure compared with that of carbon = 80,000, nitrogen = 160,000, and oxygen = 220,000. For further comparison, the figure for gold is 0.0015 and for uranium it is 0.0002. In terms of abundance of the chemical elements in seawater, hydrogen ranks second (behind oxygen) with an estimated 510 million tons per cubic mile (~109 million metric tons per cubic kilometer). Hydrogen ranks eleventh in terms of content in igneous rocks in the earth's crust, the estimate of average content being 0.13%. Although free hydrogen escaped from the earth's lower atmosphere, some of the planets appear to have significant amounts, including the atmospheres of Jupiter, Saturn, and Uranus. At an altitude of 1,000 miles (1609 kilometers) above the surface of the earth, there is a greater abundance of hydrogen atoms than of nitrogen or oxygen atoms.

Hydrogen was first identified by Cavendish in 1766. The element was named by Lavoisier in 1783. However, it was not until 1931 that a second isotope of hydrogen (deuterium) with a mass number 2 was discovered by Urey. In 1934, Rutherford, Oliphant, and Harteck prepared a third isotope (tritium) with a mass number 3. Normal hydrogen (protium) and deuterium are stable, whereas tritium is radioactive, with a half-life of 12.26 years. Tritium emits a negative electron to form ^{-3}H. It is estimated that the isotopic abundance of ^1H (protium) in natural occurring hydrogen is 99.9851% and on the basis of carbon = 12 (atomic weight scale), protium has a mass of 1.007825 amu. The isotopic abundance of ^2H (deuterium) is estimated at 0.0149% with a mass of 2.014101 amu. The artificially-prepared ^3H (tritium), $^9Be + {}^2H \rightarrow 2\,{}^4He + {}^3H$, has a mass of 3.01605 amu. Heavy water is deuterium oxide, ^2H_2O, usually written D_2O. Deuterium and deuterium oxide gained prominence largely because of their excellent properties as moderators in nuclear reactors. The ionization potential of hydrogen is 13.59765 ± 0.00022 eV. Other physical properties of hydrogen are given under **Chemical Elements.** See also **Deuteron; Deuterium.**

When ignited, hydrogen burns in air with a pale blue to colorless, nonluminous flame, yielding H_2O. When mixed with air, the flammability limit is 4–74% hydrogen. When mixed with oxygen, the flammability limit is 4–94% hydrogen. Care always must be exercised where there may be hydrogen mixtures with air or oxygen because violent explosions may occur. In sunlight or magnesium light, hydrogen combines with chlorine with violent release of energy, forming hydrogen chloride HCl. When hydrogen is heated with sodium, calcium, and several other metals, the corresponding hydride is formed. In the presence of a catalyst, hydrogen reacts with nitrogen to form ammonia NH_3. Upon heating sulfur in the presence of hydrogen, hydrogen sulfide, H_2S, is formed. At elevated temperatures, hydrogen will reduce many of the metal oxides to the metal, notably copper, iron, nickel, tin, and lead. The oxides of zinc, aluminum, and magnesium are not so reduced. Hydrogen reacts with unsaturated organic compounds in most cases to form saturated compounds. For example, in the presence of a catalyst, hydrogen will add to oleic acid $C_{17}H_{33}COOH$ to form stearic acid $C_{17}H_{35}COOH$. See also **Hydrogenation.**

Production of Hydrogen. For chemical and petroleum processes, hydrogen is an extremely high-tonnage and one of the most fundamental raw materials. Sources of hydrogen and processes for producing it are described in entry on **Hydrogen (Fuel).**

Uses. In terms of consumption, NH_3 is by far the largest user of hydrogen. Petroleum refining processes and methanol synthesis are the next largest consumers. Hydrogen needs for these uses are almost always fulfilled by hydrogen-generation capacity on the premises. What might be termed commodity hydrogen is shipped from hydrogen plants to various users. Some of the more important uses include the hydrogenation of numerous organic compounds, such as vegetable and animal oils, the oxyhydrogen and atomic-hydrogen welding applications, the reduction of several metallic oxides, such as iron, copper, nickel, cobalt, tungsten, and molybdenum, and the use of liquid hydrogen as a rocket fuel. See also **Ammonia; Hydrogenation; Methyl Alcohol; Petrochemicals;** and **Synthesis Gas.** For the potential role as a fuel, see **Hydrogen (Fuel).**

Ortho- and Para-Hydrogen. On the basis of nuclear spin, two forms of hydrogen are known: *ortho-hydrogen*, in which the two nuclei in the H_2 molecule have parallel spins, and *para-hydrogen*, in which the nuclear spins are anti-parallel. At ordinary temperatures (and above) ortho-hydrogen is present to the extent of about 75%; at lower temperatures, the ortho changes to para-hydrogen, until at very low temperatures, as that of liquid hydrogen, the para form is present to the extent of 99.7%. There is some difference in properties between the two, notably in thermal conductivity.

The transition from ortho- to para-hydrogen releases heat in amount of 168 cal/g. The heat of vaporization of liquid hydrogen is 107 cal/g. Thus, more than ample heat is released to revaporize liquid hydrogen. Knowledge of the existence of the ortho-para transition and the development of catalysts to equilibriate the liquid during liquefaction essentially have made possible the very large-scale manufacture, use, and storage of liquid hydrogen.

Below −220°C the specific heat of hydrogen is that of a monatomic gas like helium (He). Practically pure para-hydrogen may be obtained by adsorption of ordinary hydrogen, which is three-fourths ortho and one-fourth para, on charcoal at about −225°C. The mp of para-hydrogen is 0.13°C lower (ortho-hydrogen 0.04°C higher) than ordinary hydrogen, and the bp at 60 mm pressure is 0.13°C lower (ortho-hydrogen 0.04°C higher) than ordinary hydrogen. Para-hydrogen reverts slowly to ordinary hydrogen, but immediately in the presence of platinized asbestos.

Atomic Hydrogen. At high temperatures, the loss of heat from a glowing wire in hydrogen is larger than expected on regular assumptions. This is believed to be due to dissociation of ordinary hydrogen into atomic hydrogen (H). See accompanying table.

DISSOCIATION OF HYDROGEN

Temperature, °C	Pressure At 760 mm	Pressure At 1 mm
1730	0.33%	8.7%
2230	3.1	57.5
2730	34	99.3

When hydrogen is passed through an electric arc between tungsten poles, a considerable transformation into atomic hydrogen occurs, and when a stream of this gas strikes a surface a large evolution of heat takes place through recombination to ordinary hydrogen. This atomic hydrogen flame is of temperature sufficiently high to melt tungsten (mp 3,370°C). The half-life of the hydrogen atom is one-third second at 0.5 mm pressure. This reaction is endothermic, values of 98–105 kcal per mole having been reported for it. It is an active reducing agent, reducing many metallic oxides and halides to the free metals, and forming hydrides with many nonmetals. The energy of its exothermic recombination is utilized, in combination with the energy released by the oxidation of the H_2 formed, by atmospheric oxygen, in the oxyhydrogen welding process.

Ionization. The ionization potential of hydrogen is 13.59765 ± .00022 eV, and the ionization process (in the case of protium) yields an electron and a free proton. The electric field of the proton is strong, due to its small radius, so that it readily combines with polarizable atoms. Thus, in aqueous solution, it shares an unshared pair of electrons of the oxygen atom of H_2O to form H_3O^+, the hydronium ion; with NH_3 it forms NH_4^+, the ammonium ion; with phosphine it forms the phosphonium ion, PH_4^+, etc. The hydrogen atom can also add an electron, to form the hydride anion, H^-, this potential (electron affinity) being only about 0.7 eV. Hydride ions have been shown (by electrolysis, crystal structure, etc.) to exist in the hydrides which hydrogen forms with the alkali metals and some of the other metals on the left side of the periodic table. While most other hydrogen compounds are essentially covalent, the binary compounds with the halogens and some of the other elements on the right side of the periodic table exhibit a considerable degree of ionicity, varying considerably in the same group.

The hydrogen atoms in many compounds tend to be shared between the electronegative atom or group to which they are attached and similar groups on other molecules. These hydrogen bonds increase the intermolecular forces and boiling points of hydrogen fluoride, water, organic acids and alcohols, etc. A descriptive explanation of the process is the positive polarity of the H atom that is attached to the electronegative atom or group, which gives it an effective coordination number of 2, so that it can attract an unshared electron pair of a fluorine, oxygen, nitrogen, atom of another molecule. The atom having the unshared pair must be negatively polarized or easily polarizable. For example, tertiary arsines form stronger hydrogen bonds with phenols than do tertiary phosphines.

A number of hydrogen compounds ionize to yield solvated protons, i.e., $2H_2O \leftrightharpoons OH_3^+ + OH^-$, and $2NH_3$ (liq.) $\rightleftarrows NH_4^+ + NH_2^-$. Moreover, many hydrogen compounds, when dissolved in such solvents, ionize more or less completely to give solvated protons and anions. In the case of polybasic acids, ionization constants are reported for each step in this dissociation.

Hydrides. See section on hydrides in entry on **Hydrogen (Fuel);** and separate entry on **Hydride.**

Water and Acids. The properties of the most prevailing hydrogen-bearing compound, water, are given under **Water.** The characteristics of acids are attributed essentially to the presence of hydrogen ions. These topics are treated under **Acids and Bases;** and **pH (Hydrogen Ion Concentration).**

Hydrogen Under Extreme Pressure

Interest in the possible existence of hydrogen as a metal is spurred by the prospect that hydrogen may be able to conduct electricity with zero resistance near room temperature and that, because of the tremendous concentration of energy, as contrasted with liquid hydrogen, it could serve as a rocket fuel and high explosive.

In 1989, Mao and Hemley (Geophysical Laboratory, Carnegie Institution of Washington) reported on an investigation of the insulator-metal transformation in solid hydrogen at high pressure. Much earlier, theoretical calculations made by Wigner and Huntington (1935) revealed that the transition may occur in the 250-to-400 GPa (2.5-to-40 megabar) range. With the high-pressure research tools (diamond anvil cell) available today, this transition point of hydrogen has become a primary target for some researchers.

Mao and Hemley (see reference listed) reported that direct optical observations of solid hydrogen at the aforementioned pressure range and at 77K indicated that the hydrogen sample appeared nearly opaque and that optical data were consistent with a band-overlap mechanism of metallization. These findings were later challenged by Silvera (reference listed) to the effect that "Visual darkening of a sample is not sufficient evidence of metallization, just as a lustery metallic reflection is not. Good examples are the semiconductors germanium and silicon, which as thin films, are metallic in appearance. Darkening of a sample can arise from any number of physical mechanisms that cause absorption throughout the visible spectrum." Further justification, however, was given by Mao and Hemley.

In mid-1991, Badding, Hemley, and Mao reported on studies of the high-pressure chemistry of hydrogen in metals and made specific studies of the reaction between iron and hydrogen at sudden pressure-induced expansion at 3.5 gigapascals of iron samples immersed in fluid hydrogen. The investigators mention numerous specific areas of interest that may be addressed with a better understanding of the behavior of hydrogen with metallic environments, including the hydrogen degradation of ferrous metals.

Additional Reading

Badding, J. V., Hemley, R. J., and H. K. Mao: "High-Pressure Chemistry of Hydrogen in Metals: In Situ Study of Iron Hydride," *Science*, 421 (July 26, 1991).
Crawford, M.: "Accelerator Eyes for Warhead Tritium," *Science*, 469 (January 27, 1989).
Giacobbe, F. G., Iaquaniello, and O. Loiacono: "Increase Hydrogen Production," *Hydrocarbon Processing*, 69 (March 1992).
Mao, H. K., and R. J. Hemley: "Optical Studies of Hydrogen Atoms Above 200 Gigapascals: Evidence for Metallization by Band Overlap," *Science*, 1462 (June 23, 1989).
Peterson, I.: "Squeezing Hydrogen to Molecular Metal," *Science News*, 164 (March 17, 1990).
Pool, R.: "The Chase Continues for Metallic Hydrogen," *Science*, 1545 (March 30, 1990).
Ross, P., and R. Ruthen: "Hard Pressed: Squeezed Hydrogen Forms Metal with Superconducting Potential," *Sci. Amer.*, 26 (November 1989).
Silvera, I. F.: "Evidence for Band Overlap Metallization of Hydrogen—(Technical Comments)," *Science*, 863 (February 16, 1990).
Staff: "Handbook of Chemistry and Physics," 73rd Edition, CRC Press, Boca Raton, Florida (1972–1973).

HYDROGENATION. In its simplest interpretation, to hydrogenate is to add hydrogen. There are scores of examples where hydrogenation is used as a unit process throughout the chemical and process industries. Generally, the process is associated with relatively high pressure, elevated temperature, and the presence of a catalyst.

Nickel, prepared in finely divided form by reduction of nickel oxide in a stream of hydrogen gas at about 300°C, was introduced by Sabatier (1897) as a catalyst for the reaction of hydrogen with unsaturated organic substances to be conducted at about 175°C. Nickel proved to be one of the most successful catalysts for such reactions. The unsaturated organic substances that are hydrogenated are usually those containing a double bond, but those containing a triple bond also may be hydrogenated. Platinum black, palladium black, copper metal, copper oxide (Adkin catalyst), nickel oxide, aluminum, and other materials have subsequently been developed as hydrogenation catalysts. Temperatures and pressures have been increased in many instances to improve yields of desired product. The hydrogenation of methyl ester to fatty alcohol and methanol for example, occurs at about 3,000 psig (204 atmospheres) and 290–315°C. In the hydrotreating of liquid hydrocarbon fuels to improve quality, the reaction may take place in fixed-bed reactors at pressures ranging from 100 to 3,000 (7 to 204 atmospheres) psig. Many hydrogenation processes are of a proprietary nature, with numerous combinations of catalysts, temperature, and pressure possible.

Among the better known products of hydrogenation are hydrogenated vegetable and fish oils which may be hardened or solidified by catalytic hydrogenation. Some of these oils can be partially hydrogenated to clarify and deodorize them. Fatty oils, such as oleic acid, may be converted into stearic acid by hydrogenation. Through hydrogenation, peanut oil, cottonseed oil, and coconut oil can be converted to materials that taste, appear, and smell like lard; or by varying the process, they can be made to resemble tallow. Most synthetic shortenings are comprised of hydrogenated oils. Usually, hydrogenated oils will have higher melting points and lower iodine values than the natural untreated oils.

Hydrogenation of Coal and Crudes. The interest in hydrogenation has been greatly intensified since the early- and mid-1970s in connection with the synthesis of new types of fuels to augment the world energy supplies. Basically, however, the hydrogenation of coal is not a new concept, but dates back at least a half-century to the time when manufactured gas (artificial, illuminating, producer, water gas, etc.) was used prior to the more general availability of low-cost, cleaner natural gas. In 1927, a White Paper was published discussing the processes then available for production of oil from coal. One of the first large-scale applications of the Fischer-Tropsch process for the production of oil from coal was that of the South African Coal, Oil and Gas Corporation's plant in Sasolburg, Republic of South Africa, constructed in the mid-1950s and expanded and improved several times during the interim.

Similarly, sour crudes, heavy residuums, and other petroleum-base starting materials can be hydrogenated, sometimes coupled with other processes, to sweeten, reduce viscosity, and otherwise improve the materials for better use as fuels. See also **Coal; Hydrotreating;** and **Petroleum.**

HYDROGEN CYANIDE. HCN, formula weight 27.03, colorless gas with characteristic odor, very poisonous, mp −14°C, bp 26°C, critical temperature 183.5°C, critical pressure 50 atmospheres, density 0.20 g/cm^3, sp gr 0.697 (18°C). There are two isomeric forms: (1) HCN which forms cyanides, (2) HNC (inferred from its derivatives) which forms isocyanides. Hydrogen cyanide is soluble in H_2O, or alcohol, or ether in all proportions. The compound usually is marketed as an aqueous solution containing 2–10% (weight) HCN. For many process uses, it is frequently more convenient to generate HCN as needed and thus avoid storage and handling problems. HCN burns with a red-blue flame, yielding CO_2, nitrogen, and H_2O. Aqueous solutions of HCN decompose slowly, yielding ammonium formate: $HCN + 2H_2O \rightarrow HCOONH_4$. Decomposition is slowed by storage in dark locations. Peaches, apricots, bitter almonds, cherries, and plums contain some HCN derivatives in their kernels, frequently in combination with glucose and benzaldehyde as a glucoside (amylgdalin). The bitter almond fragrance of HCN and its derivatives sometimes can be detected in such kernels.

Production. Hydrogen cyanide can be prepared from a mixture of NH_3, methane, and air by partial combustion in the presence of a platinum catalyst:

$$HN_3 + CH_4 + 1.5\ O_2 + 6N_2 \rightarrow HCN + 3H_2O + 6N_2.$$

The process is carried out at about 900–1,000°C; yield ranges from 55–60%. In another process, methane (contained in natural gas) is reacted with NH_3 over a platinum catalyst at from 1,200–1,300°C, the reaction requiring considerable heat input. In still another process, a mixture of methane and propane is reacted with NH_3: $C_3H_8 + 3NH_3 \rightarrow 3HCN + 7H_2$; or $CH_4 + NH_3 \rightarrow HCN + 3H_2$. An electrically-heated fluidized bed reactor is used. Reaction temperature is approximately 1,510°C.

The high-tonnage uses of HCN are in the preparation of numerous chemical products and intermediates for organic syntheses. As a gas, HCN sometimes is applied as a disinfectant; or cellulosic disks impregnated with HCN may be used. In ore processing and metal treating, cyanides are widely used.

Hydrogen cyanide reacts with hydrogen at 140°C in the presence of a catalyst, e.g., platinum black, to form methyl amine CH_3NH_2; when burned in air, produces a pale violet flame; when heated with dilute sulfuric acid forms formamide $HCONH_2$ and ammonium formate $HCOONH_4$; when exposed to sunlight with chlorine forms cyanogen chloride CNCl, plus hydrogen chloride. An important reaction of hydrogen cyanide is that with aldehydes or ketones, whereby cyanhydrins are formed, e.g., acetaldehyde cyanhydrin $CH_3CHOH \cdot CH$, and the resulting cyanhydrins are readily converted into alpha-hydroxy acids, e.g., alpha-hydroxypropionic acid $CH_3 \cdot CHOH \cdot COOH$.

Metallic cyanides are (1) soluble, e.g., sodium cyanide NaCN, potassium cyanide KCN, calcium cyanide $Ca(CN)_2$, mercuric cyanide $Hg(CN)_2$, aurous cyanide AuCN, (2) insoluble, e.g., silver cyanide AgCN, cuprous cyanide CuCN, (3) complex, (a) decomposed by dilute H_2SO_4 and not affected by dilute NaOH, e.g., sodium silver cyanide $NaAg(CN)_2$ solution, sodium cuprous cyanide $NaCu(CN)_2$ colorless solution, (b) changed only to acid by dilute H_2SO_4 and reactive with dilute NaOH, e.g., potassium hexacyanoferrate(II) $K_4Fe(CN)_6$ yields, with dilute H_2SO_4, hexacyanoferric(II) acid, cupric hexacyanoferrate(II) $Cu_2Fe(CN)_6$ yields, with dilute NaOH, cupric hydroxide.

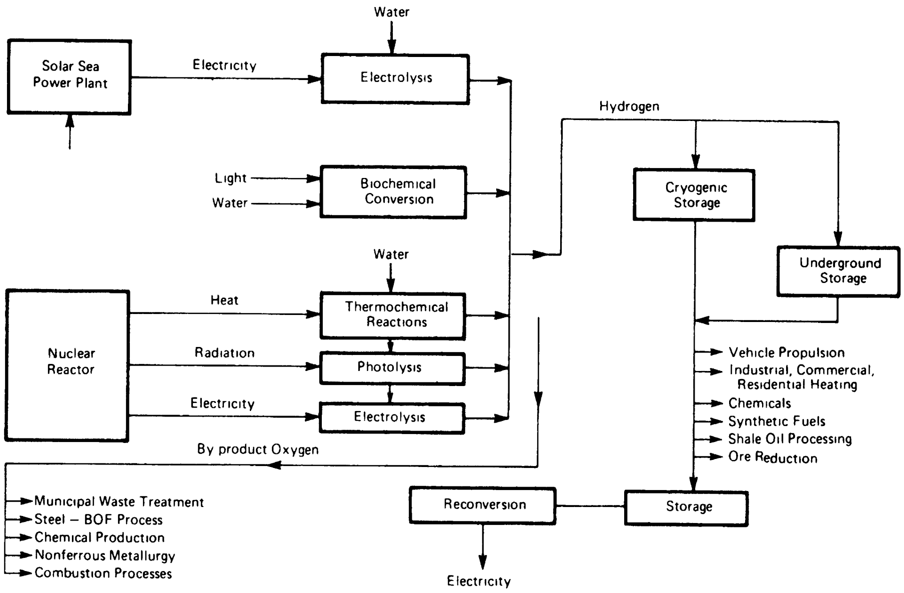
Sodium cyanide solution dissolves certain metals (1) with absorption of oxygen, e.g., gold, silver, mercury, lead, (2) with evolution of hydrogen, e.g., copper, nickel, iron, zinc, aluminum, magnesium; and solid sodium cyanide, when heated with certain oxides, e.g., lead monoxide PbO, stannic oxide SnO_2, yields the metal of the oxide, e.g., lead, tin, respectively, and sodium cyanate NaCNO. Two classes of esters are known, cyanides or nitriles, and isocyanides, isonitriles or carbylamines, the latter being very poisonous and of marked nauseating odor.

Methyl cyanide CH_3CN, bp 82°C, formed by reaction of (1) methyl iodide and potassium cyanide, (2) acetamide and phosphorus pentoxide. Methyl isocyanide CH_3NC, bp 60°C, formed by reaction (1) of methyl iodide and silver cyanide, (2) of methylamine, chloroform and NaOH solution warmed. Ethyl isocyanide C_2H_5NC, bp 78°C. Phenyl isocyanide C_6H_5NC, bp 78°C at 40 torr pressure.

Oxidation of cyanide ion (e.g., by copper(II) gives cyanogen or oxalonitrile NCCN, poisonous colorless gas, bp −21°C. This reacts with organic compounds and bases like a halogen, for example, disproportionating in aqueous alkali to cyanide and cyanate. In aqueous acid, hydrolysis to oxalamide and ultimately oxalic acid takes place. Oxidation of cyanides by oxygen donors (e.g., lead monoxide or dioxide, manganese dioxide or dichromate) a little below red heat produces cyanates.

HYDROGEN (Fuel). Because of the wide use of hydrogen in the processing industries and for the hydrogenation of various oils and fats in the food and related industries, hydrogen has become much better understood during the past several decades. For many years, hydrogen has served as a specialized fuel for certain applications, such as oxyhydrogen cutting and welding torches. But generally, until the late 1960s, the possible role of hydrogen as a major energy source fuel was rarely discussed. The word *hydrogen* took on a negative connotation with the development of the hydrogen bomb, as it also did some years ago when the hydrogen-filled dirigible Hindenburg exploded as it moved towards its mooring mast in Lakewood, New Jersey in 1937.

The probable future of hydrogen in the world's energy system was the subject of prophecy over one-hundred years ago. In 1874, Jules Verne wrote: "I believe that water will one day be employed as a fuel; that hydrogen or oxygen, which constitute it, used singly or together, will furnish an inexhaustible source of heat and light." And, in the early 1900s, Britain's Lord Haldane said: "It is axiomatic that the exhaustion of our coal and oil fields is a matter of centuries only. ... As it has often been assumed that their exhaustion would lead to the collapse of industrial civilization, I may perhaps be pardoned if I give some of the reasons which led me to doubt this proposition." Haldane envisioned net-



Major elements of a hydrogen fuel economy.

works of windmills generating the electricity needed to separate hydrogen from water. The hydrogen would then be liquefied and stored underground.

Some readily apparent advantages of hydrogen, both as a direct and an indirect fuel (discussed later) have been extrapolated into terms of a future hydrogen economy. As the result of continuing and concentrated research and development in the energy field, many experts see a hydrogen energy economy gradually emerging.

Like other energy proposals, and there have been many in the past decade or so, three factors will likely determine the pace of hydrogen energy technology: (1) the manner in which, step-by-step, hydrogen-oriented systems and subsystems will compete economically and environmentally with other energy source, conservation, and utilization proposals; (2) the pace of technological advancement in related fields, such as nuclear engineering, upon which hydrogen systems may depend; and (3) the pace of unilateral efforts on behalf of hydrogen-oriented systems, including the refinement of current planning-purpose data and opinions into actual operating information relating to hydrogen generation, transportation, conversion and/or end-utilization, and safety. Without the funding of a series of "crash programs," unilateral developments probably will be relatively slow. Most likely, the information bank for hydrogen systems will stem from an increasing awareness of the energy characteristics of hydrogen and the progressive use of hydrogen subsystems in situations where they are eminently superior.

The present concept of a hydrogen fuel economy includes a primary energy source, such as a nuclear fission or fusion reactor, a geothermal source, or a solar-powered source, with hydrogen being produced as the portable energy carrier. See accompanying figure. Thermal energy from nuclear sources would be used to generate electricity that would then be used to electrolyze water for the production of hydrogen and oxygen. The hydrogen would be distributed by pipeline to distant points of use, with storage provided by underground gas storage, or by liquefaction and refrigerated storage.

Fuel-related Background of Hydrogen. Although the abundant hydrogen isotope *protium* is the simplest known atom, it forms two diatomic molecules, namely, *ortho-hydrogen*, in which the two atomic nuclei spin in the same direction; and *para-hydrogen*, in which the nuclei spin in opposite directions. While the equilibrium composition of hydrogen gas is 75% ortho at ambient temperature, it changes to 99.8% para in the liquid state. The transition from ortho- to para-hydrogen is exothermic (168 cal/gram), so that the heat released is more than enough to revaporize liquid hydrogen (heat of vaporization 107 cal/gram). Recognition of the existence of the ortho-para transition and the development of catalysts to equilibrate the liquid during liquefaction have made possible the large-scale production, use, and storage of liquid hydrogen.

Hydrogen molecules dissociate to atoms endothermally at high temperatures (heat of dissociation about 103 cal/gram mole), in an electric arc, or by irradiation. This property is used to effect atomic-hydrogen arc welding, in which hydrogen gas is dissociated by an electric arc between two tungsten electrodes, the hydrogen atoms recombining at the metal surface to provide the heat required for welding.

Pertinent properties of hydrogen are given in Table 1.

TABLE 1. FUEL PROPERTIES OF HYDROGEN

Melting point, K	13.96
Heat of fusion at 14.0 K. calories/gram	14.0
Boiling point at 1 atmosphere. K	20.39
Heat of vaporization at 20.4K, calories/gram	107
Density. grams/cubic centimeter	
Solid at 4.2K	0.089
Liquid at 20.4K	0.071
Critical temperature K	33.3
Critical pressure, atmosphere abolute	12.8
Critical volume, cubic centimeters/mole	65.0
Critical density, grams/cubic centimeter	0.031
Heat of transition, ortho to para at 20.4K calories/gram	168
Specific heat (At constant pressure C_p, calories/gram)	
Liquid at 17.2K	1.93
Solid at 13.4K	0.63
0-200°C	3.44
Specific heat (At constant volume C_p (0-200°C) calories/gram	2.46
Specific heat: Ratio C_p/C_r (0-200°C)	1.40
Gas density, 0°C and 1 atmosphere, grams/liter	0.0899
Gas specific gravity (Air = 1.0)	0.0695
Gas thermal conductivity, 25°C (cal)(cm)/(s)(cm^2)(°C)	0.00044
Gas viscosity, 25°C and 1 atmosphere, centipoise	0.0089
Coefficient of thermal expansion per °C	0.00356
Heat of combustion at 25°C, kcal/gram mole	
Gross	63.3174
Net	57.7976
Energy release upon combustion, calories/gram	29.000
calories/cubic centimeter	2.050
joule/gram	1.21×10^5
Flame temperature, K	2.483
Autoignition temperature, K	858
Heat of formation of HF at 25°C, kcal/gram mole ΔH	−64.2
Flammability limit, percent	
In oxygen	4 to 94
In air	4 to 74

Actual and potential uses for hydrogen can be predicted by inspection of its properties. Its low density, 7% that of air, plus its high thermal conductivity, 6.7 times that of air, have led to its use as a coolant in large rotating electrical equipment. The low density reduces windage friction losses to less than 10% those with air, while its high thermal conductivity and heat capacity permit more efficient heat transfer, the result being an overall increase in generator efficiency of as much as 1%.

The high heats of reaction of hydrogen with oxygen or fluorine, plus the low molecular weights of the product gases, have made hydrogen a prime fuel for rocket propulsion, since rocket thrust increases directly with the temperature and inversely with the molecular weight of the exhaust gases. Liquid hydrogen and oxygen were used in the second- and third-stage Saturn engines in the Apollo moon flights. The low atomic weight of hydrogen has made it the preferred propellant for nuclear rockets, in which nuclear emission provides heat for exhausting hydrogen gas at high temperatures.

Some studies have indicated that the cost of transporting and distributing hydrogen by pipeline may be less than the cost of transporting and distributing electric power. Presumably existing natural gas pipelines and distribution systems can be adapted to the use of hydrogen. Although hydrogen has a net heating value of only 275 Btus per cubic foot (2448 Calories per cubic meter) as compared with 913 Btus per cubic foot (8126 Calories per cubic meter) for methane, the lower density and viscosity of hydrogen make it possible for a pipeline to deliver about the same amount of thermal energy as with methane, at a somewhat greater compression cost. The thermal energy in hydrogen can be utilized more efficiently in home heating than natural gas, because hydrogen can be burned in nonvented heaters, with no loss of heat, since its only primary combustion product is water. By using flameless catalytic heaters, nitrogen oxide formation can be eliminated. However, oxygen depletion of closed spaces will still present a hazard.

One advantage of hydrogen as a source of thermal energy, as compared with electricity, is that it can be stored for later use—it is a commodity with weight and volume. Electricity, although it can be converted into chemical energy in batteries, essentially is a form of energy that must be used as it is generated. Hydrogen, like natural gas or substitute natural gases, may be stored and transported as a refrigerated liquid, or stored as a gas under pressure in underground systems. Hydrogen also may be stored as a metallic hydride.

Categories of Energy-related Hydrogen Uses. The probable functions of hydrogen in future energy technology may be put into two major categories: (1) *direct functions* in which hydrogen serves as a fuel, that is, as the source of heat, power, and light without prior conversion to some other energy form; and (2) *indirect functions* in which hydrogen is an important component of the total energy system, but before the end-use of that energy, the hydrogen is involved in some conversion, possibly chemically, used in the creation of a synthetic fuel, such as substitute natural gas, or possibly converted into electrical energy which becomes the final end-energy used. One of the major indirect or secondary roles proposed for hydrogen is that of an energy transporter, wherein in one scheme, other forms of energy would be consumed to generate hydrogen which then would be pipelined and stored at distant points available for another conversion step—for example, converted into electrical energy as needed.

Hydrogen as Energy Source for Motive Power

Aside from their relatively low costs until the mid-1970s and continuing into the 1980s, the hydrocarbon fuels, notably gasoline and kerosine, have offered convenience in handling and transportability for use in connection with powered vehicles. And, during the past decade, the political factors that arise from the striking geographic imbalance between petroleum resources and petroleum consumption in most regions of the world have provided ample incentives to strike out for alternative sources of vehicular power.

Hydrogen, when cost competitive, can provide many of the advantages of petroleum liquids and offer the additional attraction of decreasing air pollution.

Because of its low density, the net storage volume required would be at least as much as for gasoline. The storage tank must be maintained at a temperature of −423°F (−253°C), which is the boiling point of hydrogen at atmospheric pressure. This would require insulation that would increase the overall size of the storage container. Vaporization losses from the storage tank, amounting to perhaps 2% or more per day, must be vented so that no ignition of the vented hydrogen gas can occur, and no accumulation of explosive hydrogen-air mixtures are possible. The lower explosive limit of hydrogen in air is 4%, so adequate ventilation must be provided. Fortunately, hydrogen gas, being the lightest gas with a specific gravity of 0.07 referred to air, will rise and diffuse rapidly and thus can be easily dispersed. Service stations for dispensing liquid hydrogen will require more expensive storage and pumping facilities than required for gasoline.

The estimated weights and volumes expressed in Table 2 are relative to the same energy content of gasoline. Relative weight includes that of containers. The data indicate that magnesium hydride would be at a 4.6 weight disadvantage and thus require four times the tankage in comparison with the use of gasoline in a conventional automobile. New hydrides, as described later, may change this. This would be equivalent to 450 pounds (204 kilograms) of added vehicle weight and 60 more gallons (227 liters) (2 × 2 × 2 feet storage; 0.2 cubic meter) over that required for a vehicle with a 20-gallon (76 liters) gasoline tank. Bursting upon collision for liquid storage can be overcome by using containers capable of withstanding 30 Gs, which are presently available.

TABLE 2. SOME HYDROGEN STORAGE OPTIONS FOR VEHICLES

Storage System	Relative System Weight[a]	Relative Contained Volume[a]
Gaseous phase, 2,000 psi (136 atmospheres)	~30.0	~24.0
Solid (as magnesium hydride with 40% porosity)	4.6	4.0
Liquid phase at 37°R	2.4	3.8

[a]Relative to gasoline, as unity for same energy content.

If a designer were to elect the option of using hydrogen in the gaseous phase at 2000 psi (136 atmospheres), this would require a metal container weighing some 30 times and requiring a volume of some 24 times that required for an energy equivalent volume of a hydrocarbon fuel. Also important in the total energy equation is the additional energy required to compress hydrogen (gaseous phase) or to liquefy it.

It is most likely that the first major use of liquid hydrogen as an energy source for motive power will be jet aircraft, largely because of the excellent weight advantage and the less serious nature of the boil-off loss and distribution problems as compared with other forms of transportation. City buses and long-haul motor trucks, already equipped mainly with hydride hydrogen power, have been tested in the United States and West Germany, among other countries. These may follow as candidates wherein refueling may be effected through replacement of entire storage tanks (dewars). Because the private motorcar presents the most crucial logistics problems, including the small-capacity fuel system, concern with safety, boil-off loss of fuel even when vehicle is not in use, and the education and acceptance involving millions of users, it probably will follow rather than lead the use of hydrogen in other modes of transportation. However, during the last few years, a few firms have offered hydrogen-powered private motor vehicles, set up to switch from hydrocarbon fuel to hydrogen and vice versa, but at a cost that is not competitive with mass-marketed vehicles.

Metal Hydrides. For a number of years many scientists and advanced planners have considered the possible use of metal hydrides to store hydrogen at atmospheric or reasonable pressures and at relatively low temperatures (comparable to current metal temperatures in some conventional engines). The fact that hydrogen will form hydrides with most metals has been known for many years, during which time, a number of these hydrides have been formed and tested. Until comparatively recently, magnesium hydride and a hydride of a rare-earth metal plus nickel, such as $LaNi_5$, appeared to be best suited as hydrogen storage media.

The hydride-forming reaction is exothermic and reversible: Metal + Hydrogen ⇌ Metal Hydride + Heat. Thus, when it is desired to call for the separation of hydrogen from the hydride, heat (of decomposition) is required. As may be expected, the heat of decomposition is roughly proportional to the stability of the particular hydride. It is thus evident that for a metal hydride to serve as an efficient and viable means for storing hydrogen, it should be capable of decomposition at a relatively low temperature—say 300°C or lower. At the same time, the hydride must be reasonably stable and, of course, not require a high hydrogen pressure to manufacture it. It is further evident that the metal portion of the hydride must be comparatively inexpensive and thus common and readily available in the quantities that may be required. Metal hydride storage, operating as it does in terms of hydrogen as a battery operates in terms of electricity, must be capable of easy and efficient replenishing or "recharging" cycles. In every respect, the hydride storage element must be as safe as current vehicular fuel systems.

Researchers have investigated a large number of known binary hydrides, i.e., compounds which contain one metal and hydrogen. Investigators now regard magnesium hydride (MgH_2) as a borderline possibility. This binary hydride evolves hydrogen at a pressure of one atmosphere and requires a decomposition temperature of 289°C.

In comparatively recent research, much has been learned concerning the manner in which hydride compounds hold hydrogen. It has been known for a long time, of course, that the metal portion of the hydride should be comprised of tiny particles so that there is a large surface area available for reaction. In searching for reasons why hydrides permit such a high density of hydrogen, Reilly/Sandrock (1980) have observed that it is possible to pack more hydrogen into a metal hydride than into the same volume of liquid hydrogen. When the subject metal is first exposed to diatomic hydrogen (H_2), the hydrogen atoms are adsorbed onto the surface of the metal. Immediately, some of the hydrogen is dissociated into monoatomic hydrogen (H). This permits the monoatomic hydrogen to penetrate deeply into the crystal lattice of the metal and to occupy what are known as *interstitial sites*. Investigators have found that these sites must have a critical minimum volume if they are to easily receive the hydrogen atom. Upon increasing the pressure of the hydrogen applied, the metal reaches a saturated phase—the metal hydride phase. It has been found that under certain conditions and with certain metals, the number of hydrogen atoms contained in the crystal will range from 2 to 3 times the number of metal atoms.

The most recent experimentation with metal hydrides has involved multiple-metal hydrides. It has been known for some time that hydrogen reacts with alloy metal combinations. Considerable research has gone forth in connection with ternary hydrides (2 metals + hydrogen), and one of the most promising of these compounds as of the early 1980s is iron-titanium hydride ($FeTiH_x$), where x may range from 1 to 2. Reilly/Sandrock report that the hydrogen storage capacity by weight percent of this ternary hydride is 1.75 and by volume (grams per milliliter) is 0.096. The energy density by weight is 593 calories per gram; and the energy density by volume is 3254 calories per milliliter. Thus, this ternary hydride has a higher hydrogen-storage capacity than an equal volume of liquid or gaseous hydrogen (at 100 atmospheres). Another promising intermetallic hydride is lanthanum-pentanickel hydride ($LaNi_5H_x$), although it is more costly to produce. In this hydride, x may range from 1 to 6. Both the iron-titanium hydride and the lanthanum-pentanickel hydride have low temperatures of formation and decomposition, contributing to easy charging and discharging at ambient temperature.

In connection with hydrogen engine design, it is assumed that the heat of decomposition required by the hydride can be furnished from the inevitable waste heat generated by any engine. See also **Hydride.**

Hydrogen as a Heating Fuel

The routine use of hydrogen as a heating fuel for industry and commercial-residential installations entails even greater complications and would appear to be much more dependent upon the overall economic and technical aspects of a so-called hydrogen fuel economy. From many standpoints, assuming availability, hydrogen can be an excellent fuel for almost any heating application. Hydrogen can be used in the home for cooking and heating (and even lighting) and likewise in commerce and industry. Compared with natural gas, hydrogen burns with a faster, hotter flame. Hydrogen-air mixtures are flammable over wider limits of mixtures. Hydrogen burns without producing noxious exhaust products, allowing unvented appliances except where water vapor and resulting increased humidity may be objectionable. In winter, the additional humidity can, in fact, be highly desirable. But, in humid locations in summer, the water vapor produced could be objectionable. Adequate ventilation must be provided to prevent depletion of oxygen in closed spaces.

But, generally because of the absence of hazards from carbon monoxide and other fumes, large savings could be achieved from the elimination or at least simplification of flues. Some experts suggest that not only construction costs could be lowered as the result of clean burning, but that an increase of some 30% in the efficiency of a gas-fired home heating system could be achieved. The concept of peripherally placed unflued devices, particularly through the use of catalytic "flameless" heaters, could ultimately lead to a serious revision of the widely accepted central heating concept. By maintaining the temperature of a catalytic bed as low as 100°C, the production of nitrogen oxides would be virtually eliminated.

Because hydrogen burns with a hotter flame, some design features of heating apparatus would require change. The energy content per unit mass of liquid hydrogen is about 2.75 times greater than that of of hydrocarbon fuels. On the other hand, there are only 325 Btus per standard cubic foot (2893 Calories per cubic meter) of hydrogen as compared with about 1,000 Btus per standard cubic foot (8900 Calories per cubic meter) of natural gas, thus dictating further design changes. The ignition energy of hydrogen is about 0.02 millijoules, which is less than 7% that of natural gas, a major factor in making low-temperature catalytic burners possible; also a major factor in designing for safe operation.

Despite the numerous advantages of hydrogen as a direct heating fuel, particularly in the home, the application of hydrogen must be viewed in terms of the total energy concept of an exclusively hydrogen-supplied (all-hydrogen home) installation. Where the direct use of hydrogen for heating is large, the economy will be most favorable. If a substantial amount of the hydrogen must be converted into electrical energy, as by a fuel cell, then economic justification becomes more difficult.

Lighting in the all-hydrogen home may be accomplished by condoluminescence, a cold process. A phosphor is spread on the inside of a tube similar to the conventional fluorescent lamp. Upon coming in contact with the phosphor, small amounts of hydrogen combine with the oxygen in the air to excite bright luminescence in the phosphor.

Conversion of burners and other design aspects of heating systems and appliances to pure hydrogen, or to a hydrogen-enriched natural or substitute natural gas supply, while costly and inconvenient, is certainly not in the economically insurmountable category. Similar alterations over the years were made in the United States when communities switched from manufactured gas (about 50% hydrogen) to natural gas. Such switchovers are even more recent in European communities.

As more hydrogen becomes available for transportation use and as more hydrogen is pipelined regionally or transcontinentally, depending largely on the demand placed upon the supply of hydrogen for industrial uses, it may be that the hydrogen content of community gas supplies will be progressively enriched (in a periodic, stepwise manner because of switchover problems) and thus contribute in a gradual manner to less pollution and to the conservation of natural gas.

Hydrogen as an Energy Transporter

With the possible use of hydrogen as a source of motive power in the transportation field, the *non*chemical interest in hydrogen in the total energy picture is directed to the use of hydrogen as a means or mode of storing and transporting energy. It is in this area that hydrogen directly confronts the past ever-increasing trend toward a fully-electrical energy economy. Undeniably, hydrogen energy has a major starting advantage over electrical energy, namely, hydrogen is a storable energy form. Investigations are showing that hydrogen in pipelines may cost less to transport than electricity flowing over long power lines. Thus, hydrogen may play an important future role simply as a mode of storage and transport, even though source and terminal energy conversions may be required.

Electrical power plants are most efficient when operated at constant output at full-rate load. Because of wide fluctuations in consumer load (daily and seasinally), generating rates require constant adjustment. Communication systems and some emergency systems employ batteries for interim storage of electrical energy, but these applications are minuscule when compared with the total electrical generating and distribution system. The principal means of large-scale storage is the use of pumped storage, i.e., in essence a reversible hydroelectric station wherein electrical energy is temporarily converted to a hydraulic head by pumping water to an elevated reservoir. Unfortunately, the topography has to be suitable for such an installation and thus this approach is limited to comparatively few power-generating sites. See also **Hydroelectric Power.**

The high-voltage cables required to transmit electricity from generating stations to load centers are costly. The cost of going to underground cables for transmitting bulk current ranges from 9 to 20 times that of overhead configurations. The effective use of cryogenic superconducting cables may lower underground costs considerably, but much research remains to be completed before this is possible.

Because of the tremendous volumes of fuel that can be moved in pipelines, the construction, maintenance, and operating costs of a buried pipeline are much less in terms of a percentage of the total product moved. Pipeline operations have been profitable even at the relatively low price ranges for liquid and gaseous fuels prevailing prior to the price rises of the 1970s and 1980s. Pipeline technology, of course, is well established—with several hundred thousand miles of trunklines installed and operating in the United States. These lines transport nearly 23 trillion cubic feet (0.65 trillion cubic meters) of gas. Typical pipelines range from 600 to 1,000 miles (965–1609 kilometers) in length and are up to 48 inches (1.2 meters) in diameter. Line pressures may range from 600 to 800 psi (41–54 atmospheres) but go up to 1,000 psi (68 atmospheres). A representative 36-inch (0.9 meter) pipeline will carry a gaseous fuel with the equivalent of 37,500 billion Btus (9450 billion Calories) per hour. The electrical energy equivalent would be 11,000 megawatts. By comparison, this is ten times the energy-carrying capacity of a single-circuit 500-kilovolt overhead transmission line.

The figures for pipeline transportation of pure hydrogen are not quite so attractive, but nevertheless the comparison with electric transmission costs remains highly significant.

One study shows that the pipeline transmission costs for hydrogen will range from 30% to 50% more than for natural gas. Conversion of an existing natural gas line to hydrogen service is estimated to require a rise of compressor capacity by a factor of 3.8 and compressor horsepower by 5.5.

Obviously, hydrogen transmission costs represent but one part of a total system. Should the costs of generating hydrogen in the first place, and the subsequent conversion of hydrogen into electricity at the terminal end of the system remain excessively high, then the savings in energy transportation costs, of course, become academic.

Sources of Hydrogen

The major source of chemical hydrogen over the past several decades has been natural gas. In strictly terms of chemical needs, where economic factors are favorable, natural gas has served this need well. Obviously, in terms of total energy conservation, where hydrogen is looked to as a means of conserving fossil-fuel sources, a much less costly and much more abundant hydrogen-containing raw material must be sought. The logical candidate is water. Particularly in areas of the world where hydrocarbons are not readily available, reasonably large water electrolysis installations have been made, notably in locations with low electricity costs.

In addition to electrolysis, the principal means under consideration for deriving hydrogen from water is that of thermochemical splitting. The waste heat and high temperature available from certain types of nuclear reactors would effect a series of chemical reactions, still much in the research phase, to free hydrogen and oxygen from water. Additional proposals have included the use of ultraviolet radiation from the plasma of a fusion reactor for the direct photolysis of water vapor (Department of Energy) and the use of some forms of algae, under the stimulation of light, to convert hydrogen ions to hydrogen gas by a complex chain of biochemical reactions (L. O. Krampitz, Case Western Reserve University).

Electrolysis. Because of years of operating experience, electrolysis is possibly an order of magnitude ahead of other proposals from a technological standpoint. Although simple in concept, electrolysis is costly—hence the research efforts to find other ways of splitting water carry a high incentive. Nevertheless, this side of one or more breakthroughs in other areas, most likely electrolysis operations will continue to serve as the basis for costs in extending the use of hydrogen in the relatively near term.

As of the early 1980s, industrial electrolyzers ranged in size from 500 standard cubic feet (14.2 cubic meters) of hydrogen production per day, consuming 3 kilowatts of electricity, to more than 40 million standard cubic feet (~1.1 million cubic meters) of hydrogen per day, consuming 240,000 kilowatts. Most common installations are from 10,000 to 500,000 standard cubic feet (283–14,160 cubic meters) of hydrogen per day. Two factors generally characterize an electrolyzer installation: (1) access to comparatively low-cost electricity, as found in some areas served by hydroelectric installations; and (2) need for the oxygen which accompanies the production of the hydrogen. Industrial electrolyzers usually operate at efficiencies of about 60% to 70%. Some high-pressure prototype models have reached 85%. It has been pointed out (D. P. Gregory, Institute of Gas Technology) that, in theory, electrolyzers can approach a maximum electrical efficiency of nearly 120% as the result of the ideal unit absorbing ambient heat and also converting this energy into hydrogen. A reasonable, practical target for an improved electrolyzer appears to be around 100%. Thus, the production of electrolytic hydrogen would be limited only by the efficiency of electric current generation, namely, between 35% and 45%. An estimate has been made (E. C. Tanner, Princeton University; R. Huse, Public Service Electric & Gas Co.) that the overall conversion efficiency of electricity-to-hydrogen-to-electricity will approximate 38%. The theoretical power required to produce hydrogen from water is 79 kilowatts per 1,000 cubic feet (~28 cubic meters) of hydrogen gas. One of the largest electrolyzers operating commercially is that of Cominco, Limited (British Columbia). This is a 90-megawatt installation that produces approximately 36 tons (32.4 metric tons) of hydrogen gas per day for use in ammonia synthesis. Other large plants are located in Norway and Egypt.

Two main types of electrolyzers are in commercial use: (1) Tank cells with monopolar electrodes. Porous diaphragms separate the alternate cathodes and anodes to prevent gas mixing. The anodes and cathodes are connected in parallel to keep the required voltage at approximately 2 volts and to permit high current densities. This arrangement requires a large floor area; (2) bipolar electrodes, connected in series and suitably insulated. The electrodes are cathodic on one side; anodic on the other side. This arrangement requires less floor space, is more complex, and requires high voltages.

High pressure can increase efficiency and this concept has been under development for many years. A commercial electrolyzer (Lurgi) is available which operates at a pressure of 30 atmospheres and 90°C, requiring 300 amperes of electric current at 217 volts. In the mid-1960s, bipolar cells of porous nickel electrodes were developed which operate at current densities of 800 and 1600 amperes per square foot (0.09 square meter).

In the mid-1960s, electric-high-temperature, vapor-phase electrolysis (General Electric Co.) was developed. In this process, the electrolyte is solid, porous zirconia which contains dopants. Operating temperature ranges from 500° to 800°C. A modification of the process is under development which will produce only hydrogen by consuming by-product oxygen.

Among electrolyzer design improvements that may occur are better electrodes which may result as a spinoff from fuel-cell work. There are indications that electrode improvement could cut the costs of electrolytic hydrogen by about 20% to 25%. Electrolysis looms high in consideration of utilization of ocean thermal gradients and thus these two technologies are closely interacting.

Thermochemical Splitting. The major objective is to find one or more series of chemical reactions that will result in the satisfactory separation of hydrogen (and oxygen) from water. Considerable work has been going forth at the Nuclear Research Center, Julich, Federal

Republic of Germany, where much attention has been given to sulfur- and chlorine-base thermochemical cycles. Other researchers (Institute of Gas Technology; General Electric Co.; European Atomic Energy Community) have been probing various combinations of at least 56 chemical elements, including over 700 different compounds, that may show promise in various schemes for a closed-water-splitting cycle. It is understood that approximately 20 promising schemes have emerged, mainly centered in chlorine compounds. The most frequent flaw encountered among prospective reactions is the large amount of free energy required to force one or possibly two of the series of reactions; and the appearance of reactions that produce stable compounds incapable of regeneration.

Some of these reactions would rely upon a nuclear reactor as a heat source and would not have to await the emergence of a practical, operating fusion reactor. One sequence of reactions, in particular, is of interest:

$$CaBr_2 + 2H_2O \rightarrow Ca(OH)_2 + 2HBr$$
$$Hg + 2HBr \rightarrow HgBr_2 + H_2$$
$$HgBr_2 + Ca(OH)_2 \rightarrow CaBr_2 + HgO + H_2O$$
$$HgO \rightarrow Hg + \tfrac{1}{2}O_2$$

A drawback of this sequence is its use of highly corrosive hydrogen bromide. The scheme also requires a large inventory of mercury.

Of major concern to investigators in the thermochemical splitting schemes is the availability of appropriate materials of construction. Heat exchangers between the nuclear side and the chemical side must withstand both corrosion and radioactive contamination. The conventional nickel-chromium alloys are capable up to about 1050K; exotic, but available alloys, up to about 1400K. Above these temperatures, ceramics and new alloys may have to be used. Considerable materials research along these lines is going forth at the Los Alamos Scientific Laboratory.

Conventional Hydrogen Uses. Even before its serious consideration in the fuel economy, the demand for hydrogen grew at a rate of about 15% annually since World War II. About 3 trillion standard cubic feet (~85 million cubic meters) of hydrogen (8 million tons; 7.2 million metric tons) were produced in the United States in 1970. Not including energy applications, the chemical requirements for hydrogen are expected to increase by about 7% per year through the year 2000. Among demands for hydrogen include petroleum refining, plastics, elastomers, increased desulfurization of fuel oils, increased use in iron ore reduction, aerospace uses, and hydrogen/air fuel cells. About 42% of the hydrogen produced now is consumed in ammonia production; about 38% is used in petroleum refining. The other large consumers are metallurgical and food processing.

In terms of presently nonconventional fuels that will require increasing quantities of hydrogen as new processes develop, it is estimated that (1) synthetic crude oil from coal will require 6,500 standard cubic feet (184 cubic meters) of hydrogen per barrel of oil; (2) 1,300 standard cubic feet (37 cubic meters) of hydrogen will be required per barrel of oil from shale; and (3) 1,500 standard cubic feet (42 cubic meters) of hydrogen will be required for every 1,000 standard cubic feet (~28 cubic meters) of synthetic pipeline gas produced from the gasification of coal. Petroleum refining use of hydrogen is expected to increase to 610 standard cubic feet (~17.3 cubic meters) per barrel of crude refined. Direct iron ore reduction use of hydrogen is expected to increase to 20,000 standard cubic feet (566 cubic meters) per ton (0.9 metric ton) of iron. If there were not other hydrogen sources available, the hydrogen needs could be met by using approximately 10% of the natural gas production.

HYDROGEN PEROXIDE. H_2O_2, formula weight 34.02, in pure, anhydrous form is a viscous, colorless liquid, sp gr 1.44, mp −0.89°C, bp 151.4°C. Hydrogen peroxide is soluble in H_2O in all proportions, soluble in alcohol, or ether, but not in hydrocarbons. Reagent, chemically-pure (CP) grade H_2O_2 is a solution of 90% H_2O_2 and 10% H_2O, sp gr 1.39. This concentration contains 42% active oxygen by weight. One volume yields 410 volumes of oxygen. Hydrogen peroxide solutions are high-tonnage chemicals and are supplied commercially in several strengths, ranging from 3–35% H_2O_2 by weight. Commercial grades for oxidation and bleaching normally contain 27.5–35% H_2O_2.

To reduce the tendency of H_2O_2 solutions to decompose, storage must be at comparatively low temperatures and in light-tight containers. Often, an organic material, such as acetanalide, will retard degradation. H_2O_2 has been used as an oxidizer in liquid bipropellant systems, or as a monopropellant through controlled catalytic decomposition, in supplying oxygen to various fuel mixtures for rockets and torpedoes. Low-concentration (normally 3% H_2O_2) solutions have been used for many years as antiseptics in medical applications. Bleaching is a primary outlet for H_2O_2, particularly in connection with cotton, wool, groundwood pulp—as well as hair-bleaching formulations. The compound is used as a source of gas in foaming rubber plastics. The highly reactive H_2O_2 molecule readily participates in oxidation, epoxidation, and hydroxylation reactions and is frequently used in an intermediate capacity in chemical syntheses. In restoring old paintings, H_2O_2 has been used to convert black PbS tarnish into the original white lead sulfate.

Use in Food Processing. Within recent years, there has been increased interest in the use of H_2O_2 as a bactericidal and sporicidal agent in aseptic systems used for sterilizing food processing equipment and packaging materials. Several factors affect the success of such use. At low concentration, H_2O_2 may be regarded as bactericidal, but not highly sporicidal. The latter requires concentrations of up to 35% H_2O_2. Elevated solution temperature also increases effectiveness. Hydrogen peroxide solutions can be applied at a temperature up to 95°C because of their excellent thermal stability. Such treatment must be followed by hot-air heating at about 125°C in order to dissipate H_2O_2 residuals, which must be ≤ 0.1 ppm H_2O_2. Some inorganic salts, notably cupric salts, increase bactericidal activity. Treatment with H_2O_2 and ultrasonic radiation has been shown to produce a synergistic effect on the destruction of bacterial spores. Similarly, the combination of ultraviolet radiation and H_2O_2 appears to be synergistic. In one system, a UV-irradiated solution of H_2O_2 is used. The resistance of spores varies with species. In general, the resistance of clostridial spores to H_2O_2 is lower than that of spores of bacilli. Further details are given in the Stevenson/Shafer reference listed.

Researchers have found that alkaline hydrogen peroxide renders plant fibers more digestible by ruminants and thus suggests that a number of alternative feed sources, including crop residues and other cellulosic plant biomass, may be used in animal production. Researchers at the University of Illinois and the U.S. Department of Agriculture have treated lignocellulosic residues (wheat straw, corncobs, and cornstalks) with dilute alkaline H_2O_2 solutions and have found that the fermentability of such substances increases and that the byproducts produced may be considered an acceptable energy source for the ruminant animal. Details are given in the Kerley, et al. reference listed.

Industrial Production of Hydrogen Peroxide. The traditional process for manufacturing H_2O_2 has been the electrolysis of aqueous solutions of $KHSO_4$, H_2SO_4, or NH_4HSO_4. In recent years, chemical autoxidation processes have grown in favor, largely because of energy costs. In these processes, the feedstock may be an alkylated quinone, alkylated anthraquinone, and hydroquinone solvents, together with hydrogen, air or oxygen, H_2O, and a nickel, palladium, or platinum catalyst. The process yields a 15–75% solution of H_2O_2 in H_2O, depending upon adjustment of process concentrations and conditions to provide desired concentration. The yield for this type of process is about 90% of theoretical. The process proceeds essentially in two steps. In the first step, anthraquinone contained in a solvent is hydrogenated at a temperature of about 40°C and a pressure of 1–3 atmospheres. The anthraquinone is reduced to hydroquinone (*p*-dihydroxybenzene):

$$C_6H_4{:}(CO)_2{:}C_6H_3 + H_2 \rightarrow C_6H_4{:}(COH)_2{:}C_6H_3R.$$

R is a radical such as ethyl or tertiary butyl. In the second step, the hydroquinone solution is oxidized with air or oxygen: $C_6H_4{:}(COH)_2{:}C_6H_3R + O_2 \rightarrow C_6H_4{:}(CO)_2{:}C_6H_3R + H_2O_2$. In theory, the process consumes only hydrogen, atmospheric oxygen, and H_2O. A solvent must be used that will minimize side reactions during hydrogenation while also dissolving both the hydrogenated and oxidized forms of the organic compound. Solvents referred to in this connection are benzene-methylcyclohexanol mixtures and primary and secondary nonyl alcohols. Very tight purity precautions are required because any impurities in the H_2O_2 cause spontaneous catalytic decomposition of the product. As the result of these necessary precautions, the resulting H_2O_2 is one of the purest of commercial chemicals.

The process is highly corrosive. At one time, enameled steel vessels were standard for H_2O_2 processing. Aluminum, once properly passified through pickling and treatment after fabrication, has been found satisfactory.

Hydrogen peroxide reacts (1) with alkalis to form peroxides, (2) with potassium iodide solution, in presence of ferrous sulfate, to liberate iodine. This reaction serves to indicate the presence of as small an amount as 1 part by weight of hydrogen peroxide in 25,000,000 parts of H_2O, (3) with lead sulfide PbS, brown solid, to form lead sulfate $PbSO_4$, white solid, and sometimes used to brighten the lead pigment of darkened oil paintings, (4) with lead dioxide to form lead oxide, (5) with sulfites, especially in alkaline solution, to form sulfates, (6) with nitrites to form nitrates, (7) with arsenites for form arsenates, (8) with ferrous compounds to form ferric, (9) with chromic compounds to form chromates (see **Chromium**), (10) with permanganates in acid solution to form manganous compounds plus oxygen of twice the volume available from the hydrogen peroxide, (11) with dichromates in acid solution cold to form perchromic acid, blue solution, more soluble in ether than in acid, (12) with titanic salt solutions to form pertitanic acid, yellow solution, (13) with colored organic materials, e.g., litmus, indigo, to destroy the color, and thus used for bleaching hair, silk, feathers, straw, ivory, teeth, bones, gelatin, flour. When hydrogen peroxide solution is treated with finely divided platinum or other substances, or comes in contact with rough surfaces, e.g., ground glass, oxygen is evolved (water also formed).

In the laboratory, hydrogen peroxide is prepared from barium peroxide by treatment with ice-cold dilute acid; when H_2SO_4 is used barium sulfate insoluble may be separated by filtration. Other peroxides, e.g., sodium peroxide, react similarly with acids to form hydrogen peroxide plus the salt corresponding to the peroxide and acid used. Hydrogen peroxide is formed when ether is exposed to sunlight, when a hydrogen-oxygen flame impinges on ice, and when H_2O in a quartz vessel is exposed to ultraviolet light.

HYDROGEN SCALE. 1. A thermometric scale. (See **Temperature.**) 2. Since there is no reliable method for determining the absolute potential of a single electrode, electrode potentials are measured against a reference electrode whose potential is arbitrarily taken as zero. The arbitrary zero in general use is the potential of a reversible hydrogen electrode, with gas at 1 atmosphere pressure, in a solution of hydrogen ions of unit activity, or other electrodes calibrated against the hydrogen electrode.

HYDROGEN SULFIDE. H_2S, formula weight 34.08, colorless, odorous gas, mp −82.9°C, bp −59.6°C, sp gr 1.1895 (air = 1). The gas must be handled carefully because of (1) its toxic properties (particularly dangerous because it may paralyze the olfactory nerves), and (2) its explosive tendencies (low ignition temperature of 260°C and wide flammability range from 4.3 to 44% by volume in air). Hydrogen sulfide liberates considerable heat upon burning (6,230 calories/liter at 15.6°C). The gas is produced by acid hydrolysis of many sulfides and by water hydrolysis of those elements higher in the hydrogen scale.

An aqueous solution of hydrogen sulfide is termed hydrosulfuric acid which undergoes slow atmospheric oxidation to sulfur. The acid is a strong reducing agent, usually with the separation of sulfur, e.g., with nitric acid (nitric oxide formed), with concentrated H_2SO_4(SO_2 is formed), with permanganate (manganous ion formed in the presence of acid), dichromate (chromic ion formed in the presence of acid).

Fluorine, chlorine, bromine, and iodine react with H_2S to form the corresponding halogen acid. Metal sulfides are formed when H_2S is passed into solutions of the heavy metals, such as Ag, Pb, Cu, and Mn. This reaction is responsible for the tarnishing of Ag and is the basis for the separation of these metals in classical wet qualitative analytical methods. Hydrogen sulfide reacts with many organic compounds.

The gas results from the decomposition of metal sulfides and albuminous matter and is found in the areas of mineral springs, sewers, and in some mines where it is referred to as "stink damp." H_2S also is a by-product of several industrial processes, including synthetic rubber, viscose rayon, petroleum refining, dyeing, and leather-treating operations. In the laboratory, H_2S usually is prepared by treating a sulfide with an acid, such as iron pyrites and HCl, or by heating thioacetamide $CH_3C(:S)NH_2$. Three processes are used industrially to produce H_2S in large quantities: (1) treating a sulfide with an acid, $2NaHS + H_2SO_4 \rightarrow 2H_2S + Na_2SO_4$, (2) reacting sulfur with an alkali, $4S + 2NaOH + 2H_2O \rightarrow 2H_2S + Na_2S_2O_3$, and (3) directly reacting sulfur with hydrogen, $S + H_2 \rightarrow H_2S$. Large quantities of by-product H_2S usually are converted into elemental sulfur or H_2SO_4.

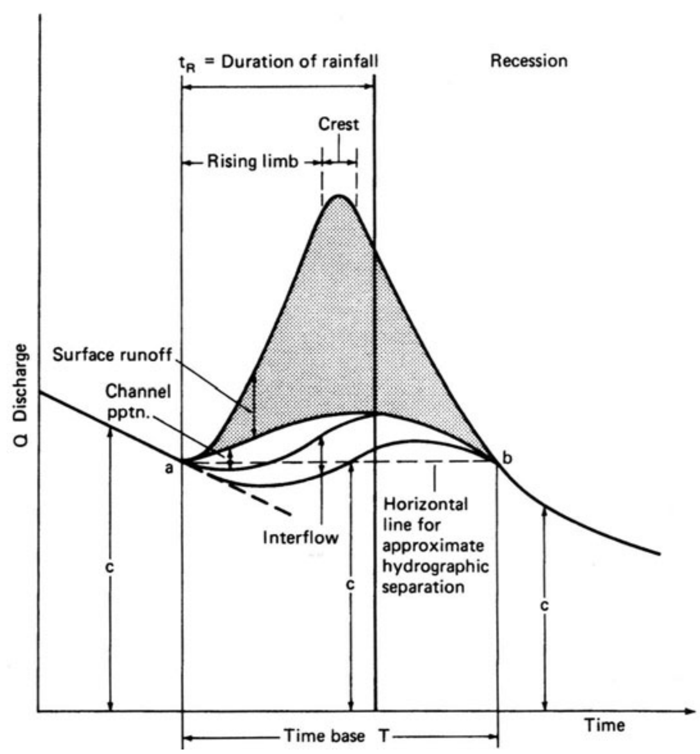
Industrial uses for H_2S include (1) the preparation of sulfides, such as sodium sulfide and sodium hydrosulfide, (2) the production of sulfur-bearing organic compounds, such as thiophenes, mercaptans, and organic sulfides, (3) the removal of Cu, Cd, and Ti from spent catalysts where the gas acts as a precipitant, (4) the formulation of extreme-pressure lubricants, and (5) the preparation of rare-earth phosphors used in color TV tubes.

See also **Coal.**

HYDROGRAPH. By graphing the discharge of a stream as ordinate against time sequence as the abscissa, a hydrograph of the stream flow is obtained. The hydrograph proves to be an important source of information in the design of sewerage and water supply systems and in the design of hydroelectric power projects. The reliability of the information it contains increases as the period of time over which the hydrograph extends is lengthened. Hydrographs extending over periods of less than 10 years are liable to be deceptive in the information they convey regarding maximim and minimum flows. The United States Geological Survey water supply papers form a valuable and important reference source for data upon which hydrographs are constructed.

Runoff and *stream flow* are synonymous. Surface runoff, interflow, and groundwater flow in varying proportions make up the total runoff in stream channels.

The *direct runoff* is that runoff which enters the stream promptly after rainfall or melting of snow. It is equal to the *surface runoff*, the *prompt subsurface runoff*, plus the *channel precipitation* which falls directly on the water surfaces of lakes and streams. Surface runoff is commonly represented in the form of a hydrograph similar to that shown by the accompanying figure.



Hydrograph parts and flow contributions: (a) and (b) are reference points; (c) is groundwater flow.

Interflow is that part of the precipitation which infiltrates the surface soil, moves laterally through the upper soil horizons as ephemeral, shallow, perched groundwater above the main groundwater level, and reaches the stream before it reaches the water table. This lateral movement results from the presence of relatively impervious horizons near the surface.

See also **Drainage Systems;** and **Hydrology.**

HYDROID. One of the two forms of individuals in the coelenterates. The polyp. This form is a tubular or sac-like individual whose body wall is composed of two cellular layers separated by a thin mesogloea. The latter contains some cells derived from the other layers but is not developed as a third cellular layer. The hydroid is usually attached to the stalk of a colony or directly to a supporting surface. Its cavity opens at the free end of the body, and the mouth is surrounded by a circlet of slender tentacles except in specialized individuals found in some colonial species. The other form of coelentenrate individual is the medusa.

HYDROKINETICS. The flowing of liquids is due to the three principal causes: pressure difference, gravity, and inertia. Bernoulli's law expresses an ideal condition fulfilled by the three components of "head" corresponding to these three causes. The value of this head (whether constant or not) is, at a given point (x, y, z) of the liquid,

$$e + \frac{p}{\rho g} + \frac{v^2}{2g} = F(x, y, z) \quad (1)$$

The terms of this expression represent lengths, usually given in centimeters or feet. The assumption of constant density requires that the product of the speed of flow by the cross section of any conserved portion of the stream shall be constant and that the streamlines (paths of the moving particles) therefore converge as the speed increases. If one could assume that the function F is really constant, or if it were possible to obtain F as a known function of the coordinates of the moving particle, then all hydrokinetic problems could be solved by applying suitable mathematics to the equation which would thus develop from (1).

Various attempts have been made to do this. Useful formulae result from assuming F constant (Bernoulli's law) and applying the equation to special cases. But when such formulas are tested, the calculated results are found to be in error, in every case indicating that appreciable energy has been lost in friction. While some improvement is obtained by introducing a friction factor, it has on the whole been found more satisfactory to employ empirical formulas adapted to each type of problem. Thus, we have the Darcy formula,

$$v = D\sqrt{\frac{d(F_1 - F_2)}{l}} \quad (2)$$

for the speed of flow in a pipe of length l and diameter d, running full, and with a difference of total head $F_1 - F_2$ at the two ends; D being a constant to be determined by experiment. Also, the Chézy formula,

$$v = C\sqrt{\frac{as}{u}} \quad (3)$$

giving the speed of flow in an inclined channel, like a ditch or a sewer; a being the cross section of the flow, u the length of channel perimeter covered by the liquid, s the fall per unit length, and C an experimental constant.

The "hydraulic grade line" is a convenient concept in connection with flow through pipes. This is an imaginary line so drawn that each point of it lies vertically above (or below) the pipe at a distance equal to the pressure head $p/\rho g$ at the corresponding point of the pipe. In the case of a siphon, part, at least, of the conduit rises above this line, which means that the pressure in this portion is less than atmospheric.

Among the more difficult problems are those of vortex motion (like a whirlpool) and turbulent flow; and the general treatment of flow through a cavity of given shape under given boundary conditions, which presents some analogies to the electric current and the conduction of heat.

HYDROLOGY. The science, or study, of water, especially in relation to its occurrence in streams, lakes, underground structures, and as snow. The study of glaciers, their origin and geological effects is usually included under the heading of Glaciology. The term hydrology is derived from the Greek meaning water, and reason, hence the science of water, including its discovery, uses, control and conservation. Since water ranks first of all the natural resources, the science of hydrology is of great practical importance. The basis of hydrology is the hydrologic cycle. All terrestrial (fresh) waters are derived from the great oceanic reservoirs through evaporation and precipitation.

Hydrology is one of several scientific disciplines that will play a major role in the "Global Change Research Program" (GCRP) announced by the U.S. National Aeronautics and Space Administration (NASA) in 1992 as a cooperative effort with other nations to use platforms and satellites in space to provide images and measurements in a concerted effort to understand better how Earth is changing, particularly as the result of human activities on Earth.

The Hydrologic Cycle. Also known as the water cycle, this is the never-ending circulation of water and water vapor over the entire earth. This circulation penetrates the three parts of the total earth system: the atmosphere (gaseous envelope above the hydrosphere), the hydrosphere (water covering the surface of the earth), and the lithosphere (solid rock beneath the hydrosphere). Solar energy and gravity provide the energy for the circulation.

Water is evaporated from the oceans and the land, with the former providing the largest amounts. The evaporated water is carried into the atmosphere, usually drifting tens to hundreds of miles before being returned to the earth as rain, snow, hail, or sleet. This precipitated water may be intercepted by plants, may run over the ground surface and into streams, may infiltrate into the ground, or fall back into the oceans. A considerable part of the water intercepted and transpired by plants and the surface runoff returns to the air by evaporation. The infiltrated water may seep down to deeper zones of the earth, forming groundwater storage which may later flow out to streams as runoff and finally evaporate into the atmosphere to complete the hydrologic cycle. Thus, the main processes involved in the hydrologic cycle are evaporation, precipitation, interception, transpiration, infiltration, seepage, storage, and runoff.

The quantity of water going through the hydrologic cycle during a given period for an area can be evaluated by the hydrologic or continuity equation:

$$I - O = \Delta S$$

where I = total inflow of surface runoff, groundwater and total precipitation
O = total outflow, which includes evapotranspiration and subsurface and surface runoff from the area
ΔS = the change in storage in the various forms of retention and interception

A qualitative representation of the hydrologic cycle is given in Fig. 1, as first depicted by R. E. Horton. The cycle is also shown diagrammatically in Fig. 2.

Magnitude of the Hydrologic Cycle. Each year, approximately 96,000 cubic miles (4×10^5 cubic kilometers or 4×10^{20} grams) of water are evaporated from the earth's surface. Of this amount, the oceans account for 84.4%, and inland water bodies and wet soils providing the remaining 15.6%. Most of the inland evaporation occurs into relatively dry air masses. Much of the water evaporated from the oceans is transported by maritime air masses (which can hold considerably more water vapor than continental air masses) to the continents, where total precipitation amounts to 24,000 cubic miles/year (100,000 cubic kilometers/year). This amount of water would cover the entire state of Texas (267,339 square miles; 692,408 square kilometers) to a depth of 475 feet (144.8 meters). Of the 24,000 cubic miles of water precipitated, 9,000 cubic miles (37.5%) returns to the sea as runoff to balance the excess precipitation over evaporation inland.

E. Reichel calculated that the mean annual precipitation for the entire world is 34 inches (86.4 centimeters), which is balanced by a comparable amount of evaporation. It is estimated that 97% of all the water in the world or over one quadrillion (10^{15}) acre-feet ($1{,}234 \times 10^{15}$ cubic meters) is contained within the oceans. If the earth were a uniform

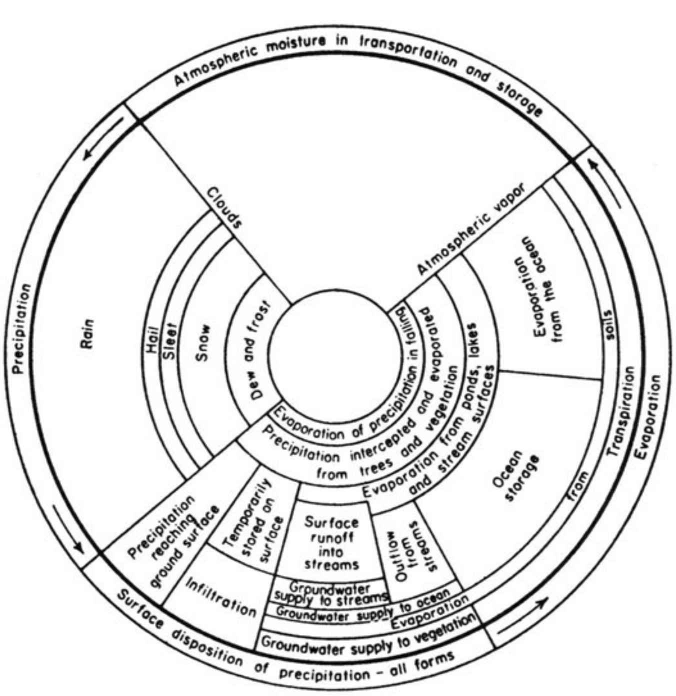


Fig. 1. The hydrologic cycle as conceived by Horton.

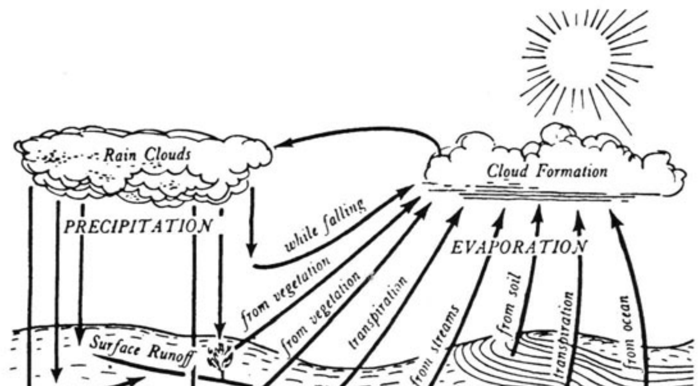


Fig. 2. The hydrologic cycle as presented by Ackermann, Colman, and Ogrosky.

sphere, this volume of water would cover the earth to a depth of 800 feet (243.8 meters), as estimated by A. Wolman.

The total volume of fresh water on the earth is estimated at 33 trillion acre-feet (4.1 × 10^{15} cubic meters; or 4.1 × 10^{21} gallons) distributed, as estimated by V. T. Chow, as follows:

Polar ice and glaciers	75%
Groundwater between 2,500 and 12,500 feet (762 and 3,810 meters)	14
Groundwater between the surface and a depth of 2,500 feet (762 meters)	11
Lakes	0.3
Soil moisture	0.06
Atmosphere	0.035
Streams	0.03

The foregoing figures are stationary estimates of distribution. Huge amounts of water pass through the atmosphere while the water content is relatively small at any given instant.

The average annual precipitation over the continental United States would amount to 30 inches (76.2 centimeters) if it were spread evenly. In actuality, the precipitation ranges from a few inches in the arid southwest to over 100 inches (254 centimeters) in parts of the Pacific northwest. Although the 17 western states contain 60% of the land area, they receive only 25% of the total precipitation. The 30 inches of water for the United States represents 4,800,000,000 acre-feet/year or 4,300 billion gallons/day. Of this amount, 21.5 inches (54.6 centimeters), or 71.7% is returned to the atmosphere by processes of evapotranspiration. The remaining 8.5 inches (21.6 centimeters) (28.3%) becomes surface and groundwater runoff into the oceans. The foregoing estimates were made by C. J. Robinove.

Mount Waialeale, Hawaii (Kauai) receives the most rain of any location in the world, averaging 460 inches (1.168 centimeters). There is also a wide range of precipitation over Canada, from about 10 inches (25.4 centimeters) in parts of Yukon to nearly 70 inches (178 centimeters) in Halifax, Nova Scotia. St. John's, Newfoundland also receives over 60 inches (152 centimeters) of precipitation.

Function of the Hydrologic Cycle. If the atmosphere and the earth were considered as separate entities, radiation and conduction fail to provide balanced heat budgets, because the earth's surface has a net gain and the free atmosphere a net loss. The link between the gain and loss is the hydrologic cycle.

Some of the heat absorbed by the earth's surface is expended in evaporation, and therefore is transferred to latent heat, which is later realized as sensible heat and released to the atmosphere when the vapor condenses to clouds. Evaporation is high where relatively cool air sweeps over warmer oceans. The highest evaporation values found in the northern hemisphere occur in the Atlantic and Pacific trade wind belts south of 30°N. High values also occur over the northwestern Pacific and North Atlantic oceans during winter when cold, dry continental air masses move over warmer waters.

The average life of water vapor molecules in air varies from an hour to several days. Latent heat is usually liberated far from the regions where evaporation occurred. This is particularly true of evaporation in the trade wind belts, which supply much of the vapor that eventually precipitates in middle and high latitudes. Thus, the circulation of water is a key part of heat tranfer from low to high latitudes and from oceans to continents.

Return of Water to the Oceans. Although there is a relatively uniform pattern of evaporation in the various latitudinal belts of the ocean, there is a marked regional imbalance in the return flow of water to the oceans. The explanation lies in the concentration of major rivers (Amazon, Mississippi, Congo, Niger, St. Lawrence, Danube, Po, Nile, and Rhine) which drain into the Atlantic Ocean and its marginal seas (Gulf of Mexico, Black Sea). In contrast, the Pacific has only a limited number of major discharge outlets (Yangtze, Hwang-Ho, Yukon, Columbia, and Colorado).

The mean annual discharges of the world's major rivers are summarized by Livingstone (U.S. Geological Survey Professional Paper 440-G) in Table 1. The information in Table 2 also provides further evidence that the Atlantic not only drains the largest portion of the earth's land surface, but has the highest proportion of land area drained to ocean area.

Basic Principles of Hydrology. Since the eighteenth century, the development of hydrologic principles has been aimed at refinement of the understanding of each distinct phase of the hydrologic cycle and of the relationship between the phases. A few of the more important principles are listed as follows, not necessarily in order of importance or discovery:

1. The recognition that groundwater moves from points of high pressure to points of low pressure (down gradient) and that gradients are often, but not exclusively, related to rock type and structure.
2. The fact that the velocity of flowing water, on the surface or underground, is governed by the differences in pressure head, or slope, and the resistance of the confining channel or of the aquifer.
3. The knowledge that water is capable of dissolving and carrying large amounts of mineral matter that changes composition as the water comes in contact with various types of potential solutes.

TABLE 1. ESTIMATED RUNOFF OF MAJOR RIVERS OF THE WORLD

River	Cubic Feet/ Second (Thousands)	Cubic Meters/ Second (Hundreds)
Rivers discharging into Atlantic Ocean		
Eastern North America		
Mississippi	620	175.5
St. Lawrence	500	141.5
South Atlantic slope	325	92.0
North Atlantic slope	210	59.4
	1655	468.4
Europe		
Danube	225	63.7
Rhine	76	21.5
Rhone	59	16.7
Dnieper	59	16.7
Elbe	24	6.8
Garonne	24	6.8
Don	24	6.8
	491	139.0
South America		
Amazon	3600	1018.8
Orinoco	600	169.8
Parana	526	148.9
Uruguay	136	38.5
	4862	1376.0
Africa		
Congo	1600	452.8
Niger	326	92.3
Orange and Zambezi	352	99.6
Nile	100	28.3
	2378	673.0
TOTAL for Atlantic Ocean	9386	2656.4
Rivers discharging into Pacific Ocean		
Columbia	345	97.6
Colorado	23	6.5
Yukon	180	50.9
Australia	354	100.2
Japan and Korea	225	63.7
Middle latitude Asian rivers	2250	636.8
TOTAL for Pacific Ocean	3377	955.7
GRAND TOTAL		
Atlantic and Pacific Oceans	12763	3612.1

4. The geologic recognition that water transports and deposits vast quantities of solid rock waste and is a major agent in the modification of land forms and in chemical alterations underground.
5. The fact that natural (underground) or artificial (surface) storage of water modifies the regimen of water in an area by changing the time of flow.

With these basic principles in mind, scientists, geographers, and engineers, who practice in the field of hydrology, are constantly attempting to refine two areas of knowledge: (1) an inventory or description of the water resources of the world—the amount of water in storage, rates and volumes of precipitation, recharge and discharge, the quantitative availability and suitability of water for use and the effects of water in terms of floods and droughts; and (2) a full understanding of water in all of its properties and cycles.

Hydrology of Coastal Terrain

In coastal districts, the fresh water in the water table migrates slowly downhill to the sea. Because of their different densities, the fresh water and salt water do not generally mix, except in the ocean where the tides, waves, and currents do the mixing. In the aquifers in coastal districts, the less dense fresh water tends to float on the more dense fresh water sometimes like an iceberg. The shape of the fresh water lens on a sandy island, assuming that fresh water is being replenished by rainfall, is shown in Fig. 3. The relationship between the thickness of the fresh water body (a) and the depth of the lowest part of the fresh water body below sea level (b) is:

$$b/a = \frac{\text{Specific gravity of fresh water}}{\text{Specific gravity of seawater}} = 40/41$$

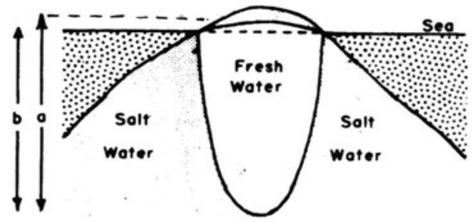


Fig. 3. Characteristic shape of the freshwater lens in islands made of uniformly permeable materials in humid areas.

Thus, for every foot the fresh water stands above sea level, the surface of the salt water lies some forty times as many feet below sea level. These figures, of course, only approximate the condition and depend upon the salinity of the seawater and the purity of the fresh water. The flow lines, i.e., the paths of water movement, for the fresh water contained within the lens are indicated in Fig. 4. Both the lens and the underlying salt water will rise and fall with the tide unless there is a barrier between the underground water and the sea. The time of the peaks and troughs of the fluctuations becomes later as traced inland, just as the time of high and low tide becomes progressively later as it is traced up the tidal portion of a river. The time between the peaks and troughs will remain the same, while the time lag will be constant for a given well.

TABLE 2. OCEANIC AND LAND-DRAINAGE AREAS
(In Millions of Square Miles and Millions of Square Kilometers)

Ocean	Area		Land Area Drained		Percent of Total Land Area	Percent of Area Drained to Ocean Area
	Square Miles	Square Kilometers	Square Miles	Square Kilometers		
Atlantic	37.8	98	25.9	67	45.3	68.5
Indian	25.3	65.5	6.6	17	11.5	26.1
Antarctic	12.4	32	5.4	14	9.4	43.5
Pacific	63.7	165	6.9	18	12.1	10.8
Interior Drainage	—	—	12.4	32	21.7	—
Total	139.2	360.5	57.2	148	100.0	

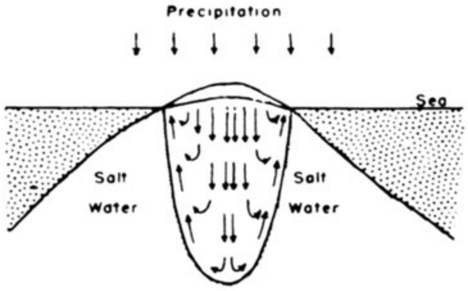


Fig. 4. Flow lines within the freshwater lens on the island shown in Fig. 3. The lens is recharged by infiltration of rain and snow into the ground, but loses water by diffusion and by flow into the sea.

Some mixing of the fresh and salt water does occur at the interface. Usually this is negligible, but it can be appreciable when favored by certain conditions. This produces a brackish water zone which may be quite thick. This zone occurs where there are considerable fluctuations in the level of the interface due to tidal action or irregular heavy rains. Thus, a strong development of a brackish zone is found in the basalt aquifers along the coast of Oahu in Hawaii. It is also increased by pumping the wells in these regions. Like the fresh water, the brackish water lens moves slowly downslope.

The worldwide rise of sea level as the glaciers melt has an important effect on the water tables in coastal areas. As the sea level rises, so does the water table and the saline water. This increases the tendency of tides to sweep saline water into rivers and it tends to push the fresh water shoreward. In the case of the Atlantic seaboard of the United States, the sea is rising at a rate equivalent to about 2 feet (0.6 meter) per century. It has been estimated that this small rise will cause the fresh water in the artesian aquifer of New Jersey to recede inland at the rate of 1 to 4 miles (1.6 to 6.4 kilometers) per century, depending on the dip of the aquifer.

The thin layer of fresh water underlain by salt water means that great care must be taken in exploiting the fresh water. The usual method of exploitation is by well from which the water flows or is pumped. When the water is taken from a well, the surface of the water table is lowered close to the well by an amount depending on the output of the well and the porosity of the aquifer, as well as other factors. As mentioned before, for every foot of lowering of the surface of the water table, the fresh-saltwater boundary moves 40 feet (12 meters) nearer the surface. Thus, it does not take a great output of water to cause the bottom of the fresh water layer to rise to the bottom of the well. Thereafter, the water produced by the well will be saline. By limiting the production, contamination can be prevented.

Particular care must be taken in the case of artesian basins in coastal regions. The quantity of fresh water that is stored is finite, and the amount of recharge is limited. Overpumping will cause influx of saline water, as has occurred in the Savannah area of Georgia and South Carolina. Only restricted pumping or artificial recharge will prevent the eventual salinization of such a productive fresh water source.

Coastal Preservation. Of hydrological concern, the preservation of ocean shorelines is of specific interest to ocean science and engineering. Giese and Aubrey (see reference listed) estimate that by the year 2025 the amount of upland lost in Massachusetts (for example) as the result of relative sea-level rise will range from 3000 to 10,000 acres (1215 to 4050 hectares). Now mostly occupied by private residences and commercial structures, "upland" refers to terrain landward of wetland that has not been altered appreciably by coastal processes (i.e., waves and tides). Giese and Aubrey observe, "Past studies of coastal upland retreat have concentrated on shore erosion and have neglected *passive submergence*, probably because such losses have been considered to be relatively small. Such is not the case, however, even at present rates of relative sea-level rise. In addition, when we consider the possible importance of measuring this loss due to relative sea-level rise, the importance of measuring this loss becomes obvious. Thus, we set out to quantify the passive retreat of upland within the coastal communities of Massachusetts."

Hyposometric curves are a convenient method for illustrating the community upland retreat rates. See Fig. 5. Calculations by Giese and Aubrey show that, by approximately the year 2025, using past sea-level rise numbers, the total upland loss for Massachusetts will be 2950 acres (1190 hectares), resulting from a sea-level rise of 0.45 feet (0.14 meters). This could triple with a sea-level rise of 1.57 feet (0.48 meters).

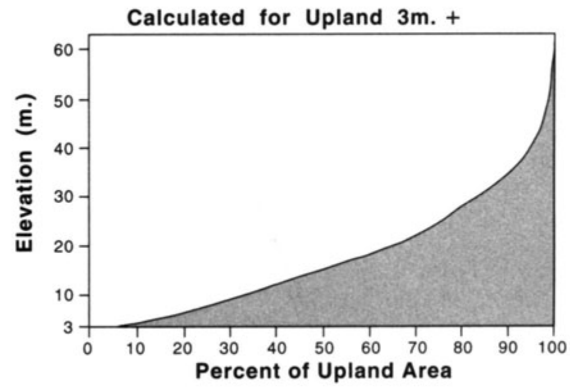


Fig. 5. Hypsometric curve for the Town of Falmouth, Massachusetts, located on Cape Cod. From the curve, one can read the percentage of upland area of the town that lies below any given elevation. Almost all Falmouth upland is less than 60 meters (197 feet) high, and about 50% of the town is less than 15 meters (49 feet) high. A hypsometric curve presents the distribution of the area of a given land (or sea-floor) surface area with respect to elevation. (*After Giese and Aubrey*.)

Within a geologic period of increasing sea level, ocean beach properties in many areas of the world are threatened. Most research has been directed toward the development of a rationale for future seashore land development programs. One example of this is the University of Florida Sea Grant Programs. This has led to the establishment of the Florida Coastal Construction Control Line, a line that delineates the 100-year coastal hazard zone within which the state has construction-permitting jurisdiction. Evidence of past erosion of beaches by sea-level rise are attested by the numerous barrier islands formed along oceanic coastlines. It is estimated that these islands were formed about 6000 to 7000 years ago during a period of similar sea-level rise. How can technology reduce the effects of sea-level rise? Some past concepts have failed; only a few have succeeded.

One concept, "beach nourishment," is the placement of large quantities of quality sand on the beach to advance the shoreline seaward. Costs are high, ranging up to \$6 million/mile (\$3.7 million/kilometer). Thus, the longevity of the nourishment is a very important economic factor.

One geologist, whose opinions have been challenged, has observed that nourished beaches erode ten times faster than natural beaches. Experience in terms of measured effectiveness is inadequate. For example, it has been observed that a beach that has been nourished may fail in its specific effects on a given beach, but that the nourishment may be passed along shore (downdrift) and assist in a small way toward nourishing beaches in adjacent areas. Dean (see reference listed) has observed, "An example is the Port Canaveral, Florida beach nourishment project (1974) in which 2.4 million cubic yards (1.8 million cubic meters) were placed over 2.1 miles (3.4 kilometers) of beach immediately downdrift of the port entrance. Recent surveys indicate nearly every grain of sand placed, although it has been transported downdrift (southward), can be accounted for. Findings like this are now being included in the preparation of cost/benefit analysis procedures for future beach nourishing projects."

Waves breaking across the surf zone are a primary function for cross-shore and longshore sediment transport. Also, of course, waves represent a significant destructive force in terms of any object located along the shore, as witnessed by hurricanes.

Dean (University of Florida), who has pioneered in the field of beach erosion, observes, "Future research agendas must include the dynamics of the surf zone and recognition of our poor understanding of this region. Although relevant knowledge has increased many fold in the last few decades, there is still much to be learned prior to development of rational design capabilities. Obvious questions include the rate of longshore and cross-shore transport under given weather conditions, the relative roles of bed load and suspended load transport, the cause of rip currents, and the mechanics of longshore bar formation."

Deltas and Estuaries. Deltas represent a special situation where freshwaters from rivers meet freshwater or seawater reservoirs, such as lakes or the oceans. See also **Estuary** and **Delta.** As pointed out by D. J. Stanley (Mediterranean Basin Project, U.S. National Museum of Natural History), "The northeastern margin of the Nile delta, including Lake Manzala, Port Said, and the northern Suez Canal, has subsided rapidly at rates of up to 0.5 centimeter per year since some 7500 years ago. This subsidence has diverted at least four major distributaries of the Nile River into this region. The combined effects of continued subsidence and sea-level rise may flood a large part of the northern delta plain by as much as 1 meter by the year 2100. The impact of continued subsidence, now occurring when sediment input along the coast has been sharply reduced because of the Aswan High Dam, is likely to be substantial, particularly in the Port Said area and as far inland as south of Lake Manzala." The aforementioned areas, with a population estimated at over 1 million, may be susceptible to flooding because it lies over one of the more rapidly subsiding parts of the delta. Continued scientific studies are needed to precisely measure present subsidence and to better determine its possible effects and thus possibly take remedial actions. In late 1990, Stanley observed, "Unless Egypt moves promptly with new coastal protection measures, the sea may advance inland by as much as 30 kilometers (18.6 miles) within the next 100 years."

Possibly the most intensely studied of the world's estuaries is Lake Michigan's Green Bay, located in the heart of North America. As pointed out by P. L. Smith and associated researchers at the University of Wisconsin Sea Grant Institute, "Green Bay is best characterized as an estuary since it functions as a nutrient trap, has exceptionally high biological productivity, and because of the thermal and chemical differences between the water of its tributaries and that of Lake Michigan. The bay's mixing process is driven by a strong wind-induced seiche coupled with a small lunar tide." Rather than one of hydrological concern, the problem with restoration activities of the estuary, which commenced in earnest in 1969, has been one of overexploitation and biological pollution. The studies and actions taken to date are elegantly described by Smith and co-researchers. See reference listed.

Hydrology of Limestone Terrain

The primary factor in the hydrology of limestone terrains is the solubility of carbonate rocks in aqueous solutions, which leads to underground networks of pipes and channels known as *karst*. Limestone terrains are defined as regions where carbonate rock formations extend from near the surface to below the water table. The soil, and the material directly beneath, strongly influence the hydrology of these terrains. An impermeable cover overlying a thick limestone sequence can cause most of the precipitation to pass out of the area as surface runoff and the effect of the limestone is minimized. A pervious soil horizon, or erosion of the impermeable cover, will facilitate infiltration to the limestone substrate.

The hydrology of a carbonate region can range from a karst area with many interconnecting passages and caves, readily absorbing surface waters and transmitting these waters fairly rapidly from source area to discharge or storage areas, to a situation where a limestone formation of low permeability acts as an aquiclude and the terrain has high surface runoff and little available groundwater. The variability of the hydrology is as significant as the unique hydrology of a fully developed karst terrain and makes the concept of an average hydrology rather hypothetical. This variability can be demonstrated by comparing an area in Texas where the Edwards limestone absorbs an estimated 156,000 acre-feet (192.4×10^6 cubic meters)/day from surface stream flow, with a chalky portion of the Cooper Marl in South Carolina which is used as an unlined public water supply conduit for Charleston. The permeability of carbonate rocks varies greatly as a function of its purity, ratio of calcium/magnesium, texture, structure, and history.

The water table in limestone terrains which have undergone extensive karstification fluctuates greatly, rising in rapid response to precipitation input and falling due to rapid flow of water through solution passages to the discharge zones. Because there may be perched flow or storage on impermeable layers and along some bedding planes, some hydrologists question the validity of the water table concept in limestone terrains.

See also **Groundwater.**

Hydrology of Semiarid Regions

The semiarid regions of the earth's surface occur as transition zones between the arid deserts and the subhumid belts. Water movement will shape the landscape, according to its geology and past topography, and will work in conjunction with wind erosion, solar insolation, temperature changes, and soils (stable or in movement), as well as with the vegetation and the animals which live thereon, to produce an ecological balance of all factors, either in a temporary or a permanent sense.

In defining arid and semiarid areas, Peveril Meigs used only three factors: humidity, season of precipitation, and temperature. An extremely arid region is defined as an area with at least one entirely rainless month per year; arid regions are defined as regions where precipitation is less than potential evoporation. A typical semiarid region is defined as one in which precipitation occurs in a cold winter, with the coldest month in the 0–10°C average range and the hottest month in the 20–30°C range. These conditions are typical of Mediterranean semiarid climate, occurring in Morocco, Algeria, Lebanon, northern Iran and also on the western coast of the United States around latitude 35°N.

Arid and semiarid lands account for over one-third of the land surface of the earth, while cultivated lands account for but one-tenth of the whole. The greatest belt of arid and semiarid regions extends across North Africa as the Sahara, through the Arabian Peninsula with the "Empty Quarter" of extreme aridity, into the Salt Desert of Iran, and the Takla Makan of Central Asia. In North America, the Great American Desert falls generally within this classification.

Precipitation in the semiarid regions is restricted and kept low by the inability of moisture-bearing winds to penetrate into, and cool down within, such regions. Zones of high pressure may prevent the entry of winds, and the great desert areas are mainly associated with this meteorological phenomenon. Such winds as do enter arid and semiarid regions may have had no opportunity of acquiring moisture by passage over oceans or sea, or they may have been forced to lose their moisture in passing over high mountains, as in the "rainshadow" deserts of Imperial Valley, California and the Jordan-Syrian steppe. Again, lack of orogenetic effects within the regions, combined with high heat reflected from the ground, may prevent cooling of the incoming winds so that no moisture condenses to form clouds of precipitation, as in the coastal deserts of Chile, southern California, Morocco, and western Australia.

Almost all precipitation in the semiarid and arid regions occurs as rain, except in higher altitudes. Dew and even hoarfrost are also of importance, and are due to the great differences between day and night temperatures. Where infiltration conditions are good, as over coastal sand dunes, or suitable vegetation exists, such dew may make a permanent addition to the useful water resources of the area because of this type of moisture intake by certain vegetation.

Wind-wells have been reported in use in some areas of the Crimea and the south of France, wherein airborne moisture condenses on rapidly cooling stones. It has been reported that the Byzantines had irrigated vines by planting them at the base of an octahedron of open stones, the upper pyramid above ground surface to condense moisture and the lower inverted pyramid leading the condensate down to the vine root.

High evaporation and high transpiration in vegetated areas is the dominant hydrological characteristic of the semiarid and arid zones. Both transpiration and evaporation are high because abundant heat energy is supplied to change the limited amounts of liquid water into water vapor, either directly or through the biological processes. In this way, the heat balance of the area is maintained.

The inability of surface waters to maintain themselves against evaporation has a longterm effect in that it permits the formation of basins of inland drainage. Any basin of closed drainage will cease to exist if the average annual storage of surface water exceeds evaporation from its central lake system, for then the lake will rise and spread each year till it overtops the lowest point of the encircling water divide over which it will discharge to the ocean level and also cut its way down so as to reduce the size of the lake. Basins of closed drainage are characteristic of the semiarid lands of the world. The ability of the Nile, the Euphrates-Tigris, the Indus, the Colorado, and similar rivers, to keep open their basins is due to the fact that the amount of incoming surface waters (originating in nonarid regions) exceeds the evaporation losses.

Precipitation is never pure water, but contains salts and gases in solution. The salts are dissociated into cations, mainly calcium, magnesium, sodium, and potassium, while the anions are bicarbonate, chloride, and sulfate. Carbon dioxide is the principal dissolved gas. These elements in solution in precipitation may be of marine or terrestrial origin. It has been estimated that the annual precipitation of sea salts to be about three kilograms per hectare for the drier steppe regions south of the Sahara; as two kilograms per hectare in the Kalahari; and one kilogram per hectare for parts of Iraq and Iran. Full evaporation of the water which carries these salts will result in their deposition more or less where they fell; surface runoff will concentrate them in the central evaporating pans in basins of closed drainage, while infiltration to the aquifers may be with water which already is far from pure. Thus, in Syria, the chemical composition of the precipitation may be altered from a starting figure of about 20 ppm to concentrations of 100 to 200 ppm by evapotranspiration and leaching of precipitates in the zones of precipitation. Thus, D. J. Burdon and S. Mazloum report that the recharge waters to aquifers in Syria may contain from 50 to 200 ppm of total soluble salts.

In some aquifers, such as the Fars Formation of Iraq-Syria (of lagoonal facies), such soluble salts will be very abundant, while in other aquifers, such as the continental arkositic sandstones of the Sahara and Arabia, soluble minerals are almost completely absent. When the amount of groundwater flowing through the aquifer is large, such soluble salts tend to be removed and the aquifer flushed and cleaned out; likewise, fast-moving groundwater will flush an aquifer quicker than slow-moving water. Since the amount and often the rate of movement of groundwater in the semiarid zone tends to be small, mineralization by dissolution of the aquifer tends to be high.

At the point of natural discharge from aquifers, springs or marshy ground appears. If the spring is large, a perennial river carries off the discharge, and an oasis is formed, or else a large city, such as Damascus (fed by the Barada River flowing mainly from Ain Figeh) comes into existence. If the discharge is small or diffuse, a saline marsh tends to form, of which one of the greatest is the Qatarra Depression in Egypt, the probable discharge zone for the sandstone aquifer of the Western Desert of Egypt.

Serious studies and efforts have been underway for several ways to improve the overall hydrological conditions of semiarid zones, including: (1) Surface management—directed to making use of the water before it is lost by evaporation—increasing transportation through useful vegetation; (2) control of storage of surface runoff—often intensive and short-lived—possibly spreading the water over large areas by diverting it from the wadi or stream bed and controlling it behind earth banks in such a way that its flow velocity is never sufficient to erode the retaining structures; (3) control of aquifers—use of proper techniques in connection with extraction from galleries, wells, and bore holes; (4) underground storage of groundwater—borrowing some of the successful techniques used in storing surplus natural gas and oil underground; (5) possibly using weather modification techniques and through the construction of dams and other holding means to allow large surfaces of water to form upwind of the semiarid area under consideration (Would introduction of the Mediterranean Sea to the Qatarra depression increase precipitation along the Alexandrian coast?); and (6) desalting of brackish waters.

Hydrology of Volcanic Terrain

Volcanic terrains are made of rocks erupted from volcanos and intrusive rocks that congealed below the surface. The eruptives fall into three main categories:

1. Basalt, a dark-colored rock low in silica and high in ferromagnesian minerals;
2. Rhyolite, a light-colored rock high in silica and low in ferromagnesian minerals; and
3. A whole series of rocks of intermediate composition between basalt and rhyolite, such as andesite, dacite, latite, and trachyte.

The silica-rich magmas commonly are more explosive and tend to form steep cones close to their vents. Such vents include Mt. Lassen, California, Mt. Hood, Oregon, and Mt. Ranier, Washington. The basalt, being more fluid, forms plains, such as the Snake River Plain of Idaho and the Columbia River plateau of Washington, Oregon, and California. Basalts form the high islands of the Central Pacific, of which Mauna Loa and Kilauea volcanos on Hawaii are well known. If poured out molten, they are *flows*; if blown out, they are *pyroclastics*; if solidified in cracks or other voids in the crust, they are *intrusives*; if deposited as fragments in a vent, they are *throat breccias*.

The hydrology of volcanic terrain depends largely upon the permeability of the rocks present. Of the nearly 70 first-magnitude springs in the United States that discharge more than 100 cubic feet/second (2.83 cubic meters/second), 36 of these issue from basalt. Big Springs, Idaho, near Yellowstone Park, discharges about 180 cubic feet/second (5.1 cubic meters/second) from spherolitic obsidian at the terminous of a blocky silica-rich lava flow in an ancient caldera. Several other large springs issue in the same area from silicious lavas, presumably filling ancient valleys.

During the 1950s, wells were developed that were sufficient to irrigate large areas of pineapple on the island of Lanai, Hawaii. Also, large quantities of water have been developed in the Hawaiian islands by tunnels penetrating dike complexes. A large development of groundwater from basalt has been done by way of wells, tunnels, and shafts on Oahu.

Additional Reading

Brown, A. C., and A. McLachlan: "Ecology of Sandy Shores," Elseiver, Amsterdam, 1990.

Czaya, E.: "Rivers of the World," Van Nostrand Reinhold, New York, 1982.

Dean, R. G.: "Managing Sand and Preserving Shorelines," *Oceanus*, 49 (Fall 1988).

Giese, G. S., and D. G. Aubrey: "Losing Coastal Upland to Relative Sea-Level Rise: 3 Scenarios for Massachusetts," *Oceanus*, 16 (Fall 1987).

Goff, J. C., and B. P. J. Williams, Editors: "Fluid Flow in Sedimentary Basins and Aquifers," Blackwell Scientific, Palo Alto, California, 1987.

Gore, R., et al.: "Between Monterey Tides," *Nat'l. Geographic*, 2 (February 1990).

Hamilton, D. P.: "Death of the Nile Delta?" *Science*, 1084 (November 23, 1990).

Hauck, G. F. W.: "The Roman Aqueduct of Nîmes," *Sci. Amer.*, 98 (March 1989).

Hey, R. D., Bathurst, J. C., and C. R. Thorne: "Gravel-Bed Rivers," Wiley, New York, 1982.

Lowenstein, F.: "The Rising Tide," *Technology Review (MIT)*, 17 (July 1988).

Raghunath, H. M.: "Ground Water Hydrology," Wiley, New York, 1982.

Seymour, R. J., Editor: "Nearshore Sediment Transport," Plenum, New York, 1989.

Smith, P. L., et al.: "Estuary Rehabilitation: The Green Bay Story," *Oceanus*, 12 (Fall 1988).

Stanley, D. J.: "Subsidence in the Northeastern Nile Delta: Rapid Rates, Possible Causes, and Consequences," *Science*, 407 (April 22, 1988).

HYDROLYMPH. The watery body fluid or blood of lower invertebrates. It carries nutriment to organs and tissues and removes waste; it has no respiratory function generally, though may contain proteins able to function as oxygen carriers.

HYDROLYSIS. A chemical reaction in which water reacts with another substance to form two or more substances. This involves ionization of the water molecules as well as splitting of the compound hydrolyzed, e.g., $CH_3COOC_2H_5 + H{\cdot}OH \rightarrow CH_3COOH + C_2H_5OH$. Examples are conversion of starch to glucose by water in the presence of suitable catalysts; or the conversion of sucrose (cane sugar) to glucose and fructose by reaction with water in the presence of an enzyme or acid catalyst; or conversion of natural fats into fatty acids and glycerin by reaction with water, as occurs in one stage of soap manufacturing; or the reaction of the ions of a dissolved salt to form various products, such as acids, complex ions, etc. See also **Cellulose Ester Plastics (Organic); Organic Chemistry;** and **Starch.**

HYDROMAGNETIC EQUATIONS. The time-dependent equations which describe the behavior of a plasma in a magnetic field, assuming that the plasma is a compressible fluid and the plasma pressure P is a scalar. These equations are:

$$\rho \frac{d\mathbf{V}}{dt} = \mathbf{j} \times \mathbf{B} + q\mathbf{E} - \nabla P + \rho \mathbf{g} \tag{1}$$

$$\nabla \cdot (\rho \mathbf{V}) = -\frac{\partial \rho}{\partial t} \tag{2}$$

$$\mathbf{E} + \frac{\mathbf{V}}{c} \times \mathbf{B} = \frac{1}{\sigma}(c\mathbf{j} - q\mathbf{V}) \tag{3}$$

$$\frac{1}{P}\frac{dP}{dt} = \frac{\gamma}{\rho}\frac{d\rho}{dt} \tag{4}$$

$$\nabla \times \mathbf{B} = 4\pi \mathbf{j} + \frac{1}{c}\frac{\partial \mathbf{E}}{\partial t} \tag{5}$$

$$\nabla \cdot \mathbf{B} = 0 \tag{6}$$

$$\nabla \times \mathbf{E} = -\frac{1}{c}\frac{\partial \mathbf{B}}{\partial t} \tag{7}$$

$$\nabla \cdot \mathbf{E} = 4\pi q \tag{8}$$

The first equation is a force equation including gravitational forces. The second equation is a statement of mass conservation, while the third is analogous to Ohm's law. Number four is a statement of the adiabatic condition of the motion where γ is the ratio of specific heats of the plasma. The next four equations are the familiar Maxwell equations with no distinction made for **B** and **H** and **D** and **E** because all currents and charges are treated explicitly. The electromagnetic quantities are given in mixed Gaussian units and the conductivity σ in esu.

HYDROMETER. See **Specific Gravity.**

HYDRONIUM ION. An ion found in water and all its solutions, which has the formula H_3O^+ and which consists of a proton combined with a water molecule. It has been established that hydrogen ions do not exist free in aqueous solution, but are present as hydronium ions. Formation of such ions is statistically rare, resulting from the interaction of water molecules in a ratio of 1 to 556 million.

HYDROPHILIC. Having a strong tendency to bind or absorb water, which results in swelling and formation of reversible gels. This property is characteristic of carbohydrates, such as algin, vegetable gums, pectins, starches, and of complex proteins, such as gelatin and collagen. See also **Colloid System;** and **Detergents.**

HYDROPHOBIC. Antagonistic to water; incapable of dissolving in water. This property is characteristic of oils, fats, waxes, and many resins, as well as of finely divided powders, such as carbon black and magnesium carbonate. Some interesting concepts are explored in "The Hydrophobic Effect and the Organization of Living Matter," by C. Tanford, *Science*, **200**, 1012–1018 (1978). See also **Colloid System.**

HYDROPHONE. A transducer which responds to water-borne sound waves and, if of electroacoustic design, produces equivalent electric waves as output. Types of hydrophones include:

Line Hydrophone. A directional hydrophone consisting of a single straight line element, or an array of contiguous or spaced electroacoustic transducing elements disposed on a straight line, or the acoustic equivalent of such an array.

Split Hydrophone. A directional hydrophone in which electroacoustic transducing elements are so divided and arranged that each division may induce a separate electromotive force between its own electric terminals.

Directional Hydrophone. A hydrophone the response of which varies significantly with the direction of incoming sound.

HYDROPHYTES. Sometimes called water plants, these plants can grow only where there is an abundance of water, essentially growing in water or saturated soil. Those hydrophytes which are flowering plants are probably those which have reverted to an aquatic habitat. The reverting land plants may first have become marsh plants and then gradually developed into definite hydrophytes.

An aqueous environment presents conditions far more constant than an aerial one does. In the tropics, such conditions permit the plants to grow throughout the year. In colder regions there is a definite winter period when growth must cease. Many hydrophytes of temperate regions merely sink to the bottom and remain dormant during the winter. Others accumulate food reserves in rhizomes, which remain rooted in the bottom and renew growth in the spring. Still others form winter buds, consisting of large apical buds surrounded by many closely packed leaves containing much reserve food material. A few hydrophytes form small tubers.

The stems of hydrophytes contain a very small amount of vascular tissue, since support is largely afforded by the water, and conduction is not a great problem. In many of these plants the stem is very porous so that the plant floats in the water. The leaves of hydrophytes are of two types. Submerged leaves are thin and of various shapes; some, like eel grass leaves, are long and ribbonlike; others, like bladderworts, are finely dissected; while others are reduced to awl-shaped structures of small size. Floating leaves are usually large, undivided, and with stomata on the upper surface.

Reproduction in hydrophytes occurs both asexually and sexually. The flowers of nearly all hydrophytes are wind and insect pollinated, apparently a hangover from the time when they lived on land. A few have become modified to such an extent that pollination takes place on the surface of the water, the pollen floating about thereon and eventually reaching the stigma. A small number of hydrophytes are pollinated under water.

Nearly all algae and many fungi are hydrophytes; so also are some of the higher plants. In the flowering plants there are many water plants, such as the water lilies, bladderwort, eel grass and pondweeds. Many are very interesting plants; several are aquarium plants, serving to oxygenate the water; few are of any economic value. See **Algae;** and **Fungus.**

HYDROPONICS. The soilless culture of plants. In this technique, plants are grown with their roots immersed in a solution containing the necessary mineral salts or rooted in a sand medium which is kept moistened with such a solution. In one version of the method, the plants are supported in a matrix of peat, excelsior or some similar material on a wire screen with their roots dipping into the solution below. Aeration of the solution must also be provided if the best results are to be obtained. In another method, the plants are rooted in a medium of sand, gravel, or some similar material contained in a shallow tank into which the solution is automatically pumped at suitable intervals. Between pumpings, the solution gradually drains back into a reservoir tank.

The elements known to be necessary in chemically detectable amounts for the development of plants are carbon, oxygen, hydrogen, nitrogen, phosphorus, sulfur, potassium, magnesium, calcium, iron, manganese, boron, copper, zinc, and perhaps molybdenum. The first three of these elements are obtained by the plant from atmospheric gases or from water absorbed from the soil. The others are all absorbed in the form of mineral salts from the soil. Of the elements absorbed as salts the iron, manganese, boron, copper, zinc, and molybdenum are required in relatively minute quantities and are often called micronutrient elements. The principal elements which must be provided in the form of dissolved salts in hydroponic techniques, therefore, are nitrogen, phosphorus, sulfur, potassium, calcium, and magnesium.

Numerous solutions have been devised for use in the solution or sand culture of plants on both large and small scales. One solution which has been widely and successfully used for such purposes is made as follows: To each liter of water (preferably distilled or rain) add 1 M solution of the following salts as indicated: 1 cubic centimeter KH_2PO_4, 5 cubic centimeters KNO_3, 5 cubic centimeters $Ca(NO_3)_2$, and 2 cubic centimeters $MgSO_4$. To this solution then add 1 cubic centimeter per liter of a solution of micronutrients made as follows: 2.5 grams H_3BO_3, 1.8 grams $MnCl_2 \cdot 4H_2O$, 0.1 gram $ZnCl_2$, 0.05 gram $CuCl_2 \cdot 2H_2O$, and

0.075 gram MoO_3 per liter of distilled water. Also add to each liter of the solution made as described first, 1 cubic centimeter of a 0.5% solution of iron tartrate. The solution must be replaced with a fresh one at suitable intervals, and it is often necessary to add more of the iron solution between replacements.

Crop yields of at least some kinds of plants fully equal to those obtained on fertile soils can be obtained by hydroponic methods. The raising of crops by this method, however, probably will prove to be economically sound only for certain intensive types of agriculture or under certain special conditions. Some greenhouse floricultural and horticultural crops are now being grown successfully by this method. In regions where there is no soil, or where the soil is extremely infertile, but in which the climate is suitable to the development of plants, it seems likely that hydroponic techniques may prove useful. They have been used with some success, for example, on some of the coral islands of the Pacific Ocean.

See also **Aquaculture.**

HYDROSPHERE. The discontinuous envelope of water, both fresh and salt, which covers a major portion of the lithosphere. The bulk of the hydrosphere is contained within the deeper depressions of the surface of the earth. These depressions are termed ocean basins, and the water within them, oceans. Since the ocean basins are not large enough to hold the entire hydrosphere, seas are formed by the overflow of the oceanic waters on the continents. Geologists classify seas as epicontinental (epeiric) or relict. Technically, lakes, rivers and underground waters are also part of the hydrosphere. In general, therefore, the term hydrosphere is used mainly to distinguish the watery covering of the earth from the lithosphere on which, and in which, in part, it rests. See also **Earth; Hydrology; Ocean;** and **Polar Research.**

HYDROSTATIC PRESSURE. The pressure created by a superimposed layer of a liquid is hydrostatic pressure. The intensity of hydrostatic pressure is commonly expressed as pounds per square inch. A head of 2.31 feet of fresh water creates a hydrostatic pressure of 1 pound per square inch. At a given depth of immersion in water, the pressure acts with equal intensity in all directions, that is, hydrostatic pressure is not directional in effect. Hydrostatic pressures are measured by means of pressure gauges of the Bourden tube type, or manometers of the U-tube type. Hydrostatic pressure sometimes is referred to as hydrostatic head.

In compressible flow, the hydrostatic pressure must be defined more carefully. A suitable definition is in terms of the Helmholtz free energy, $A = U - TS$,

$$p = -\left(\frac{\partial A}{\partial(1/\rho)}\right)_T$$

HYDROSTATICS. This branch of mechanics has to do with the equilibrium of liquids and the laws relating to liquid pressure. A study of these laws makes it clear that the components of pressure in a liquid fall naturally into two classes, according to the way in which they are produced; namely, (1) pressures due to forces applied externally, as by the atmosphere or by the piston of a pump, and (2) those due to causes operating throughout the body of liquid, such as gravity or inertia.

Pascal's law applied only to the first class, and states that any pressure in an enclosed liquid, originating in forces applied at its boundary, is communicated with unaltered intensity to all parts of the liquid. A familiar illustration of this fundamental law is the hydraulic press, which consists of two communicating cylinders, usually of different diameter, fitted with pistons, the force acting upon one piston and the force exerted by the other being in proportion to their areas.

The pressure in an enclosed liquid due to its own weight, on the other hand, increases uniformly with the depth below its highest point, and is equal to the product of the depth by the weight per unit volume. For fresh water, the pressure at depth h feet is 62.4 h pounds per square foot. The total pressure of water against a submerged plane are area is equal to the intensity of pressure on the center of gravity times the area.

Problems of flotation, draft, and buoyant stability always involve the density of the liquid and the volume and shape of the floating object. A floating body of mass m in a liquid of density ρ, will float with a volume v submerged, v determined by the relationship $v = (m/\rho)$. That is, a floating body displaces a volume of liquid having the same weight as the body.

The buoyancy may be said to be the force which is equivalent to the weight of the liquid displaced by the submerged portion of the floating object. Buoyancy and weight do not, in general, act in the same vertical line. The weight acts at the center of gravity of the floating object, the buoyancy at the center of gravity of the displaced liquid, called the center of buoyancy. The relative positions of the buoyancy and the weight when a floating object is disturbed from an upright floating position, determines whether it is stable or unstable flotation. If the vertical drawn through the center of buoyancy passes above the center of gravity of the body, there is a righting moment, and the body is stable, whereas if it passes below the center of gravity, it is unstable in that the buoyancy tends to tip the object still further. The intersection of the line of buoyancy with the axis of symmetry of the floating body is the metacenter, and the distance from the metacenter to the center of gravity is the metacentric height. The latter is used to measure the stability of a hull. Another case of flotation is illustrated by the balance of the hydrometer. It is apparent from the above that with a given weight, the volume of immersion varies inversely with the density of the liquid. In other words, a floating body rides higher in a denser liquid. This fact is put to use in the hydrometer, which has a given weight and which is immersed in fluids to measure their density. The hydrometer is calibrated to read the volume submerged directly in terms of density of the liquid.

An important general principle of hydrostatics is that which determines the free liquid surface in equilibrium. The direction of the surface at any point is perpendicular to the resultant of all forces acting upon a particle at that point. Thus, if only gravity is acting, the surface is horizontal or "level"; but if there are capillary forces, or if the external pressure is not uniform, the surface is inclined. An interesting case is that of a liquid rotating uniformly in a cylindrical tub; the surface then assumes the form of a paraboloid of revolution, symmetrical about the vertical axis.

Hydrostatic Equation. The form assumed by the vertical component of the vector equation of fluid motion when Coriolis, earth curvature, frictional, and vertical acceleration terms are considered negligible compared with those involving the vertical pressure force and the force of gravity. Thus,

$$\frac{\partial p}{\partial z} = -\rho g$$

where p is the pressure, ρ the density, g the acceleration of gravity, and z the geometric height.

In compressible flow, the hydrostatic pressure must be defined more carefully. A suitable definition is in terms of the Helmholtz free energy, $A = U - TS$,

$$p = -\left(\frac{\partial A}{\partial(1/\rho)}\right)_T$$

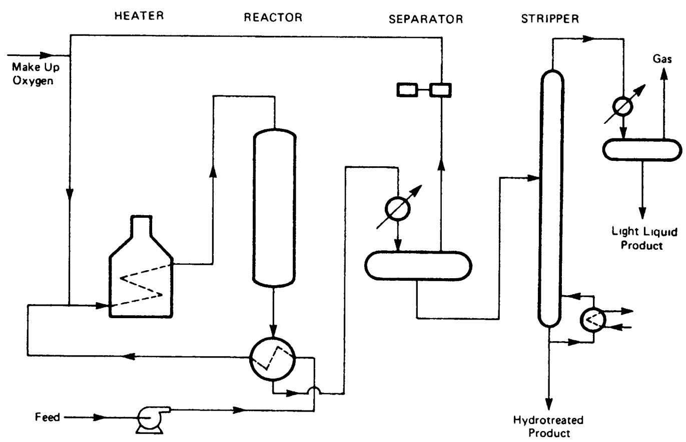
Hydrostatic Equilibrium. The state of a fluid whose surfaces of constant pressure and constant mass (or density) coincide and are horizontal throughout. Complete balance exists between the force of gravity and the pressure force. The relation between the pressure and the geometric height is given by the hydrostatic equation. The analysis of atmospheric stability has been developed most completely for an atmosphere in hydrostatic equilibrium.

Hydrostatic Pressure. The pressure created by a superimposed layer of liquid is hydrostatic pressure. See **Hydrostatic Pressure.**

HYDROTREATING. A specialized kind of hydrogenation in which the quality of liquid hydrocarbon streams is improved by subjecting them to mild or severe conditions of hydrogen pressure in the presence of a catalyst. The objective is to convert undesirable material in the feedstock to either desired materials or easily disposed byproducts, on a highly selective basis. As of the early 1980s about 45% of the crude oil refined in the United States is hydrotreated. Some applications of hydrotreating include: (1) improvement of the burning quality of jet fuels, kerosines, and diesel fuels; (2) purification of light aromatic byproducts from pyrolysis operations; (3) pretreatment of naphtha feeds for catalytic reforming units; (4) reduction in sulfur content of residual fuel oils; (5) pretreatment of catalytic cracking feeds and cycle oils by removal of metals, sulfur, nitrogen, and reduction of polycyclic aromatics; (6) desulfurization of distillate fuels; (7) upgrading of lubricating oil quality; and (8) improvement of color, odor, and storage stability of various fuels.

Some of the specific reactions involved include: (1) hydrogenation of monoaromatics to naphthenes to improve burning quality of certain fuels; (2) removal of nitrogen as ammonia from its organic combinations; (3) removal of oxygen from its organic combinations as water; (4) hydrogenation of polycyclic aromatics so that only one aromatic ring remains in the molecule; (5) hydrogenation of diolefins and olefins to paraffins or naphthenes; (6) removal of sulfur from its organic combinations in various types of sulfur compounds by hydrodesulfurization to form hydrogen sulfide; and (7) decomposition and removal of organometals, such as arsenic compounds in naphthas, by retention of these metals on the catalyst. Vanadium and nickel also can be removed.

In the hydrotreating process shown by the accompanying diagram, the liquid feed is preheated by exchange with the reactor effluent. It is then heated to the desired reactor-inlet temperature in a fired heater. At this point, recycle hydrogen joins the feedstock. An excess of hydrogen is used to suppress accumulation of deactivating carbonaceous deposits on the catalyst. Fresh makeup hydrogen enters the process to maintain a sufficient supply and also pressure on the system. Cooled effluent from the reactor goes to a separator vessel at which point the recycle or net hydrogen is removed. The liquid then goes to a stripper or stabilizer where hydrogen, hydrogen sulfide, ammonia, water, and light hydrocarbons dissolved in the separator liquid are removed. The stabilized hydro-treated liquid, free of dissolved, unwanted contaminants, is routed to subsequent processing or to product fuel blending.



Representative hydrotreating unit used in petroleum industry. (*UOP Inc.*)

It is interesting to note that there are over 25 proprietary versions of this basic process. Numerous modifications are required, depending upon the nature of the feedstock and desired end-products.

Technical Staff, UOP Inc., Des Plaines, Illinois.

HYDROXYLAMINE. H_2NOH, formula weight 33.02, white, odorless solid, mp 33°C, bp 56°C (22 mm pressure), explosive, soluble in all proportions in H_2O or alcohol. Hydroxylamine is: (1) A weak base forming with acids soluble salts that decompose more or less violently when heated, e.g., hydroxylamine hydrochloride (hydroxylammonium chloride, $H_2NOH \cdot HCl$), mp 151°C, nitrate $H_2NOH \cdot HNO_3$, hemisulfate $H_2NOH \cdot \frac{1}{2}H_2SO_4$. Dihydroxylamine oxalate and trihydroxylamine phosphate are insoluble in H_2O. Hydroxylamine hydrochloride is soluble in alcohol. (2) A weak acid forming with bases soluble salts, e.g., sodium hydroxylamite H_2NONa. Hydroxylamine salt solution is a powerful reducing agent, more especially in alkaline than in acid solution, for example, cupric salt solutions changed to cuprous oxide, silver salt solutions to silver, mercuric chloride solution to mercurous chloride, ferric salt solutions (in acid) to ferrous. Ferrous hydroxide in sodium hydroxide is, however, oxidized by hydroxylamine to ferric hydroxide plus NH_3.

Hydroxylamine reacts with carbonyl group =CO of aldehydes, ketones or quinones, yielding *oximes*, white solids, of definite melting point and used in identification of aldehydes and ketones, e.g., acetaldehyde oxime $CH_3CH{:}NOH$:

Beta-phenylhydroxylamine, N-phenylhydroxylamine, is a white solid, slightly soluble in water, very soluble in alcohol or ether, forms salts with acids, e.g., beta-phenylhydroxylamine hydrochloride $C_6H_5NHOH \cdot HCl$, upon exposure to air the water solution forms azobenzene $C_6H_5N{:}NC_6H_5$. Beta-phenylhydroxylamine reacts (1) with oxidizing agents, such as chromic acid or ferric chloride, to form nitrosobenzene C_6H_5NO, (2) with reducing agents, such as tin plus hydrochloric acid, to form aniline $C_6H_5NH_2$, (3) with alkaline cupric salt solution (Fehling's solution) at room temperature to form cuprous oxide, (4) with ammonio-silver salt solution (Tollen's solution) at room temperature to form silver, (5) in the presence of hydrochloric acid to form paraminophenol $HO \cdot C_6H_4 \cdot NH_2(1,4)$.

Beta-phenylhydroxylamine is formed by reduction of nitrobenzene (1) by zinc and calcium chloride or ammonium chloride solution, (2) by electrolysis in acetic acid plus sodium acetate solution.

Diphenylhydroxylamine is prepared by reaction of nitrosobenzene and phenylmagnesium bromide in anhydrous ether, followed by treatment with H_2O (magnesium hydroxybromide also formed).

When hydroxylamine reacts with aldehydes, the resulting compounds are termed *aldoximes* as, for example, acetaldoxime. $CH_3 \cdot CHO + H_2NOH \rightarrow CH_3 \cdot CH{:}N \cdot OH$ (acetaldoxime) + H_2O. Hydroxylamine reactions with ketones produce *ketoximes*. $(CH_3)_2CO + H_2NOH \rightarrow (CH_3)_2C{:}\,N \cdot OH$ (dimethylketoxime) + H_2O.

The lower aldoximes are essentially odorless, volatile liquids, and miscible with H_2O in all proportions. The higher members are only slightly soluble. Ketoximes have similar properties.

HYDROZOA. A class of the phylum *Coelenterata* composed chiefly of small animals without common names. Many species are colonial and the colonies of a few, such as the Portuguese man-of-war, are quite large.

The class differs from the other coelenterates in the occurrence of both hydroid polyps and medusae in the same species, usually in alternating generations. In many colonies, additional specialization occurs among the polyps for the performance of different functions; the gonozooids, gastrozooids, and dactylozooids of the siphonophores are such individuals. Hydrozoan medusae differ from jellyfishes and are called medusoids. In some cases, they are specialized forms which remain attached to the colony and show no resemblance to medusae, but the free-swimming forms differ from medusae only in details of structure. The medusoids are sexual reproductive individuals.

Relatively few species of hydrozoans live in fresh water. *Hydra*, the most widely known genus, includes a number of species without a medusa stage. The polyps are solitary and carry on both asexual and sexual reproduction. Several freshwater medusae for which no polyp stage has been discovered are also known from lakes in Europe, Africa, and the Americas. The marine species are numerous.

The class can be conveniently classified as follows:

Order 1. *Hydroidea*. Fixed zoophyte stage.
  Suborder a. *Anthomedusae* (*Athecata*). Polyps and reproductive zooids not protected.
  Suborder b. *Leptomedusae* (*Thecata*). Polyps protected by hydrothecae and reproductive zooids by gonothecase.
Order 2. *Hydrocorallina*. Massive skeleton of calcium carbonate secreted from coenosare—Hydroid corals.
Order 3. *Trachylinae*. No fixed zoophyte stage, all members being locomotive medusae.
  Suborder a. *Trachymedusae*. Tentacles from margin of umbrella and gonads develop in connection in the radial canals.
  Suborder b. *Narcomedusae*. Tentacles from ex-umbrella away from margin and gonads develop in connection with the manabrium.
Order 4. *Siphonophora*. Pelagic forms; colony usually exhibits polymorphism of its zooids.

HYENA (*Mammalia, Carnivora*). A large animal slightly resembling a wolf, but more closely related to the civets. The Aard-Wolf (*Protelinae*) and the Hyena (*Hyaeninae*) make up a small, special grouping in the order of *Carnivora*. The aard-wolf is believed by some authorities to bridge the gap between the Hyaenines and the Viverrines. See also **Viverrines**. Some years ago, investigators believed the aard-wolf was a type of civet and belonged with the Viverrines. This animal is difficult to describe, appearing something like a clumsy dog, with a rather fox-like face, with wooly hair along the back in a form of a permanently erected crest, giving it something of a skunk-like appearance. The aard-wolf differs from the hyena, in that it has five toes on the forefeet; four on the hind feet. Because the animal's teeth are widely set and reduced in number and size, the principal diet is comprised of insects or very decomposed meat or newly born animals. The aard-wolf is found uncommonly in eastern Africa (north of the Kalahari Desert) and on the west coast of Africa as far north as Angola.

At one time, hyenas ranged over much of Europe in climes south of Scotland and the Scandinavian countries and reached eastward through eastern Europe and central Asia. They are found today in most parts of Africa south of the great deserts.

The Striped Hyena (*Hyaena*) is found in Africa as described, as well as in fairly large numbers in northern India. It is also characterized by a crest along its back and by striping on the flanks, with cross-stripes on the legs. The animal is sturdily built, with massive head and somewhat disproportionately long front legs.

The Spotted Hyena (*Crocuta*) is larger than the striped hyena and its limbs are better proportioned. The animal is extremely powerful and can crack large bones, including those of the elephant and hippopotamus. It can put up a good fight with the Big Cats. The striped hyena is considered of very poor disposition, sometimes described as sneaky and generally unpleasant. However, except when in danger, it is described as cowardly. In nature, the animal is extremely dirty, carrying around a most offensive odor. Surprisingly, however, the animal is reported to make an excellent, docile, and trustworthy pet if taken young and trained.

The hyenas are known for their bloodcurdling howling and noise like an insane laugh which usually occurs at the end of a barking streak. On flat ground, their speed exceeds that of a horse.

HYGROMETRY AND PSYCHROMETRY. These are instrumental methods for measuring humidity. Humidity can be expressed in a variety of different forms: wet bulb temperature; percent relative humidity (% RH); vapor pressure; mixing ratio; dew/frost point; grains per pound; grams per kilogram; and parts per million, among others. These parameters can be measured by a number of different instruments, each capable of accurate measurement under certain conditions and within specific limitations.

Definition of Humidity. Unless one is routinely working with humidity measurements, there is a tendency to overlook the fact that humidity is water gas, behaving in accordance with the ideal gas laws. One of the easiest ways to put humidity in its proper perspective is through application of Dalton's law of partial pressures to the most commonly encountered gas—*air*.

Dalton's law states that the total pressure P_m exerted by a mixture of gases or vapors is the sum of the pressure of each gas if it were to occupy the same volume by itself. The pressure of each individual gas is called its *partial pressure*. The total pressure of an air-water gas mixture, containing oxygen, nitrogen, and water, is equal to the sum of the partial pressures of each gas:

$$P_m = P_{N_2} + P_{O_2} + P_{H_2O} + \cdots$$

Therefore, the partial pressure of water vapor in air is directly related to the measurement of humidity. This vapor pressure varies from 1.22×10^{-3} mb (millibar) of mercury (0.122 Pascal) at the −75°C frost point of "bone dry" arctic or industrial dry air—to 1.013×10^3 mb of mercury (0.1013×10^6 Pascal) at the 100°C dew point of saturated hot air in a product drier. This is a change of almost a million to one over the span of interest in industrial humidity measurement.

The ideal humidity instrument would be a linear, wide-range pressure gage, specific to water vapor and employing a primary or fundamental measuring method. Such an instrument, although physically possible, would be cumbersome. Most humidity measurements are made by some secondary instrument which is responsive to humidity-related phenomena.

Humidity Parameters. The humidity parameters most often encountered in scientific and industrial applications are given in the accompanying table. In addition to these common parameters, numerous other formats exist for use in narrow applications or specific technologies. However, most of these are variations of the parameters listed.

The psychrometric chart provides a quick means for converting from one humidity format to another because dew point, relative humidity, ambient temperature, and wet bulb temperature can be conveniently related to each other on a single sheet of paper. The psychrometric chart has long been the basic tool of air conditioning engineers. A chart of this type is given in the entry entitled **Psychrometric Chart.** Psychrometric charts are available for higher temperatures and humidities and are quite useful in drier and condensation system design. Charts are also available for lower temperatures, but these tend to be less useful because wet bulb measurements are difficult to make with any accuracy at temperatures below –7°C.

HUMIDITY MEASUREMENT METHODS

Parameter	Description	Units	Typical Applications
Wet bulb temperature	Minimum temperature reached by a wetted thermometer in an airstream	°F or °C	High temperature driers, air conditioning, meteorology, test chambers
Percent relative humidity	The ratio of the actual vapor pressure to the saturation vapor pressure, with respect to water, at the prevailing dry bulb temperature	0–100%	Monitoring conditioning rooms, test chambers, pharmaceutical and food packaging
Dew/frost point	Dew point is the temperature to which the air must be cooled to achieve saturation. If the temperature is below 32°F, it is called the frost point	°F or °C	Heat treating, annealing atmospheres, drier control, instrument air monitoring, meteorological/environmental measurements
Volume or mass ratio	Parts per million (ppm) by volume is the ratio of the partial pressure of the water vapor to the partial pressure of the dry carrier gas. PPM by weight is identical to ppm by volume, but the ratio changes according to the molecular weight of the carrier gas.	ppm_v, ppm_w	Used primarily to insure dryness of industrial process gases such as air, nitrogen, oxygen, methane, hydrogen, etc.

Wet Bulb/Dry Bulb Measurements

Psychrometry has long been a popular method for monitoring humidity, primarily due to its simplicity and inherent low cost. A typical industrial psychrometer consists of a pair of matched electrical thermometers, one of which is maintained in a wetted condition. Water evaporation cools the wetted thermometer, resulting in a measurable difference between it and the ambient, or dry bulb measurement. When the wet bulb reaches its maximum temperature depression, the humidity is determined by comparing the wet bulb/dry bulb temperatures on a psychrometric chart. In a properly designed psychrometer, both sensors are aspirated at an airstream rate between 4 and 10 meters per second for proper cooling of the wet bulb, and both are thermally shielded to minimize errors from radiation.

A properly designed and utilized psychrometer, such as the Assman laboratory type, is capable of providing accurate data. However, very few industrial psychrometers meet these criteria and are limited to applications where low cost and moderate accuracy are the underlying requirements. The psychrometer does have certain inherent advantages: (1) The psychrometer is capable of highest accuracy near 100% RH. From an accuracy standpoint, it is superior to most other humidity sensors near saturation. Since the dry bulb and wet bulb sensors can be connected differentially, this allows the wet bulb depression (which approaches zero as the relative humidity approaches 100%) to be measured with a minimum of error. (2) Although large errors can occur if the wet bulb becomes contaminated or improperly fitted, the simplicity of the device affords easy repair at minimum cost. (3) The psychrometer can be used at ambient temperature above 100°C, and the wet bulb measurement is usable up to 100°C.

Major limitations of the psychrometer include: (1) As relative humidity drops below about 20% RH, the problem of cooling the wet bulb to its full depression becomes difficult. The result is impaired accuracy below 20% RH, and few psychrometers work at all below 10% RH. (2) Wet bulb measurement at temperatures below 0°C are difficult to obtain with any high degree of confidence. Automatic water feeds are not feasible, because of freezing. (3) Because a wet bulb psychrometer is a source of moisture, it can only be used in environments where added water vapor from the psychrometer exhaust is not a significant component of the total volume. (4) Generally speaking, psychrometers cannot be used in small, closed volumes.

Percent Relative Humidity

Percent relative humidity is the best known and perhaps the most widely used method for expressing the water vapor content of air. Percent relative humidity is defined as the ratio of the prevailing water vapor pressure e_a to the water vapor pressure if the air were saturated, e_s, multiplied by 100:

$$\% \text{ RH} = (e_a/e_s) \times 100$$

The term "percent relative humidity" appears to be derived from the invention of the hair hygrometer in the 17th Century. The hair hygrometer operates on the principle that many organic filaments, such as hair, goldbeater's skin, and even nylon, change length as a nearly linear function of the *ratio* of *prevailing water vapor pressure* to the *saturation vapor pressure*.

Basically, percent relative humidity is an indicator of the water vapor saturation deficit of the gas mixture, rather than an indicator of sorption, desorption, comfort, or evaporation. A measurement of % RH without a corresponding measurement of dry bulb temperature is not of particular value, since the water vapor content cannot be determined from % RH alone.

Sensors for Measuring % RH. Over the years, devices other than the simple hair hygrometer have evolved which permit a direct measurement of % RH. These devices are, for the most part, electrochemical sensors which offer a degree of ruggedness, compactness, and remote electronic readout ability not afforded by hair devices.

Two widely used electronic % RH sensors are the Dunmore element and the Pope cell. The Dunmore sensor employs a bifilar-wound inert wire grid on an insulative substrate which is coated with a lithium chloride solution of a controlled concentration. The hygroscopic nature of this salt causes it to take up water vapor from the surrounding atmosphere. The ac resistance of the sensor is an indication of the prevailing % RH. Dunmore cells are excellent RH sensors, but, because of the characteristics of lithium chloride, are usually designed to cover a narrow range of interest. For example, a single sensor may cover from 40 to 60% RH and the sensor output is usable only in that range. See Fig. 1(a).

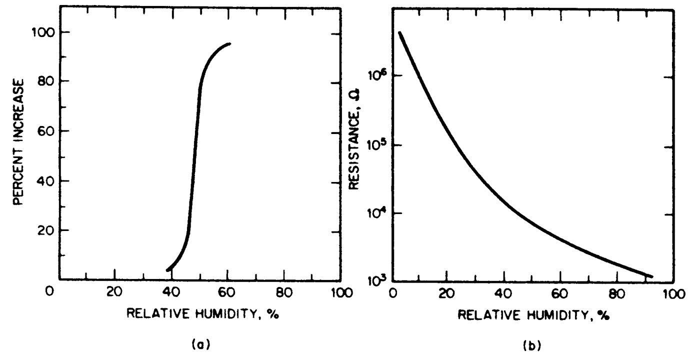


Fig. 1. Resistance characteristics of typical Dunmore and Pope sensors. (a) Dunmore sensors are limited to a narrow range of humidity. This sensor operates between 40 and 60% RH. (b) Pope sensors operate over a wide humidity range, but output impedance of the sensor varies from 1000 ohms (100% RH) to several megohms (10% RH), which complicates readout circuitry.

Wide-range Dunmore sensors can be made with a cluster of narrow range sensors in a common housing, mated with an electrical matching network. This arrangement, however, usually results in a rather bulky sensor.

The Pope cell employs a similar bifilar conductive grid on an insulative substrate. In this sensor, the substrate is made from polystyrene, which has been treated in a prescribed fashion with sulfuric acid. This results in sulfonation of the longer-chain polystyrene molecules. Because the sulfate radical (SO_4) is highly mobile in the presence of hydrogen ions (available from the water molecule in vapor form), the $(SO_4)^{2-}$ ions can detach and take on H^+ ions, thereby altering the surface resistivity of the sensor as a function of humidity.

In both the Dunmore and Pope sensors, the element is arranged in an ac-excited Wheatstone bridge so that only alternating current flows through the grid. Direct current excitation of either the Dunmore or Pope elements polarizes the sensor, eventually causing loss of calibration.

The Pope sensor has one significant advantage over the Dunmore sensor in that the Pope unit is a wide-range sensor, typically covering 15% RH to 99% RH in a single element. See Fig. 1(b). Considerable attention must be given to readout circuitry for the Pope sensor because the resistance varies in a nonlinear fashion from 1000 ohms to several megohms.

Dew Point Hygrometry

Dew point measurements are widely used in scientific and industrial applications when precise measurement of water vapor pressure is needed. Dew point, the temperature at which water condensate begins to form on a surface, can be accurately measured from −75°C to +100°C across the entire range of humidity with a condensation (chilled mirror) hygrometer.

Three types of instruments have received wide acceptance in dew point measurements: (1) the saturated salt dew point sensor; (2) the condensation-type hygrometer; and (3) the aluminum oxide sensor. Many other instruments are used in specialized applications, including pressure ratio devices, dewcups, and fog chambers. The latter are manually operated.

Saturated Salt Dew Point Sensors. The saturated salt (lithium chloride) dew point sensor is widely used because of its inherent simplicity, ruggedness, and low cost. Both the United States and Canadian government weather services use this type of sensor for most official groundbased humidity measurements. However, some of these are being converted to the more accurate condensation hygrometers.

The principle of the saturated salt dew point sensor is based on the relationship that the vapor pressure of water is reduced in the presence of a salt. When water vapor in the air condenses on a soluble salt, it forms a saturated layer on the surface of the salt. This saturated layer has a lower vapor pressure than water vapor in the surrounding air. If the salt is heated, its vapor pressure increases to a point where it matches the water vapor pressure in the surrounding air and the evaporation/condensation process reaches equilibrium. The temperature at which equilibrium is reached is directly related to the dew point.

A saturated salt sensor is constructed with an absorbent fabric bobbin covered with a bifilar winding of inert electrodes and coated with a dilute solution of lithium chloride. Lithium chloride (LiCl) is often used as the saturating salt because of its hygroscopic nature, which permits application in relative humidities between 11 and 100%.

An alternating current is passed through the winding and salt solution, causing resistive heating. As the bobbin heats, water evaporates into the surrounding air from the diluted LiCl solution. The rate of evaporation is determined by the vapor pressure of water in the surrounding air. When the bobbin begins to dry out, due to evaporation of water, resistance of the salt solution increases. With less current through the winding, because of increased resistance, the bobbin cools and water begins to condense, forming a saturated solution on the bobbin surface. Eventually, equilibrium is reached and the bobbin neither takes on nor loses any water.

Properly used, a saturated salt sensor is accurate to ±1°C between dew point temperatures of −12 and +38°C. Outside these limits, small errors may occur as a result of the multiple hydration characteristics of lithium chloride, which may produce ambiguous results at 41°C, −12°C, and −34°C dew points. Maximum errors at these ambiguity points are 1.4, 1.6, and 3.4°C, respectively, but actual errors encountered in typical applications are usually less.

Applications. The saturated salt sensor has certain advantages over other electrical humidity sensors, such as % RH instruments. Because the salt sensor operates as a current carrier saturated with Li and Cl ions, addition of contaminating ions has little effect on its behavior compared to a typical RH sensor, which operates "starved" of ions and is easily contaminated. A properly designed saturated salt sensor is not easily contaminated since, from an ionic standpoint, it can be considered precontaminated.

If a saturated salt sensor does become contaminated, it can be washed with an ordinary sudsy ammonia solution, rinsed and recharged with lithium chloride. It is seldom necessary to discard a saturated salt sensor if proper maintenance procedures are observed.

Limitations of saturated salt sensors include: (1) relatively slow response time; and (2) a lower limit to the measurement range imposed by the nature of lithium chloride. The sensor cannot be used to measure dew points when the vapor pressure of water is below the saturation vapor pressure of lithium chloride, which occurs at about 11% RH. In certain gases, ambient temperatures can be reduced, increasing the RH to above 11%; but the extra effort needed to cool the gas usually warrants selection of a different type of sensor. Fortunately, a large number of scientific and industrial measurements fall above this limitation and are readily handled by the sensor.

Condensation-Type Hygrometers. The condensation-type dew point hygrometer is one of the most accurate and reliable of sensors for humidity measurements, and has the widest range. These features are achieved, however, through increased complexity and cost. In the condensation-type hygrometer, a surface is cooled (either thermoelectrically, mechanically, or chemically) until dew or frost begins to condense out. The condensate surface is maintained electronically in vapor pressure equilibrium with the surrounding gas, while surface condensation is detected by optical, electrical, or nuclear techniques. See Fig. 2. The surface temperature is then the dew point temperature, by definition.

The largest source of error in a condensation hygrometer stems from the difficulty in measuring condensate surface temperature accurately. Typical industrial versions of the instrument are accurate to ±0.2°C over very wide temperature spans. Laboratory models offer accuracies up to ±0.1°C.

Wide span and minimal errors are two main features. A properly designed condensation hygrometer can measure dew points from 100°C down to frost points of −75°C.

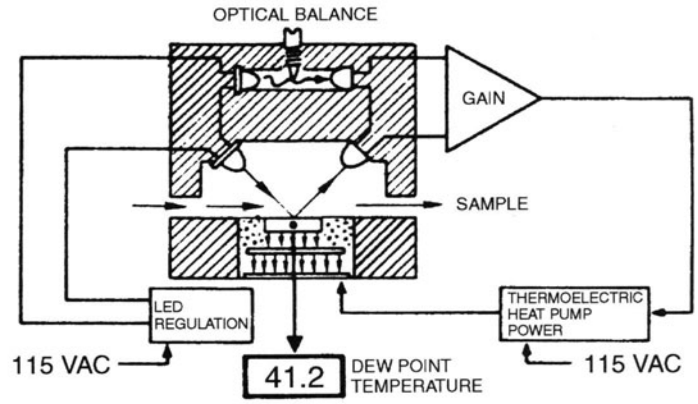


Fig. 2. Dew is detected in a condensation hygrometer by cooling a surface until water begins to condense. Condensation is detected optically or electronically. The signal is fed into a control circuit which maintains the surface temperature at the precise dew point.

Response time of a condensation dew point hygrometer is usually specified in terms of its cooling/heating rate, typically 1.5°/second, making it considerably faster than a saturated salt dew point sensor and nearly as fast as most electrical % RH sensors. Perhaps the most significant feature of the condensation hygrometer is its fundamental measuring technique, which essentially renders the instrument self-calibrating. For calibration, it is only necessary to manually override the surface-cooling control loop, causing the surface to heat, and witness that the instrument recools to the same dew point when the loop is closed. Assuming that the surface temperature measuring system is calibrated, this is a reasonable and valid check on the instrument's performance.

Because of its fundamental nature and superior accuracy and repeatability, this kind of instrument is widely used as a secondary standard (National Bureau of Standards) for calibrating other lower level humidity instruments.

The inert construction of the condensation hygrometer makes it virtually indestructible. Although the instrument can become contaminated, it is easy to wash and return to service without impairment of performance or calibration.

The condensation (chilled mirror) hygrometer measures dew/frost temperature. Unfortunately, many applications require measurement of % RH, water vapor in parts per million, or some other humidity parameter. In such cases, the user must decide whether to employ the fundamental, high accuracy condensation hygrometer and convert the dew/frost point measurement to the desired parameter, or use lower level instrumentation to measure these parameters directly. In recent years, microprocessors have been developed which can be incorporated in the design of a condensation hygrometer, resulting in instrumentation which can offer accurate measurements of humidity in terms of almost any humidity parameter.

Electrolytic Hygrometer. A typical electrolytic hygrometer utilizes a cell coated with a thin film of phosphorous pentoxide (P_2O_5), which absorbs water from the sample gas. See Fig. 3. The cell has a bifilar winding of inert electrodes on a fluorinated hydrocarbon capillary. Direct current applied to the electrodes dissociates the water, which is absorbed by the P_2O_5, into hydrogen and oxygen. Two electrons are required to electrolyze each water molecule and thus the current in the cell represents the number of molecules dissociated. A further calculation, based on flow rate, temperature and current, yields the parts per million concentration of water vapor.

In order to obtain accurate data, the flow rate of the sample gas through the cell must be known and constant. Since the ppm calculation is partially based on flow, an error in the flow rate causes a direct error in measurement.

A typical sampling system for insuring constant flow is shown in Fig. 4. Constant pressure is maintained within the cell. Sample gas enters

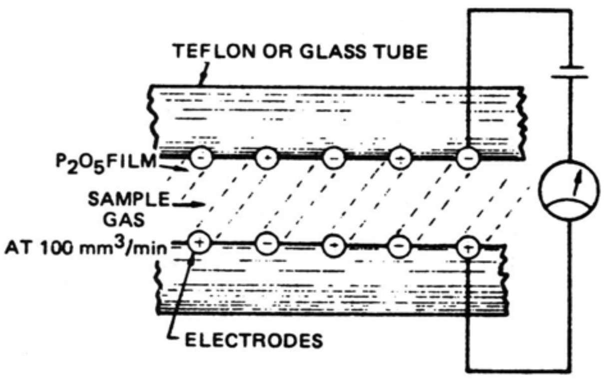


Fig. 3. An electrolytic hygrometer dissociates water, absorbed by P_2O_5, into hydrogen and oxygen by electrolysis. Since two electrons are required to electrolyze a molecule of water, the amount of current used by the hygrometer relates to parts per million of water vapor.

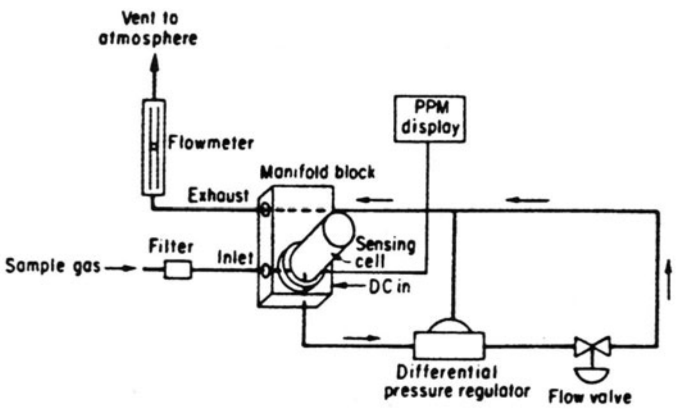


Fig. 4. Calculation of the water vapor content in an electrolytic hygrometer is dependent on precise control of the flow rate. This arrangement controls the sample pressure across the cell, ensuring correct flow regardless of input pressure fluctuations.

the inlet, passes through a stainless steel filter, and enters a stainless steel manifold block. It is very important that all components prior to the sensor be made of an inert material, such as stainless steel, to minimize contamination. After passing through the sensor, the sample gas pressure is controlled by a differential pressure regulator which compares pressure of the gas leaving the sensor with the pressure of the gas venting to atmosphere through a preset valve and flowmeter. In this way, constant flow is maintained even though there may be nominal pressure fluctuations at the inlet port.

A typical electrolytic hygrometer can cover a span from 0 to 2000 ppm with an accuracy of $\pm 5\%$ of the reading, more than adequate for most industrial applications. The sensor is suitable for most inert elemental gases and organic and inorganic gas compounds that do not react with P_2O_5.

Electrolytic hygrometers cannot be exposed to high water vapor levels for any long period of time because this results in a high usage rate for the P_2O_5 and high cell currents.

Aluminum Oxide Moisture Sensor. This type of sensor is a capacitor, formed by depositing a layer of porous aluminum oxide onto a conductive substrate, and then coating the oxide with a thin film of gold. The conductive base and the gold layer become the capacitor's electrodes. Water vapor penetrates the gold layer and is absorbed by the porous oxidation layer. The number of water molecules absorbed determines the electrical impedance of the capacity which is, in turn, a measure of water vapor pressure.

Advantages of the aluminum oxide sensor are: (1) small size and suitability for in situ use; (2) it can be used very economically in multiple sensor arrangements; (3) suitability for very low dew point levels without the need for sensor cooling (as required in condensation-type sensors—(typically, dew points down to $-100°C$ can be measured without serious difficulty); (4) the unit covers a wide span.

Limitations of the aluminum oxide sensor include: (1) the sensor is a secondary measurement device and must periodically be calibrated to accommodate aging effects, hysteresis, and contamination; and (2) sensors require separate calibration curves, which are typically nonlinear.

Aluminum oxide humidity instruments are available in a variety of types, ranging from a low-cost, single-point system, including portable battery operated models, to multipoint microprocessor based systems with capability to compute and display humidity information in different parameters, such as dew point, %RH, etc.

The aluminum oxide sensor is also used for moisture measurements in liquids (hydrocarbons). Because of its low power usage, it is suitable for use in explosion proof installations. These sensors are frequently used in petrochemical applications where low dew points are to be monitored on line and where the reduced accuracies and other limitations are acceptable. The advantages of the sensor must be weighted against the fact that accuracy is lower than with any of the fundamental measurement sensor types. As a secondary measurement device, it can provide reliable data only if kept in calibration and if damage due to incompatible contaminants is avoided.

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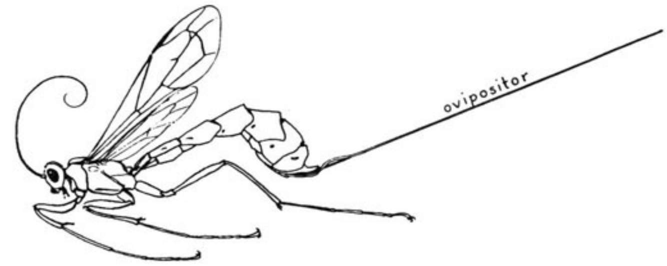
HYGROSCOPIC. 1. Pertaining to a marked ability to accelerate the condensation of water vapor. In meteorology, this term is applied principally to those condensation nuclei composed of salts that yield aqueous solutions of a very low equilibrium vapor pressure compared with that of pure water at the same temperature. Condensation on hygroscopic nuclei may begin at a relative humidity much lower than 100% (about 75% for sodium chloride); while on so-called non-hygroscopic nuclei, which merely furnish sufficiently large (by molecular standards) wettable surfaces, relative humidities of nearly 100% are required.

2. Descriptive of a substance, the physical characteristics of which are appreciably altered by effects of water vapor. The hygroscopicity of certain materials has been advantageously utilized in humidity measurement and control devices; for example, the hair element of a hair hygrometer.

HYMENOPTERA. One of the large orders of insects, including ants, bees, wasps, sawflies, and many species without common names. The mouth is formed for biting or for biting and sucking and the wings, when present, are four in number and membranous. Metamorphosis is complete. The order includes plant-eating, parasitic, and predacious species, and in the ants and bees displays some of the finest examples of social organization. The order includes about 70,000 species.

Owing to its extent and diversity this division of the insects includes many species of economic importance. Some of the sawflies and gall wasps are harmful to plants and, on the other hand, the fig insects are beneficial and the galls produced by some gall wasps are of commercial value. Many parasitic species are of undoubted value in holding in check important insect pests. Ants are sometimes very troublesome and the large carpenter bee sometimes damages wood in construction. The most important single species is the honeybee, which is of great value



The degrees of specialization represented by members of *Hymenoptera* are exemplified by this ichneumon wasp, which incorporates a greatly extended proboscis for placing egges in the bodies of other insects, notably caterpillars. (*USDA photo.*)

as a producer of honey and wax and in the cross pollination of fruit trees.

The main families of Hymenoptera include:

Andrenidae (also called *Halictidae*)	Mining bees, sweat bees
Apidae	Honeybees
Anthophoridae	Anthophorid bees
Bombidae (also called *Bremidae*)	Bumblebees
Braconidae	Braconid wasps
Cephidae	Stem sawflies
Ceratinidae	Small carpenter bees
Chalcididae	Chalcid wasps
Chrysididae	Cuckoo wasps
Colletidae (also called *Hylaeidae*)	Bifid-tongued bees or plaster bees
Cynipidae	Gall wasps
Dryinidae	Dryinid wasps
Eumenidae	Mud or potter wasps
Evaniidae	Ensign wasps
Formicidae	Ants
Ichneumonidae	Ichneumon wasps
Megachilidae	Leaf-cutting bees and mason bees
Multilidae	Velvet ants
Nomadidae	Cuckoo bees
Pompilidae (also called *Psammocharidae*)	Spider asps
Proctotrupidae (also called *Serphiodea*)	Egg-parasite wasps
Prosopidae (also called *Hylaeidae*)	Obtuse-tongued bees or wasplike bees
Scoliidae	Vespoid digger wasps
Siricidae	Horn-tails
Sphecidae	Digger wasps, mud-daubers, thread-waisted wasps
Tenthredinidae	Sawflies
Vespidae	Social wasps, paper-nest wasps, hornets, yellow jackets
Xylocopidae	Large carpenter bees

HYPABYSSAL. A general term sometimes used by structural geologists and petrologists to designate those igneous rocks such as sills and dikes which have congealed under less pressure than the plutonic or deep-seated rocks, but under greater pressure than the effusive rocks (lavas).

HYPERACTIVITY (Children). Professionally, this disorder of some children is termed *Attention Deficit-Hyperactivity Disorder* (ADHD). The disorder may be described in lay terms as an *excessively rambunctious behavior*. ADHD affects, in varying degrees, an estimated 5 million children in the United States. The disorder is most common among boys.

ADHD is poorly understood. At one time, the disorder was considered to be a purely psychological problem. Today, ADHD is considered a physical disorder, and some experts believe it may be inherited. It has been proposed, but not proved, that ADHD children lack certain neurotransmitters (chemical "messengers" that transmit signals within the brain). This insufficiency may have a genetic base. As with most scientifically unknown situations, there is a tendency to suspect nearly anything within reason as a cause. Thus, diet, lead poisoning, food additives, allergies, and other factors have been suspected by not proved to date as a cause of ADHD. One factor that has been well established is that ADHD is not associated with brain damage or impaired intelligence. About half of ADHD children outgrow the disorder.

Some clinics that specialize in their attempts to treat ADHD use a dual approach involving medication and modification of behavior and environment. Medications are selected to control the child's hyperactivity.

Parental control of the ADHD child is important. Some authorities suggest:

1. *Set limits*—Establish a system of rewards and punishment that is consistent with the child's behavior. Consistency of attention and approach to the child is very important.
2. *Encourage a sense of responsibility*—An ADHD child should not be removed from responsibility simply because of its actions or reactions.
3. *Monitor educational needs*—If the child falls behind in such subjects as reading and mathematics, remedial education classes should be encouraged wherever available. Testing can reveal the need for special education. Close cooperation between teacher, parent, and child must be given a very high priority.

Only after careful analysis by one or more physicians should stimulants, on the one hand, or tranquilizers, on the other hand, be considered.

In diagnosing ADHD, one or more of the following characteristics will be determined:

1. What sets ADHD children apart from other feisty and inattentive children with boundless energy is the intensity and persistence of ADHD behavior. The child acts much younger than his/her chronological age.
2. Easy distraction and a very short attention span, sometimes with extreme mood swings. Although there is a pattern of going from one project to another in many children, the ADHD child in most cases will do this persistently. However, in rarer cases, an ADHD child may become deeply absorbed in certain pursuits.
3. Hyperactivity may become evident at an early age, with feeding and sleeping problems and unexplained crying. Drumming fingers, shuffling feet, and the inability to sit still are commonly manifested.
4. The ADHD child is impulsive, acting on the spur of the moment. Although there are exceptions, untidiness and risk-taking behavior are common.
5. Attention-demanding behavior. The ADHD child desires to be center stage. Actions may include virtual non-stop talking, whining, badgering, teasing, and bossing of other children. But, in some cases, the ADHD child may be "cold" emotionally and quite unresponsive to affection or discipline.

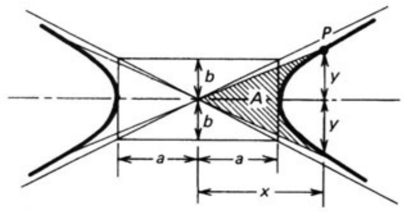
ADHD is a serious but manageable condition when parents, teachers, and friends are aware of the disorder. With proper supervision of a physician knowledgeable and experienced in handling ADHD cases and a supportive home environment, the ADHD child can enjoy the many positive aspects of childhood.

HYPERBOLA. A conic section obtained by a plane cutting both nappes of a right-circular conical surface. It is the locus of a point which moves so that the difference of its distances from two foci is a constant. Its eccentricity is greater than unity.

The standard equation may be taken as $x^2/a^2 - y^2/b^2 = 1$. The curve is a central conic for it is symmetric about both the X- and Y- axes when placed in this standard position and the coordinate origin is its center. The transverse axis, coincident with the X-axis, is of length $2a$; the conjugate axis, along the Y-axis, has length $2b$ ($b < a$). The distance from the center of the hyperbola to either focus is $\sqrt{a^2 + b^2}$ the eccentricity, $e = \sqrt{a^2 + b^2}a$; the length of the latus rectum is $2b^2/a$; the equations for the directrices are $x = \pm a/e$, the same as for the ellipse. The distance from any point on the hyperbola to a focus is a focal radius and the differences between any two focal radii equals $2a$. The lines $y = \pm bx/a$ are asymptotes to the hyperbola. If the length of the transverse axis becomes equal to that of the conjugate axis ($a = b$), the curve is an equilateral or rectangular hyperbola. In this case, the asymptotes are perpendicular to each other. If the coordinate axes are rotated so that they coincide with the asymptotes, the equation for the rectangular hyperbola becomes $xy = a^2/2$, a form which is familiar to students of physical chemistry as Boyle's law.

The polar equation of the hyperbola is $r = a(e^2 - 1)/(e \cos \theta - 1)$ and its parametric equations are $x = a \cosh u$, $y = b \sinh u$ or $x = a \sec \phi$, $y = b \tan \phi$. Its evolute is similar to that of the ellipse $X^{2/3} - Y^{2/3} = 1$, where $X = ax/e^2$, $Y = by/e^2$.

With reference to the accompanying diagram, the shaded area $= ab \log_e(x/a + y/b)$. In an equilateral hyperbola, $a = b$, in which case the



Major parameters of hyperbola.

shaded area = $a^2 \log((x + y)/a) = a^2 \log(a/(x - y)) = a^2 \sin h^{-1} (y/a)$, or $a^2 \cos^{-1}(x/a)$.

See also **Conic Section.**

HYPERBOLIC FUNCTION. Combinations of $e^{\pm z}$ with properties similar to those of the trigonometric functions. They are defined by:

$$\sinh z = (e^z - e^{-z})/2 = z + \frac{z^3}{3!} + \frac{z^5}{5!} + \cdots$$

$$\cosh z = (e^z + e^{-z})/2 = 1 + \frac{z^2}{2!} + \frac{z^4}{4!} + \cdots$$

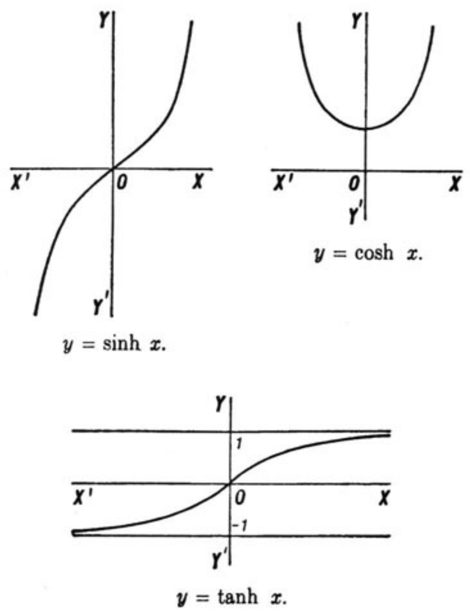
$$\tanh z = \sinh z/\cosh z;\ \coth z = 1/\tanh z$$

$$\text{sech}\, z = 1/\cosh z;\ \text{csch}\, z = 1/\sinh z$$

If n is a positive integer, $i = \sqrt{-1}$, $u = n\pi i$; $\sinh u = \tanh u = 0$; $\cosh u = (-1)^n$; $\sinh(z + u) = (-1)^n \sinh z$;

$$\cosh(z + u) = (-1)^n \cosh u.$$

For real z, the hyperbolic functions are related to the hyperbola in the same way that the trigonometric functions are related to a circle. If $x^2 \pm y^2 = a^2$, with a plus sign, is the equation for a circle of radius a, or, with a minus sign, for a rectangular hyperbola, the parametric equations are $x = a \cos \phi$, $y = a \sin \phi$ and $x = a \cosh z$, $y = a \sinh z$, respectively. The equations $x = a \sec \phi$, $y = a \tan \phi$ also apply to the hyperbola, and ϕ is the same angle as that for the circle. Comparison of these results shows that $\cosh z = \sec \phi$, $\sinh z = \tan \phi$, with $-\pi/2 < \phi < \pi/2$. Further relations are obtained from the equations defining the hyperbolic functions: $\text{sech}\, z = \cos \phi$, $\text{csch}\, z = \cot \phi$, $\tanh z = \sinh \phi$, $\coth z = \csc \phi$. The equation $\sinh z = \tan \phi$ determines ϕ as a function of z. It is called the gudermannian of z, and thus $\phi = \tan^{-1} \sinh z = \text{gd}\, z$.



Major hyperbolic functions.

Again, with real z, hyperbolic and circular (trigonometric) functions are related as follows: $\sinh iz = i \sin z$; $\cosh iz = i \cos z$; $\tanh iz = i \tan z$. Additional formulas, similar to those familiar from trigonometry, are: $\cosh^2 z - \sinh^2 z = 1$; $1 - \tanh^2 z = \text{sech}^2 z$; $\cosh^2 z + \sinh^2 z = \cosh 2z$; $2 \sinh z \cosh z = \sinh 2z$.

The inverse hyperbolic functions are also denoted in a manner similar to that for the inverse trigonometric functions. Thus, if $y = \sinh z$, the inverse function is the angle whose hyperbolic sine is y, or $z = \sinh^{-1} y = \text{arc sinh}\, y$. The following relations may be obtained from the definitions of the various functions:

$$\sinh^{-1} z = \ln(z + \sqrt{z^2 + 1})$$

$$\cosh^{-1} z = \ln(z \pm \sqrt{z^2 + 1});\quad z \geq 1$$

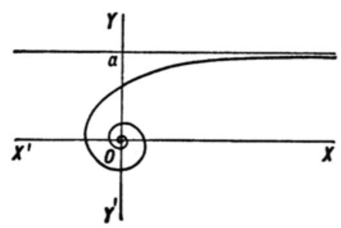
$$\tanh^{-1} z = \frac{1}{2} \ln \frac{1 + z}{1 - z};\quad z^2 < 1$$

$$\coth^{-1} z = \frac{1}{2} \ln \frac{z + 1}{z - 1};\quad z^2 > 1$$

$$\text{sech}^{-1} z = \ln \frac{1 \pm \sqrt{1 - z^2}}{z};\quad 0 < z \leq 1$$

$$\text{csch}^{-1} z = \frac{1 \pm \sqrt{1 + z^2}}{z}$$

HYPERBOLIC SPIRAL. A transcendental plane curve, also known as a reciprocal spiral, with polar equation $r\theta = a$ and thus inverse to Archimedes' spiral. It begins at an infinite point from the pole, but as it winds around it never reaches the pole. It has an asymptote $y = a$. Its equation can also be taken as $xt = a \cos t$, $yt = a \sin t$, where t is a parameter.



Hyperbolic spiral.

HYPERBOLOID. A central quadric surface with one or two negative terms in its equation. If there is only one, so that $x^2/a^2 + y^2/b^2 - z^2/c^2 = 1$, the surface is a hyperboloid on one sheet. It is given this name because any point on the surface may be reached from any other point on the surface. A plane parallel to the XY-plane cuts out an ellipse, but if the sections are parallel to the XZ- or YZ-planes the results are hyperbolas. When $a = b$, the sections by planes z = constant are circles and the surfaces can be generated by revolving the hyperbola, $x^2/a^2 - z^2/c^2 = 1$ about its conjugate axis, the Z-axis.

If there are two negative terms in the equation, $x^2/a^2 - y^2/b^2 - z^2/c^2 = 1$, the surface is a hyperboloid of two sheets, separated into two parts symmetrically located above and below the planes x = constant. Traces parallel to the XY- and XZ-planes are hyperbolas and traces parallel to the YZ-planes are ellipses, provided $x > a$. When $b = c$, the sections by planes x = constant are circles and a surface of revolution results when the hyperbola $x^2/a^2 - y^2/c^2 = 1$ is rotated about its X- or transverse axis.

See also **Quadric Surface.**

HYPERCONJUGATION. The description of the properties of a molecule in terms of resonance structures in which an atom or group is not joined by any sort of bond to the atom to which it is ordinarily considered linked. Also called no-bond resonance. The hypothesis of hyperconjugation has been advanced to interpret some properties of substances containing but 1 double bond by analogy with those of substances containing conjugated double bonds. Consider a substance with a terminal structure $H_3C{-}CH{=}CH{-}$. . . . One of the possible resonating structures of this group is

$$H_3{\equiv}C{-}CH{=}CH{-}{-}H_3{=}C{=}CH{-}CH{-}$$

the dotted line indicating two unpaired electrons with opposite spins.

HYPEREUTECTIC ALLOY. An alloy with a composition falling on the right of the eutectic point of a binary phase diagram that freezes with a structure containing some eutectic.

HYPERFINE STRUCTURE. In general, a set of very closely spaced lines in atomic spectra or other kinds of spectra. There may be many causes of hyperfine structure: (1) for a single atomic species or nuclide, the occurrence of spectral lines as doublets, triplets, etc., due to the interaction, or coupling, of the total angular momentum of the orbital electrons with the nuclear spin and associated magnetic moment; (2) for an element consisting of several isotopes, the occurrence of components for each spectral line that is observable under high resolution, each isotope contributing one or more components. This type of hyperfine structure is often called isotope structure to differentiate it from the first type of hyperfine structure discussed above. See also **Atomic Spectra.**

HYPERGEOMETRIC DISTRIBUTION. A distribution of a discrete random variable generally associated with sampling from a finite population without replacement. The frequency of r "successes" and $n - r$ "failures" in a sample of n so drawn from a population of N in which there are N_p "successes" and N_q "failures" $(p + q = 1)$ is

$$\frac{1}{N^n}\binom{n}{r}(N_p)^{[r]}(N_q)^{[n-r]}$$

where $N^{[r]} = N(N - 1)\ldots(N - r + 1)$. As N tends to infinity the distribution tends to the ordinary binomial form. The distribution derives its name from the fact that the probability generating function may be put in the form of a hypergeometric series.

HYPERONS. These are subatomic particles that are more massive than nucleons (protons and neutrons). *Strangeness* is a property of elementary particles found useful in classifying hyperons. Each particle is assigned a strangeness quantum number S which is related to the electric charge Q, the isospin number T, and the baryon number B by the formula $Q = T + (S + B)/2$. ($T = \frac{1}{2}$ for a proton and $-\frac{1}{2}$ for a neutron; other particles may have $T = 0$ or $T = 1$, depending on the type.) Strangeness is conserved in reactions involving the strong interaction. The selection rules resulting from strangeness conservation are important in understanding why some reactions take place much more slowly than others. See also **Particles (Subatomic).**

HYPEROPIA. See **Vision and the Eye.**

HYPERSONIC FLOW. In aerodynamics, flow of a fluid over a body at speeds much greater than the speed of sound and in which the shock waves start at a finite distance from the surface of the body.

HYPERSTHENE. The mineral hypersthene is an orthorhombic pyroxene, chemically a ferro-magnesian silicate, differing from enstatite in that the iron content is considerable (FeO being greater than 15%). A general formula is $(Mg, Fe)SiO_3$. It is usually found as a massive mineral, whose crystals tend to be prismatic or tabular in habit. It has a distinct prismatic cleavage; fracture, uneven; brittle; hardness, 5–6; specific gravity, 3.42–3.84; luster, pearly to somewhat metallic; color, brownish-green, brown, greenish-black to grayish-black; streak, grayish-brown; translucent to opaque. Hypersthene is often associated with labradorite in gabbro and norite and in extrusive rocks like andesite. It is occasionally encountered in meteorites. Hypersthene is associated with pyrrhotite in Bavaria, with labradorite on the Isle St. Paul, Labrador. It is also found in Montmorency County, Quebec; and in the United States in the rocks of the Cortlandt series in the Hudson River Valley, and the andesites of Colorado and northern California. Superb crystals of exceptional size and quality have been found growing into and within the almandinepyrope garnets at Gore Mountain, North River, New York. The rarity of hypersthene in crystal form makes this occurrence noteworthy. The word hypersthene comes from the Greek words meaning *strong* or *tough*.

See also **Pyroxene.**

HYPERTENSION (High Blood Pressure). Commonly regarded and treated as a disorder in itself, high blood pressure may be more accurately described as a major symptom of a complex of disorders. Not all of these disorders are present in one person. There is a variety of patterns of these underlying disorders which readily explains what was a puzzle for many years—namely, the manner in which different people with the symptom of hypertension react differently to various drug therapies. Considering the numerous body systems—circulatory, nervous, endocrine, excretory, among others—that interact in different ways to produce a universal symptom (hypertension), it would indeed be surprising if all persons with high blood pressure did react precisely in the same manner to therapy.

Considering the heart as a pump, (1) the *systole* is the period of the heart's contraction, or the contraction itself—the systolic pressure represents the highest arterial blood pressure; (2) the *diastole* is the period of the heart's dilation—the diastolic pressure represents the lowest arterial blood pressure that occurs between the pulse waves. The arterial blood pressure in humans is usually measured in the arm at the brachial artery, preferably with the patient seated or lying down with the arm slightly flexed and at heart level. In a thorough examination for hypertension, multiple readings will be made—both arms and legs. With the aid of a stethoscope, the examiner will determine both systolic and diastolic pressure with the *sphygmomanometer*, an instrument whose pressure scale is calibrated in millimeters of mercury. See **Manometer;** and **Sphygmomanometer.**

Statistically, over decades, expected ranges for these pressures in healthy persons have been established. These ranges have become established standards against which individual readings are compared and from which a diagnosis of high blood pressure (*hypertension*) or low blood pressure (*hypotension*) is made. The pressures in the arteries and the veins were first measured as early as 1733 by Stephen Hales, who used a rather crude measurement technique in making determinations of these pressures in a mare. By 1828, Hales had developed a method using a U-tube manometer, the prototype of current instruments.

The statistical averages for blood pressure of healthy adults are:

Systolic	110–120 mm mercury
Diastolic	65–80 mm mercury

Some observers have reported that the systolic pressure is higher in men than in women. The normal upper limits of systolic pressure are

140 mm mercury (men)
130 mm mercury (women)

In a conservative approach to hypertension, treatment is indicated as follows:

Age	*Systolic Pressure*
Under 35	Greater than 140 mm mercury
35–59	Greater than 150 mm mercury
60+	Greater than 160 mm mercury

According to Koch-Wester (1973), *hypertension* exists when the systolic pressure exceeds 150 mm mercury and the diastolic pressure is greater than 90 mm mercury. *Borderline* hypertension has been defined as the intermittent elevation of systolic or diastolic pressure above the accepted normal value for a person's age and sex.

Hypertension has been variously estimated to affect between 20 and 35 million persons in the United States alone and thus it is considered the most common of the chronic disorders. Of the millions of people affected, it is estimated that, as of the early 1980s, only about 50% of

EFFECT OF VARIOUS RISK FACTORS (Including Hypertension) ON OCCURRENCE OF CARDIOVASCULAR DISEASE
(Within eight years in a 45-year-old male)

Risk Factor Effects	Glucose Intolerance	Cholesterol Level[1]	Smoke Cigarettes	Left Ventricular Hypertrophy	Probable Cases per Thousand at a Systolic Blood Pressure of (Millimeters Mercury): 105	135	165	195
ONE RISK FACTOR PRESENT								
None present	No	Low	No	No	22	35	54	84
Glucose intolerance (GI)	Yes	Low	No	No	39	61	95	143
High cholesterol (HC)	No	High	No	No	44	68	105	158
Smoking (SM)	No	Low	Yes	No	38	59	91	138
Left ventricular hypertrophy (LVH)	No	Low	No	Yes	60	93	141	208
COMBINED MULTIPLE RISK FACTORS PRESENT								
GI + HC	Yes	High	No	No	145	214	304	411
GI + HC + SM	Yes	High	Yes	No	229	323	433	550
GI + HC + SM + LVH	Yes	High	Yes	Yes	460	577	686	778

NOTES: The data are based upon a follow-up of patients in the Framingham Study. Framingham males in the study had the following characteristics at the age of 45 years:

Average systolic blood pressure, 131 millimeters mercury; average serum cholesterol level, 234 milligrams/100 milliliters; 0.7% have definite left ventricular hypertrophy as shown by an electrocardiogram; 3.9% have glucose intolerance. Considering these average values, the probability of having cardiovascular disease within 8 years is 75/1000.

[1]Low cholesterol level is considered 185 milligrams/100 milliliters; high cholesterol level, 335 milligrams/100 milliliters.

cases have been diagnosed and are known to the individuals. Of the remaining 50% of cases, only about half are being treated. Hypertension is considered a major health problem because the disorder predisposes individuals to debilitating and often fatal diseases, the major categories of which are heart attack and heart diseases, stroke, and kidney failure. The risk of these consequences is *greatly reduced* with proper therapy for lowering blood pressure.

The foregoing is exemplified by a study made in Framingham, Massachusetts several years ago, which took into account five risk factors: (1) Glucose intolerance (indicative of diabetes), (2) cholesterol level, (3) cigarette smoking, (4) left ventricular hypertrophy (increase in volume of a tissue or organ produced entirely by enlargement of existing cells), and (5) hypertension. See accompanying table.

Primary and Secondary Hypertension. Traditionally, authorities in the field have made a distinction between two forms of hypertension. *Primary* or *essential* hypertension is a disorder of unknown etiology. This form accounts for approximately 90% of all cases of hypertension. In *secondary* hypertension, a cause for the disorder can be identified. The causes of secondary hypertension are many and include: various drugs, such as amphetamines, oral contraceptives, estrogens, steroids, and thyroid hormones; increased intracranial pressure; certain tumors, such as pheochromocytoma; primary aldosteronism (abnormal aldosterone secretion by adrenal cortex, causing excessive loads of potassium and muscular weakness); several renal diseases, such as chronic pyelonephritis, diabetic nephropathy, glomerulonephritis, gout, polycysystic disease, vasculitis, and renovascular hypertension, among others. Hypertension is also associated with toxemia of pregnancy, acute pulmonary edema, acute myocardial infarction, dissecting aortic aneurysm, and cerebral hemorrhage.

Primary (Essential) Hypertension

Although considerable progress has been made during the past few years in understanding the complex root causes which contribute to primary hypertension, much of this information has stemmed from observing the actions of various drugs used in the therapy of hypertension.

The Renin-Angiotensin-Aldosterone System. In the early 1900s, Tigerstedt and Bergman suggested that *renin*,[1] a proteolytic enzyme elicited by ischemia of the kidneys or by diminished pulse pressure, played an important role in blood pressure homeostasis and in the pathogenesis of hypertension. In a revival of interest in the role of renin, which commenced during the 1960s, considerable new information has been gained, with an increasing implication of the kidneys in hypertension. Additional reninlike enzymes have been identified and their possible functions are being researched. In a rather complex pathway, renin cleaves renin substrate to yield *angiotensin I*, a decapeptide and apparently quite inactive physiologically. Through further enzymatic action during its passage through the pulmonary circulation, angiotensin I is cleaved to produce *angiotensin II*. This substance is the most potent vasoconstrictor known, causing constriction of the arterioles (small arteries that branch to form the capillaries—the smallest of blood vessels at the sites where the blood exchanges nutrients and waste products with the tissues).

Angiotensin also stimulates adrenocortical production of aldosterone, which promotes the reabsorption of sodium and water by the renal tubules. The resulting augmentation of the fluid content of the circulatory system elevates the blood pressure. The action of the angiotensins also involves the nervous system, releasing the neurotransmitter norepinephrine by the nerve terminals of the sympathetic nervous system and by the adrenal medulla (inner portion of adrenal gland). This action potentiates the action of the norepinephrine, which causes increased blood pressure as the result of constriction of the arterioles. This complex is known as the *renin-angiotensin-aldosterone system*. When this system is functioning normally, a decrease of pressure of the blood flowing through the kidneys will stimulate renin release, while an increase of that pressure "signals" the kidney to halt the release of renin as well as angiotensin, an action accomplished by a feedback mechanism. Thus, any increase in blood pressure will be transient. But, when the feedback apparatus dysfunctions, chronic hypertension may result.

Statistics show that about 90% of persons with primary hypertension have elevated levels of renin in the blood, although they usually do not exhibit symptoms of kidney damage. One researcher has found that persons with high plasma renin activity run the highest risk of heart attack, stroke, and kidney failure. The reverse situation holds for those persons with low plasma renin activity.

Some researchers have learned that angiotensin II must combine with specific receptors on target organs before its effects can be produced. Most peptides are angiotensin antagonists and thus, by binding with the

[1]Not to be confused with rennin, an enzyme secreted by the glands of the stomach which causes curdling of milk.

receptors, prevent angiotensin from binding. These antagonist substances have been synthesized by several investigators (Cleveland Clinic Foundation; Washington University Medical School). One of these antagonists is the octapeptide called Saralasin® (first synthesized by Norwich Pharmacal Company). It has been found that intravenous injection of this drug will lower the blood pressure to near-normal levels if the cause of hypertension is renin. This kind of information has proved helpful in determining the most effective drug therapy for primary hypertension.

Other investigators have found that the destruction of a portion of the mid-brain (*subnucleus medialis*) will negate the blood pressure response, probably because the area is involved in the control of peripheral resistance to blood flow. Apparently angiotensin II has the ability to cross the blood-brain barrier. More recently, some researchers have suggested that angiotensin is synthesized in the brain. Much of the experimentation to date has been carried on with laboratory dogs.

Involvement of the nervous system in hypertension has attracted much interest in recent years and ultimately may prove or disprove the association of stress with hypertension. Other investigators have found that prostaglandin E_2 tends to decrease blood pressure by countering the angiotensin-induced constriction of the blood vessels.

Additional Reading

Bonna, K. H., et al.: "Effect of Eicosapentaenoic and Docosahexaenoic Acids on Blood Pressure in Hypertension: A Population-Based Intervention Trial from the Tromose Study," *N. Eng. J. Med.*, 795 (March 22, 1990).

Calhoun, D. A., and S. Oparil: "Current Concepts—Treatment of Hypertensive Crisis," *N. Eng. J. Med.*, 1177 (October 25, 1990).

Fackelmann, K. A.: "High-Pressure Hormone," *Science News*, 344 (December 1, 1990).

Grimm, R. H., Jr., et al.: "The Influence of Oral Potassium Chloride on Blood Pressure in Hypertensive Men on a Low-Sodium Diet," *N. Eng. J. Med.*, 569 (March 1, 1990).

Grobbee, D. E., et al.: "Coffee, Caffeine, and Cardiovascular Disease in Men," *New Eng. J. Med.*, 1026 (October 11, 1990).

Guyton, A. C.: "Blood Pressure Control—Special Role of the Kidneys and Body Fluids," *Science*, 1813 (June 28, 1991).

Kaplan, N. M., Brenner, B. M., and J. H. Laragh, Editors: "New Therapeutic Strategies in Hypertension," Raven Press, New York, 1989.

Kem, D. C., and R. D. Brown: "Renin—From Beginning to End," *N. Eng. J. Med.*, 1136 (October 18, 1990).

Kilgour, F. G.: "William Harvey," in *Scientific Genius and Creativity*, W. H. Freeman, New York, 1987.

Panza, J. A., et al.: "Abnormal Endothelium-Dependent Vascular Relaxation in Patients with Essential Hypertension," *New Eng. J. Med.*, 22 (July 5, 1990).

Salisbury, D.: "Hypertension: A Discriminating Disease," *Technology Review (MIT)*, 22 (October 1990).

Sytkowski, P. A., Kannel, W. B., and R. B. D'Agostino: "Changes in Risk Factors and the Decline in Mortality from Cardiovascular Disease—The Framingham Heart Study," *N. Eng. J. Med.*, 1635 (June 7, 1990).

Vane, J. R., Anggard, E. E., and R. M. Botting: "Regulatory Functions of the Vascular Endothelium," *N. Eng. J. Med.*, 27 (July 5, 1990).

HYPOBARIC (Controlled-Atmosphere) SYSTEMS. Sensitive materials, notably fresh foods, normally cannot withstand long periods of transportation and storage prior to consumption. Over the years, much of the effort extended toward offering produce in marketplaces far distant from the source was concentrated on reducing the time for delivery. Thus, the extensive use of air express and air freight. Conventional refrigeration systems for trucks and railway cars also were especially adapted for use during transport. But even with all of these improvements in technology, certain transporting feats (as, for example, shipping midwestern pork to the California and even Hawaii markets) were difficult to achieve. The controlled-atmosphere hypobaric concept has greatly extended the potential for distant shipping of delicate, perishable materials (not necessarily limited to foodstuffs).

In 1964, the Institute of Food Technologists annual award was given in recognition of the development of a controlled-atmosphere storage system. Essentially, the process was designed to reduce the rate of deterioration of certain fruits and vegetables in refrigerated storage by reducing the oxygen level, increasing the carbon dioxide level, and maintaining the relative humidity close to 100%. In an initial design, the conditions were created by using a home-furnace size catalytic generator which burned natural gas or propane gas to create the atmosphere that essentially halts the natural respiration of the stored products. Later, the gas generator was replaced by cryogenic liquefied gases, allowing additional flexibility and the creation of any desired gas mixture. Atmospheres can be tailored to particular perishables. For example, an atmosphere of 15–20% carbon dioxide and 80–85% nitrogen is optimal for strawberries. For iceberg lettuce, an atmosphere of 8–10% oxygen, less than 10% carbon dioxide, with the remainder nitrogen is used. As of the early 1980s, the system has been installed on over ten thousand rail cars and 7000 sea vans.

In 1979, another IFT award was given in recognition of a hypobaric transport and storage system for fresh meats and meat products. Hypobarics is defined as a precisely controlled combination of low pressure, low temperature, high humidity, and ventilation which, when properly applied, extends up to six times the length of time a perishable commodity remains fresh. This makes possible the shipment of perishable items by way of relatively low-cost surface transportation to distant points. In developing the concept, it was observed that refrigerated storage of fruits in closed containers will result in accumulation of gases generated by the fruit, i.e., ethylene and carbon dioxide, an atmosphere which hastens ripening and spoilage. Although ventilation of fruit containers can prevent accumulation of the gases, the gases are not removed from within the product itself—with no prevention of accumulation of gases within the cells of the fruit. The researchers made the supposition that by drawing a partial vacuum on a closed vessel containing the fruit, the low pressure would increase the diffusivity of the gases, thus promoting release and removal of the gases. At the same time, a reduction of pressure would reduce the oxygen concentration, thus retarding respiration and attendant spoilage. Combined with refrigeration, this would decelerate the metabolic processes, not only of the fruit, but also of any bacteria present. Humidification of the chamber would prevent any drying of the fruit. After testing the concept on bananas and other perishables, the system was patented.

Generally the storage temperature for meats is about -1°C, and up to 10 or 12°C for various fruits and vegetables. In all cases, the relative humidity is controlled at about 95%. Pressure ranges between 10 and 80 millimeters of mercury. Lower pressures are maintained for meats and seafoods; somewhat higher pressures for fruits and vegetables.

HYPOCHLORITES. When chlorine is reacted with an alkali, a hypochlorite is formed. These compounds are very high-tonnage chemicals for sanitizing and bleaching purposes. Commercial sodium hypochlorite $NaClO$ usually is available in two strengths (1) the familiar household liquid bleach which contains about 5.25% (weight) $NaClO$, and (2) commercial bleach which contains about 13% (weight) $NaClO$. The latter compound sometimes is referred to as 15% bleach because the chlorine content is approximately 150 grams/liter of available chlorine. The term "liquid chlorine" usually refers to a solution of $NaClO$ (up to 10%) used in the swimming-pool trade. "Dry chlorine" is part of the registered trade-mark of a proprietary calcium hypochlorite product containing 70% available chlorine. See also **Bleaching Agents.**

Sodium hypochlorite normally is manufactured in batches by diluting caustic soda to the proper starting concentration. This is approximately 6.8% NaOH for the 5.25% bleach; and about 18.5% NaOH for the 15% bleach. After cooling the caustic soda solution, chlorine gas is added through a sparger pipe until the desired concentration is reached. This usually is determined by making a series of titration analyses. Bleaching powder $CaOCl_2$ is made by passing chlorine gas over slaked lime. This was the first type of chlorine bleaching agent made and dates back to 1799. The product usually contains about 30% available chlorine. Over the years, it was used extensively in the bleaching of textiles and for sanitizing even though the compound is unstable and difficult to use. The original bleaching powder largely has been replaced by an improved calcium hypochlorite product which contains about 70% available chlorine. The compound essentially is a calcium hypochlorite dihydrate and, in one process, is made by chlorinating a slurry of lime and caustic soda. The crystals which precipitate out are mixed with calcium chloride and chlorinated lime. When warmed, the calcium hypochlorite dihydrate precipitates, with sodium chloride remaining in solution. After filtering, the cake is dried, granulated, sized, and packaged. In addition to use in swimming pools, products of this type are used widely for water purification, algae control, and sanitation. On a very high-tonnage basis, calcium hypochlorite $Ca(ClO)_2 \cdot 4H_2O$ is used for pulp bleaching in the paper industry. Bleach liquor containing from 20–40% available chlorine may be produced in batches or continuously. In a continuous system, the flow of chlorine is controlled by making frequent (or continuous) measurements of oxidation-reduction potential.

A common means of detecting hypochlorites is the production of a blue color (caused by free iodine) with starch iodide paper by hypochlorites in weakly alkaline solution. Silver nitrate also precipitates part of the hypochlorite in solutions as white silver chloride.

Hypochlorous Acid. This compound, HOCl, is prepared by the reaction of (1) chlorine monoxide Cl_2O with H_2O, (2) sodium hypochlorite and an acid, excess acid yielding chlorine and oxygen, and (3) chlorine with mercuric oxide suspended in water, mercuric chloride being formed simultaneously. Hypochlorous acid is a yellow solution of characteristic odor. It decomposes upon standing, the rate depending upon (1) concentration, (2) exposure to light, (3) presence of a catalyst (cobaltous hydroxide, for example, promotes the evolution of oxygen), and (4) acidity or alkalinity. Hypochlorous acid is a powerful oxidizing agent and sometimes used as a bleaching agent for organic colors.

Perchloric Acid. This compound, $HClO_4$, is a colorless, fuming, oily liquid, miscible with H_2O, volatile under diminished pressure. A maximum constant-boiling solution (203°C, 760 millimeters Hg) results when the concentration of $HClO_4$ reaches 73% in H_2O. Cold dilute perchloric acid reacts with such metals as zinc and iron, yielding hydrogen gas and the corresponding perchlorate in solution; is stable from the point of view of oxidation and reduction (except that iodine is oxidized to periodic acid, with liberation of chlorine, ferrous salt solutions to ferric, titanous salt solutions to titanic). Concentrated hot perchloric acid, on the other hand, is a powerful oxidizing agent, exploding violently in contact with charcoal, paper, alcohol; causes serious wounds in contact with the skin. Prepared by distilling ammonium perchlorate with HNO_3 and HCl.

Metallic perchlorates are soluble in water, except that potassium perchlorate is slightly soluble. Potassium perchlorate is, however, insoluble in alcohol containing perchloric acid, a property made use of in the qualitative recognition and quantitative estimation of potassium in salt solutions. Perchlorates, when heated, evolve oxygen and leave the chloride as a residue. Potassium perchlorate decomposes at 400°C.

HYPOCYCLOID. A special case of a cyclic curve, thus a higher plane curve and, in particular, the case where a circle of radius r rolls around inside a fixed circle of radius R. Its parametric equations are

$$x = (R - r)\cos\phi + r\cos\frac{(R - r)\phi}{r}$$
$$y = (R - r)\sin\phi - r\sin\frac{(R - r)\phi}{r}$$

Reference to the corresponding equations for the epicycloid, where the circle rolls around the outside of the fixed circle, will show that the hypocycloid (R, r) is identical with the hypocycloid $(R, R - r)$ or the epicycloid $(R, r - R)$.

Considerations similar to those used for the epicycloid will also show that the curve may or may not repeat itself and that it will produce cusps when its generating point touches the fixed circle. The special case is that in which $R = 4r$ has four cusps, and is called the asteroid.

See also **Asteroid (Mathematics);** and **Curve (Higher Plane).**

HYPODERMIS. The cellular layer of the integument (integumentary system) in the invertebrates, which secretes the outer cuticula.

HYPOEUTECTIC ALLOY. An alloy to the left of the eutectic point in a binary phase diagram that freezes with a structure containing some eutectic.

HYPOFLUORITE. Any compound containing the group—OF. The simple anion FO^- is unknown. A number of covalent hypofluorites are known, including such compounds with carbon, oxygen, nitrogen, sulfur, chlorine and arsenic (uncertain), CF_3OF, CF_3COOF, C_2F_5COOF, NO_2OF, OF_2, O_2F_2, O_3F_2, SF_5OF, FSO_2OF, ClO_3OF and possibly AsF_4OF. These are all powerful fluorinating agents. They react violently with water yielding OF_2 as one product. The oxygen fluorides O_3F_2 and O_2F_2 decompose about −158°C and −100°C, respectively, the former into the latter and the latter into the elements. Nitryl and perchloryl hypofluorites (fluorine nitrate and fluorine perchlorate) easily detonate. The perfluoracyl hypofluorites are much more stable but may also decompose violently. The others appear to be stable.

HYPOGENE. Originated by the geologist Charles Lyell for all igneous rocks which assumed their form, fabric and texture at great depths beneath the surface of the lithosphere.

HYPOIODOUS ACID AND HYPOIODITES. Hypoiodous acid (HOI) is a greenish-yellow solution, of characteristic odor. It is unstable, and cannot be distilled unchanged.

Prepared by reaction (1) of iodine and mercuric oxide (see **Mercury**) suspension in water, mercuric iodide being simultaneously formed, (2) of sodium hypoiodite and an acid, excess acid yielding iodine.

Sodium hydroxide solution reacts with iodine to form iodide and hypoiodite, the latter decomposing in a few hours at ordinary temperatures to form iodide and iodate.

HYPONITROUS ACID AND HYPONITRITES. Hyponitrous acid $H_2N_2O_2$ is a white solid, explosive even at as low a temperature as 0°C, soluble in water, more soluble in ether, can thus be extracted from water solution by ether and the latter evaporated, water solution decomposes quickly into nitrous oxide plus water. Hyponitrous acid is nonreactive with hydriodic acid (a strong reducing agent), but reactive with permanganic acid (a strong oxidizing agent) to form nitrous or nitric acid.

Prepared (1) by reaction of silver hyponitrite $Ag_2N_2O_2$ and hydrogen chloride in anhydrous ether, an evaporation of the resulting solution, (2) by reaction of hydroxylamine H_2NOH plus nitrous acid HONO.

Sodium hyponitrite $Na_2N_2O_2$ is formed (1) by reaction of sodium nitrate or nitrite solution with sodium amalgam (sodium dissolved in mercury), after which acetic acid is added to neutralize the alkali. Sodium stannite ferrous hydroxide, or electrolytic reduction with mercury cathode may also be utilized, (2) by reaction of hydroxylamine sulfonic acid and sodium hydroxide. Silver hyponitrite is formed by reaction of silver nitrate solution and sodium hyponitrite.

HYPOPHOSPHORIC ACID AND HYPOPHOSPHATES. Hypophosphoric acid (H_2PO_3 or $H_4P_2O_6$) is a solid, melting point 55°C, decomposing in solution to form phosphorous plus phosphoric acids. Hypophosphoric acid is used in solution and is a reducing agent, but only with strong oxidizing agents, such as potassium permanganate; and the acid is unaffected by zinc and dilute sulfuric acid (distinction from phosphorous acid). Dehydration of hypophosphoric acid does not yield phosphorus tetroxide; hydration of phosphorus tetroxide does not yield hypophosphoric acid but phosphorous plus phosphoric acids.

Hypophosphoric acid is formed by reaction (1) of yellow phosphorous and potassium permanganate in sodium hydroxide medium, (2) of red phosphorus and calcium hypochlorite solution, (3) also one of the products of slow oxidation at ordinary temperatures of phosphorus in moist air.

There are recorded the following sodium hypophosphates: Na_2PO_3 (or $Na_4P_2O_6$), $NaHPO_3$ (or $Na_2H_2P_2O_6$), $Na_3H(PO_3)_2$ (or $Na_3HP_2O_6$), and $(NaH_3PO_3)_2$ (or $NaH_3P_2O_6$). There is evidence in support of each of the formulas H_2PO_3, $H_4P_2O_6$ for hypophosphoric acid.

Ester: Dimethyl hypophosphate $(CH_3)_2PO_3$ or $(CH_3O)_2PO$. See also **Phosphorus.**

HYPOPHOSPHOROUS ACID AND HYPOPHOSPHITES. Hypophosphorous acid (H_3PO_2, or $H{\cdot}PO_2H_2$) is a colorless liquid, melting point 26.5°C, density 1.493.

Hypophosphorous acid is miscible with water in all proportions and a commercial strength is 30% H_3PO_2. Hypophosphites are used in medicine.

Hypophosphorous acid is a powerful reducing agent, e.g., with copper sulfate forms cuprous hydride Cu_2H_2, brown precipitate, which evolves hydrogen gas and leaves copper on warming; with silver nitrate yields finely divided silver; with sulfurous acid yields sulfur and some hydrogen sulfide; with sulfuric acid yields sulfurous acid, which reacts as above; forms manganous immediately with permanganate.

Hypophosphorous acid is formed by reaction of barium hypophosphite and sulfuric acid, and filtering off barium sulfate. By evaporation of the solution in vacuum at 80°C, and then cooling to 0°C, hypophosphorous acid crystallizes.

Sodium hypophosphite $NaPO_2H_2$, the only sodium hypophosphite, is formed (1) by reaction of yellow phosphorus and sodium hydroxide solution (phosphine simultaneously formed), (2) by reaction of hypophosphorous acid and sodium hydroxide, and evaporating. Sodium hypophosphite, upon heating, yields sodium phosphate and sodium phosphide. Common tests for the hypophosphites are as follows:

1) Zinc reduces dilute sulfuric acid solution of hypophosphites to phosphine recognizable by odor (difference from phosphates).

2) Barium chloride produces no precipitate (difference from phosphites). See also **Phosphorus.**

HYPOPLASIA. Defective or insufficient development of any tissue. Thymic hypoplasia, also known as DiGeorge's syndrome, results from embryopathy of third and fourth pharyngeal pouch area. There are deficiencies of cell-mediated immunity (CMI) and impaired antibodies. Attendant features of the condition are hypoparathyroidism, abnormal feces, and cardiovascular abnormalities. See also **Immunology and Immunization.**

HYPOPROTHROMBINEMIA. Lack of adequate amounts of prothrombin in the blood resulting in tendency to hemorrhage from impairment of the clotting mechanism.

HYPOSULFUROUS ACID AND HYPOSULFITES. Hyposulfurous acid $H_2S_2O_4$ is a yellow solution rapidly oxidized in air to sulfurous acid and then to sulfuric acid. Commercially known as hydrosulfurous acid and its salts as hydrosulfites (but not to be confused with "hypo" which is sodium thiosulfate).

Hyposulfurous acid is a powerful reducing agent, e.g., with copper sulfate forms cuprous hydride Cu_2H_2, brown precipitate, which evolves hydrogen gas and leaves copper on warning, with silver nitrate yields finely divided silver, with permanganate yields manganous compounds. Hyposulfurous acid is formed by reaction of sodium hyposulfite and an acid.

Sodium hyposulfite, sodium hydrosulfite $Na_2S_2O_4 \cdot 2H_2O$ is formed (1) by reaction of zinc and sulfurous acid (or sodium hydrogen sulfite), yielding zinc hyposulfite and then converted by sodium chloride into sodium hyposulfite, (2) by electrolysis of sodium hydrogen sulfite and then addition of sodium chloride.

Sodium hyposulfite is used to bleach sugar, indigo, wood pulp. With moist hydrogen sulfide, sulfur is precipitated and sodium thiosulfate simultaneously formed.

HYPOTENSION. When the systolic arterial pressure is consistently below 100 millimeters of mercury, low blood pressure (hypotension) is said to exist. Many healthy individuals have a blood pressure that is somewhat below average. A moderately low value is usually considered conducive to longer life. When no cause for the low pressure can be found, the condition is referred to as *essential hypotension*. There often are no significant symptoms.

In *orthostatic* or *postural hypotension*, the regulatory mechanism does not function properly so that a person with this condition may suffer unconsciousness simply in changing from a reclining or sitting position to a standing position—as the result of the action causing an abnormal drop in blood pressure. Some normal individuals may from time to time experience a slight giddiness when standing up quickly, but the severe changes in postural hypotension are such that they should be called to the attention of a physician.

Frequently, unrelated diseases, largely degenerative in nature, may cause hypotension as a secondary symptom. Such conditions include acute fevers, Addison's disease, heart failure, hypothyroidism, malnutrition, hyperinsulinism, and anemia. Sometimes associated with transient hypotension are internal hemorrhage, shock, fainting, and anesthesia. In most situations of this type, the blood pressure returns to normal upon removal of the original causative condition. In most instances, hypotension is of major significance only when the blood pressure falls below that required to produce adequate filtration through the kidneys.

See also **Heart and Circulatory System (Human); Hypertension (High Blood Pressure);** and **Shock Syndrome.**

HYPOTHERMIA AND COLD-RELATED INJURIES. The human body contains water in and around its cells. When water inside the cells freezes, the cells burst and die. When water outside the cells freezes, water is drawn out of the cells, causing damage. Lower temperatures also make the blood thicker (viscous) and the blood vessels narrower, resulting in poor circulation and tissue damage.

Several factors affect risks for cold weather injuries. Most of these are self-evident, but still can be overlooked. They include: (1) inappropriate clothing, (2) inactivity, (3) age of the person exposed, (4) lack of customization to cold weather, (5) prior cold injury, (6) cardiovascular disease, (7) use of alcohol, (8) a concurrent accident which may temporarily superexclude attention to a less serious condition, (9) victim is in a state of shock, (10) high altitude, (11) lack of sleep, (12) poor nutrition, (12) lack of fitness, and (13) dehydration. Different levels of injury occur, depending on extremes of weather, the time exposed, and how well the body is protected.

Chillblains. The mildest form of cold-related injury. This occurs with repeated exposure of bare skin to weather ranging from 32° to 60°F (−1° to 15.6°C). The skin swells, turns red, and itches.

Frostbite. This a serious cold injury. Severity depends on temperature, wind, and duration of exposure. Superficial frostbite involves only the skin. A waxy appearance is common, and blisters appear in 1 to 3 days, followed by generalized swelling of the affected area. Deep frostbite damages not only affect the skin, but also deeper tissues and even bone. The nose, ears, fingers, toes, penis, buttocks, and chin are the parts of the body most commonly frostbitten. Thorough rewarming can limit the extent. Severe cold injuries resemble burns and usually are treated in a similar manner.

Hypothermia. Considered a major emergency, the victim may be found semi-conscious or unconscious, often some distance from shelter. Cardiac arrest may have occurred, in which case cardiopulmonary resuscitation measures should be applied for a very long time. This occurs when the body core temperature is below 85°F (29°C). Extreme measures must be taken to improve the airway and ventilation. If transfer to hospital cannot be made immediately, the victim's hands and forearms should be immersed in water maintained at about 113° to 118°F (45° to 48°C) and controlled by a thermometer if possible. As a guide, the water should feel uncomfortable but bearably hot to the rescuer's elbow. A conscious victim should be given hot drinks. At hospital, some authorities prefer whole-body immersion in hot water at 113° to 118°F (45° to 48°C). Other authorities indicate that nothing will succeed if the victim's rectal temperature continues to fall after rescue and that whole-body immersion may be counterproductive. Core temperature can be raised by gastric lavage with warm water containing dextrose. For more detail in treatment of hypothermia, reference to the *Merck Manual* (frequently updated) is suggested.

Trench Foot or Immersion Foot. This injury occurs after prolonged exposure to wet, cold weather, usually ranging from 32° to 50°F (0°C to 10°C). See separate article on **Trench Foot.**

HYPOTHESIS. A tentative assumption, usually based upon some reasonable concept, made in order to generate interest in obtaining proof and to consider the consequences of the assumption.

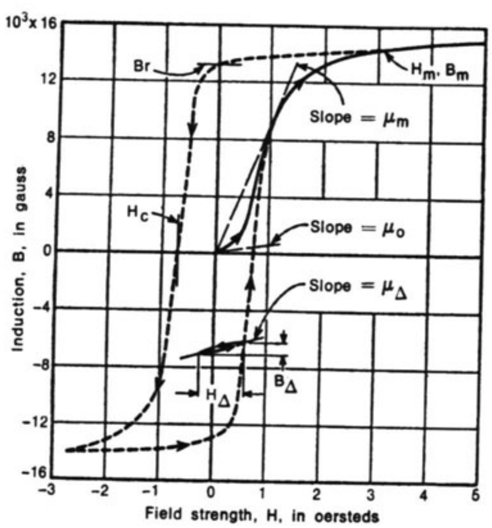
HYRAXES (*Hyracoidea*). A very small group of *Mammalia*, hyraxes are small animals and are of two genera: Dassies (*Procavia*) and Tree-Hyraxes (*Dendrohyrax*). These rabbit-shaped animals are popularly termed Coneys, a term used in the Bible. Classification of these animals has been a problem for zoologists over the years, finally solved by creating a separate small group. At one time, they were considered to be

Adult cape hyrax with young. (*New York Zoological Society.*)

rodents closely associated with guinea-pigs. At another time, they were classified with the Pachyderms, at a time when elephants, rhinoceroses, and hippopotamuses were all grouped together. In the mean time, all of the aforementioned mammals have been reclassified, Pachyderm being an obsolete term.

Hyraxes are nocturnal in nature and are considered highly aggressive and essentially mean, attempting to bite anything that gets close to them. These animals also are quite noisy and can issue a number of different sounds, including whistles and screams. The fur contrasts in color on their mid-backs. The fur is thick and coarse. They are good jumpers. Vacuum cups in their padded feet enable them to cling to vertical surfaces. They have daggerlike teeth. The Dassies are found in Africa south of the Sahara Desert. The Coney mentioned in the Bible is found in the Sinai Peninsula, Palestine, and Syria. They are rock dwellers and live in fur-lined nests. The Tree-Hyraxes prefer a mountain habitat and frequently are found at relatively high altitudes—7,000 to 10,000 feet (2100 to 3000 meters). They are omnivorous. They prefer closed-canopy forests in the central and west-central regions of Africa. See accompanying photo.

HYSTERESIS. In general, the phenomenon exhibited by a system whose state depends on its previous history. This term usually refers to magnetic hysteresis, of importance in alternating-current machinery. When a ferromagnetic material such as iron is placed in a magnetic field, a certain amount of energy is involved in bringing about its magnetization. If the field is a rapidly alternating one, the material may become noticeably warm. It appears that the repeated changes of orientation in whatever it is within the substance that responds to the reversals of field are opposed by something like viscous friction.



Hysteresis loop (dotted). Some important magnetic quantities are shown

A quantitative study of the process indicates that, as the field intensity H increases, the magnetic induction B also increases in a manner characteristic of the substance. This is conveniently represented by a graph, which is called the magnetization curve (see figure). Its initial slope is the initial permeability (μ_0). If H is carried to some maximum value H_m and then reduced (to $-H_m$), B follows the dotted hysteresis curve. B does not fall off as it was built up (solid line); the residual induction B_r is the induction remaining when H has been reduced to zero; the reverse H needed to reduce B to zero is called the coercive force (H_c). From this point the cycle proceeds to describe the closed curve shown by the dotted lines, which is called the hysteresis loop. The initial portion (solid line) is not retraced. The amount of energy converted into heat is proportional to the area of the cycle.

Electric hysteresis is a somewhat analogous phenomenon exhibited by dielectrics in the electric field and gives rise to heating in capacitors.

Some solids exhibit what is called elastic hysteresis, in which the variables corresponding to H and B in the magnetic case are the stress and the strain or deformation. Elastic bodies such as metals operating at stresses below the proportional limit also undergo hysteresis.

Hysteresis energy is that energy used per cycle of operation to overcome the effect of hysteresis.

HYSTERECTOMY. Total or partial removal of the uterus.

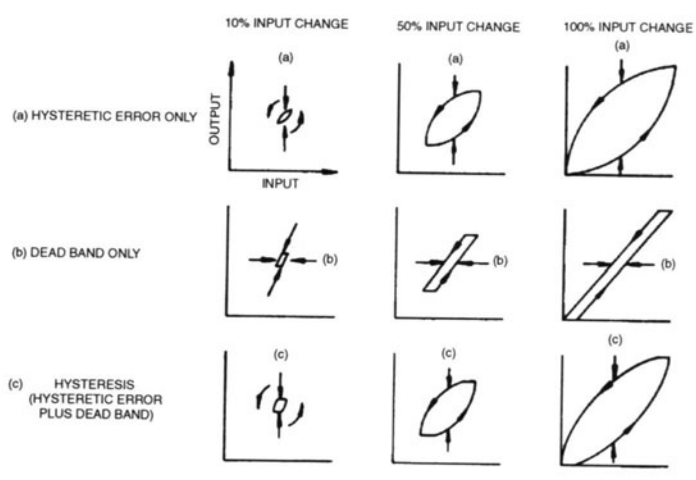
HYSTERESIS DISTORTION. The distortion of voltage and/or current waveforms in circuits containing magnetic components, which is caused by the non-linear hysteresis effect.

HYSTERESIS HEATER. An induction device in which a charge or a muffle about the charge is heated principally by hysteresis losses due to a magnetic flux which is produced in it. A distinction should be made between hysteresis heating and the enhanced induction heating in a magnetic charge.

HYSTERESIS (Instrument). With reference to industrial and scientific instruments, the Scientific Apparatus Makers Association defines hysteresis as:

1. When used as a performance specification, the maximum difference for the same input between the upscale and downscale output values during a full range traverse in each direction. See (c) of accompanying diagram. This is a common usage definition which includes hysteretic error and dead band. That portion of the difference which is dependent on the history of prior excursion is hysteretic error, while that portion due to dead band may be determined by a conventional dead band test.

2. When describing a physical property, that property of an element evidenced by the dependence of the value of the output, for a given



Hysteretic error, dead band, and hysteresis.

excursion of the input, upon the history of prior excursions and the direction of the current traverse. Some reversal of the output will occur on any small reversal of the input if a device exhibits hysteretic error without dead band.

Hysteretic Error. That portion of hysteresis due to energy absorption in the elements of a measuring instrument. It is obtained by subtracting the value of dead band from the corresponding value of hysteresis for a given input. See (a) of accompanying diagram. The energy absorbed is conceived as produced by molecular friction and appears as heat in dynamic cycling when cyclic mechanical force is applied to a spring or cyclic magnetizing force to a magnetic material.

See also **Backlash;** and **Core Loss.**