L

LABARIA (*Reptilia, Sauria*). A poisonous South American snake belonging to the pit vipers. It ranges from eastern Brazil north into the Guianas. Related to the jararaca.

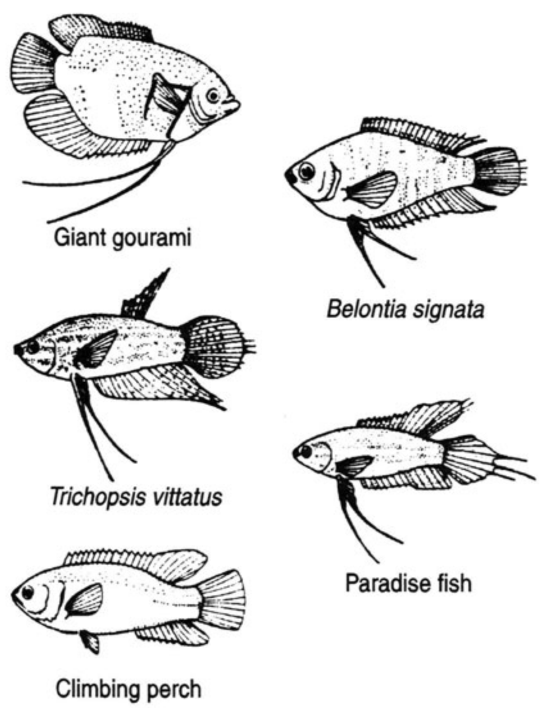
LABRADOR CURRENT. An ocean current that flows southward from Baffin Bay, through the Davis Strait, thence southeastward past Labrador and Newfoundland. East of the Grand Banks, the Labrador current meets the Gulf Stream, and the two flow east separated by the cold wall.

LABYRINTH FISHES (*Osteichthyes*). Of the suborder *Anabantoidea*, and family *Anabantidae*, this is a group of tropical freshwater fishes, quite small in size, and usually found in Africa and southeast Asia. The pelvic fins bear a long slender filament, apparently sensory. They are named because of their highly-specialized breathing apparatus. In using this labyrinth-type anatomical mechanism, the fish draws in a bubble of air from which it extracts the oxygen, and when the fish next surfaces, it expels old air out of the gill covers. This is an accessory apparatus which enables these fishes to tolerate water deficient in oxygen. One member of this family is the walking fish (sometimes called climbing perch) (*Anabas testudineus*) which attains a length of about 10 inches (25 centimeters). It is found in Malaya, the Philippines, and India. Its name stems from the fact that it apparently can walk for considerable distances on land, possibly seeking another body of water. In walking, the fish uses its gill plates, which fortunately are equipped with spiny edges, serving as "feet." The fish can cover about 10 feet (3 meters) in a minute by this method of locomotion.

Another labyrinth fish is the *Betta splendens* (Siamese fighting fish), found in Thailand. The males of this species are renowned for their ability to fight other males of their species. They have a rather dull color and achieve a length of about 2 inches (5 centimeters). The so-called "gourami" (*Osphronemus goramy*) is the largest of the labyrinths, attaining a length of 2 feet (0.6 meter). The fish is used as food in the Orient. Because of its unusual habits, the kissing gourami (*Helostoma temmincki*) is popular among tropical-fish hobbyists. There are several varieties of gourami. Favorites among fanciers are the genus *Colisia*, paradise fishes (genus *Macropodus*), and the croaking gourami (*Trichopsis vittatus*), the males of which make an odd croaking noise, particularly at night, when they come to surface for air.

See accompanying illustration.

LACCOLITH. An intrusive type of igneous rock. Studies of the forms of igneous rock masses have shown that the openings followed by volcanic lavas in their upward journey toward the surface are of two dominant types, *tubular* and *tabular*. Tubular openings are approximately circular in outline. These openings may vary from a few centimeters to many meters. Dikes are thin, tabular, parallel-walled masses of igneous rocks. Essentially, they have a vertical position and are formed by the injection of lava into fissures and joints in rocks. Sills are another form, but they are flat and essentially horizontal. Laccoliths, although somewhat similar to sills, differ from them in that the overlying beds are arched. The horizontal area occupied by a laccolith usually is smaller than that occupied by a sill. It is believed that the lava



Species of labyrinth fishes.

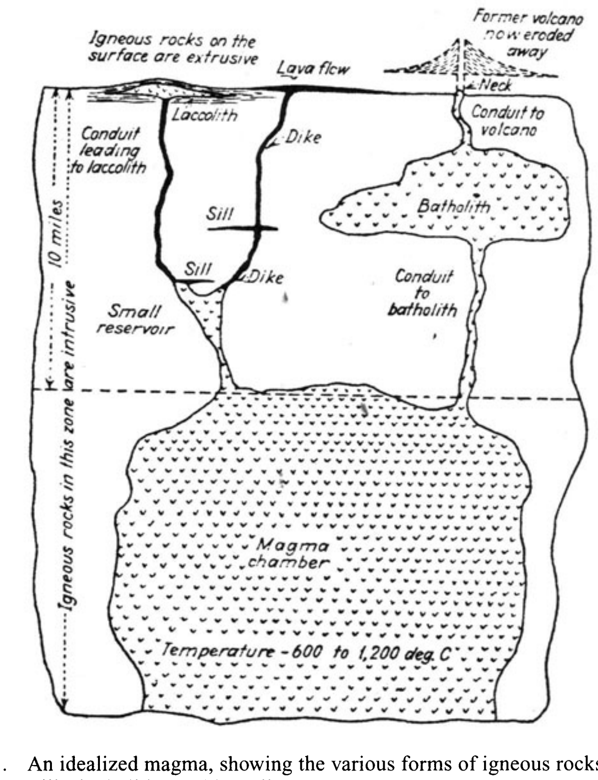


Fig. 1. An idealized magma, showing the various forms of igneous rocks, such as dikes, sills, batholiths, and laccoliths.

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

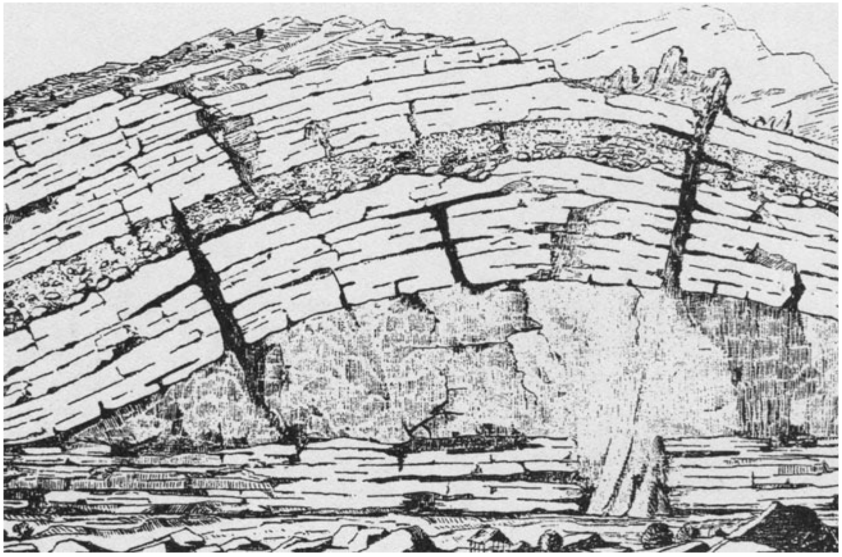


Fig. 2. An idealized laccolith, with associated dikes and sills.

forming a laccolith was too viscous to flow far between the beds of the rock and thus pushed them up to form a dome. A laccolith generally is fed through a conduit. Laccoliths may merge into sills, and they are also commonly associated with dikes. Several laccoliths may occur in the same area. Laccoliths form numerous buttes and mountains in the western United States. See Figures 1 and 2.

The formal definition (American Geological Institute) is "A concordant igneous intrusion with a known or assumed flat floor and a postulated dikelike feeder somewhere beneath its thickest point. It is generally lenslike in form and roughly circular in plan, less than 5 miles in diameter and from a few meters to several hundred meters in thickness."

LACHRYMATOR. A chemical substance that causes tears to form in the eyes. See **Chlorinated Organics.**

LACQUER. See **Paints and Coatings; Resins (Natural).**

LACTIC ACID. Alpha-hydroxy-propionic acid, H—$C_3H_5O_3$, formula weight 90.05, colorless liquid, mp 18°C, bp 122°C, sp gr 1.248, miscible with water, alcohol, or ether in all proportions. The substance exists in two forms: (1) *dextro* lactic acid, which rotates the plane of polarized light to the right; and (2) *levo* lactic acid which rotates the plane of polarized light to the left. A mixture of these two forms is ordinary lactic acid, which does not rotate the plane of polarized light. Ordinary lactic acid is termed dextrolevo lactic acid. Lactic acid is a product of corn refining.

Lactic acid was one of the first biological substances to be investigated from the standpoint of the existence of the two optically active forms.

Lactic Acidosis. Lactic acid is the cause of one of many possible disorders in human acid-base metabolism. Lactic acidosis represents an accumulation of lactic acid in the blood and tissues. This condition gradually depletes the natural buffers in the body and there is a consequent lowering of pH. As described in the entry on **Glycolysis,** lactic acid is the end product of that process. Lactic acid blood levels are determined by at least four factors. The rate of generation of lactic acid; the rate of transport from tissues to plasma and from plasma to the liver (point of utilization of lactic acid); the rate of utilization; and excretion of lactic acid by the kidneys. Normally, all of these functions are maintained in balance to give a normal blood lactate concentration of about 1 mEq/1.

On the generation side, three factors are involved. (1) The availability of oxygen is a major controlling determinant of lactic acid generation because, as adenosine triphosphate (ATP) generation from oxidative phosphorylation diminishes, the cells naturally respond with a greater rate of glycolysis. This increases tissue lactate levels and ultimately lactate blood levels. See also **Phosphorylation (Oxidative).** (2) If, as may be caused by various factors, there is an increase in pH, the activity of phosphofructokinase will increase (this is the rate-limiting enzyme of glycolysis). With increases in pH, the enzyme is more active and more lactate is formed. (3) Factors which affect the biological oxidation-reduction potentials also influence the rate at which glucose is metabolized to lactate.

Fundamental predisposing conditions causing an increased generation of lactate include: decreased tissue perfusion associated with shock, which may occur in cardiac arrest; increased skeletal muscle activity (the rate of glycolysis increases with exercise; this also may be associated with convulsive states that may follow severe exercise—brought about by increased blood lactate concentrations); large tumors, since tumors (leukemias, lymphomas, etc.) may have an increased rate of glycolysis even in the presence of a sufficient supply of oxygen; and both cyanide and carbon monoxide poisoning, which can increase lactate levels because of insufficient oxygen supply.

On the utilization side, there are a number of influencing factors. The liver is the principal lactic acid utilization center. In liver failure, a surplus of lactate builds up, a condition which may be associated with reduced hepatic perfusion, hepatocyte failure, and hepatocytes replaced by tumor. Blood lactate concentrations are elevated in persons with diabetic ketoacidosis. The observed elevation of lactic acid levels in cases of alcoholism is not fully understood: the condition may increase generation or by decreasing utilization. The latter effect is now favored by many authorities, the theory being that ethanol completes for electrons in the liver, thus decreasing utilization of lactic acid in that organ.

LACTIC ACIDOSIS. See **Lactic Acid.**

LACTOSE. See **Carbohydrates.**

LACTULOSE. See **Sweeteners.**

LADY BEETLE (*Insecta, Coleoptera*). Small oval beetles, strongly convex and with relatively small legs. The common name applies chiefly to the more common red species, marked with black and white, but the family *Coccinellidae* to which they belong contains many others.

Lady beetles are found on plants and trees and deposit their eggs on the underside of leaves. They are practically round in shape and quite small, approximately $\frac{1}{8}$ inch (3 millimeters) across. The color varies with the species. They are harmless and often carried around by children as "pets."

The wormlike maggots eat plant lice. Aphids are a favorite food and thus the presence of lady bugs helps with gardening and growing flowers. See also entry on **Beneficial Insects.**

LAG (Angle of). When two related quantities, such as an alternating voltage and an alternating current, vary sinusoidally with time and have the same frequency, they may be expressed as

$$Q_1 = A\begin{Bmatrix}\sin\\ \cos\end{Bmatrix}(\omega t + \phi)$$

$$Q_2 = B\begin{Bmatrix}\sin\\ \cos\end{Bmatrix}\omega t$$

where A, B, and ω are constants. It is then said that Q_2 lags (behind) Q_1 and ϕ is known as the angle of lag if it is positive. If ϕ is negative its magnitude is the angle of lead and Q_2 is daid to lead Q_1.

LAGOON. See **Estuary.**

LAGRANGE FORMULA FOR INTERPOLATION. Used when $(n + 1)$-pairs of values are given for $y = f(x)$, but not necessarily at equally spaced increments of x or y. Let the given number pairs be (x_0, y_0), (x_1, y_1), . . ., (x_n, y_n). Then for any desired value of x within this interval,

$$y = y_0L_0^{(n)}(x) + y_1L_1^{(n)}(x) + \cdots + y_nL_n^{(n)}(x)$$

where

$$L_i^{(n)}(x) = \frac{(x - x_0)\cdots(x - x_{i-1})(x - x_{i+1})\cdots(x - x_n)}{(x_i - x_0)\cdots(x_i - x_{i-1})(x_i - x_{i+1})\cdots(x_i - x_n)}.$$

The quantities L_i, known as Lagrange coefficients, are independent of y; hence they may be calculated once for a given set of x values and used unchanged to obtain results for varying y. Moreover, it will be found that they remain unchanged with a change of variable to $u = (x - a)/h$, where h and a are constants.

Because of the symmetry in the equation, x and y may also be interchanged so that inverse interpolation may be effected.

See also **Interpolation.**

LAGRANGIAN COORDINATES. Sometimes called material coordinates. A system of coordinates by which fluid parcels are identified for all time by assigning them coordinates which do not vary with time. Examples of such coordinates are (a) the values of any properties of the fluid conserved in the motion; or (b) more generally, the positions in space of the parcels at some arbitrarily selected moment. Subsequent positions in space of the parcels are then the dependent variables, functions of time and of the Lagrangian coordinates.

Few observations in meteorology are Lagrangian: this would require successive observations in time of the same air parcel. Exceptions are the constant-pressure balloon observation, which attempts to follow a parcel under the assumption that its pressure is conserved, and certain small-scale observations of diffusion particles. See also **Eulerian Coordinates.**

LAGRANGIAN FUNCTION. Also called kinetic potential, the difference between the kinetic energy and the potential energy of a dynamic system. It is generally symbolized by L.

LAGUERRE DIFFERENTIAL EQUATION. The linear equation $xy'' + (1 - x)y' + ny = 0$, having a simple pole at the origin. Its solutions are the *Laguerre polynomials*. Differentiation of the equation k times and replacement of the kth derivative by y gives

$$xy'' + (k + 1 - x)y' + (n - k)y = 0$$

which is the associated Laguerre equation with solutions as associated Laguerre polynomials. These functions occur in the quantum mechanical problem of the hydrogen atom.

The associated polynomials may be defined by the equivalent expressions

$$L_n^{(k)}(x) = \frac{e^x x^{-k}}{n!}\frac{d^n}{dx^n}(e^{-x}x^{n+k})$$

$$= \sum_{i=0}^{n}\binom{n+k}{n-i}\frac{(-x)^i}{i!}$$

The special case of $k = 0$ gives the Laguerre polynomials

$$L_n(x) = 1 - \binom{n}{1}x + \binom{n}{2}\frac{x^2}{2!} - \binom{n}{3}\frac{x^3}{3!} + \cdots$$

Both kinds may also be expressed in terms of the Gauss hypergeometric series and by generating functions. They are also related to the Hermite polynomials and the Bessel functions.

LAKE. See **Earth; Limnology,.**

LAKES (Colors). See **Colorants (Foods); Dyes (Textile).**

LAMBDA PARTICLE. A hyperon with a rest-mass energy of 1115.6 MeV, an isospin quantum number zero, an angular momentum spin quantum number $\frac{1}{2}$, and a strangeness quantum number 1. Symbol, λ.

LAMBERT. See **Units and Standards.**

LAMBERT PROJECTION. The Lambert modified conformal conic projection (commonly known as the Lambert, and sometimes as the Gauss conformal) has been used for many years in the construction of maps. During this century, this projection gained favor rapidly for use in constructing charts for air navigators.

The projection is actually a mathematical type, but may be quite accurately described as a conical projection. It differs from the simple conical in that the cone is not tangent to the earth's surface, but cuts it on two latitude parallels known as standard parallels. This type of projection is particularly valuable for portraying large longitudinal areas, e.g., the entire United States. The graticule of the Lambert chart shows parallels of latitude as concentric circles centered at the nearer pole, and the meridians of longitude as straight lines converging on this pole. From simple geometric considerations, it is obvious that the meridians and parallels must be perpendicular. The angle of convergence of the meridians, and therefore the radii of the parallels, depends upon the distance of the nearer pole from the center of the area.

The great advantage of the Lambert projection is that the scale of distance is uniform, for all practical purposes, all over the chart. To indicate the accuracy of this statement, consider the Lambert projection on which the series of aeronautical charts of the United States are constructed by the Coast and Geodetic Survey. The standard parallels for these charts are N 45° and N 33°. The scale of distance, if considered as unity on the standard parallels, is 0.994 at the central parallel, expands to 1.010 at the extreme north boundary of the United States, and to 1.023 at the tip of Florida. Comparing the Lambert with the mercator distance scales, we find that if we consider the mercator scale as unity at N 39°, the scale at the northern limits of the United States is 1.154, whereas, at the tip of Florida, it is 0.846.

The nonorthogonal graticule of the Lambert chart is a distinct disadvantage for general navigational problems, since neither the rhumb line nor the great circle is straight. However, the uniformity of scale is of such great advantage that air navigators have begun to use this type of chart even in preference to the mercator, particularly when navigating in good visibility over land, or where good radio aids are available.

A straight line between two points on a Lambert chart is referred to as a Lambert line. Since the meridians on the graticule are convergent, this will not be a rhumb line. However, the distance measured along this line will be less than that along the rhumb line, and only slightly greater than the great-circle distance between the two points. This line is frequently used by aviators during conditions of good visibility. The standard procedure in using the Lambert line is to draw a straight line on the chart and pick out conspicuous landmarks separated by about 25 miles. The rhumb line indicated by measurement of the angle between the Lambert line and the meridian nearest the starting point is then followed. This will lead to a point at some distance from the first landmark but within visibility. From this first landmark, a new heading is adopted for the second, and so on to destination. If a rhumb line is desired for the entire route, the Lambert line is drawn as before, and the course is measured from the meridian halfway between the point just left behind and the point to be arrived at. This rhumb line will appear as a curve on the Lambert chart, but can be laid down by any one of a number of standard methods with sufficient accuracy for the selection of landmarks.

The convergence of the meridians prevents the use of the Lambert chart for graphical solution of dead-reckoning problems, and introduces difficulties in plotting lines of position obtained either by radio bearings or by celestial observation. The ease with which such problems can be solved on the mercator chart seems to cast doubt on any statement to the effect that "the Lambert Chart will completely supersede the mercator for all navigational purposes."

See also **Course; Great-Circle Course; Line of Position; Mercator Sailing; Navigation;** and **Rhumb Line.**

LAMBERT'S COSINE LAW. The intensity from a surface element of a perfectly diffuse radiator is proportional to the cosine of the angle between the direction of emission and the normal to the surface. An element of a surface which obeys this law will appear equally bright when observed from any direction.

LAMB SHIFT. The displacement between the $2S_{1/2}$ and $2P_{1/2}$ levels of hydrogen, which in the absence of radiative corrections would be zero due to the Coulomb degeneracy. The experimental value obtained by Lamb and Rutherford

$$E_{2S_{1/2}} - E_{2P_{1/2}} = 1057.8 \pm 0.1 \text{ megahertz}$$

is in agreement with the theoretical value. Of this, 27 MHz arises from vacuum polarization, the rest from self-energy corrections. The term is now used to indicate the displacement of any bound state level due to radiative corrections. See also **Field Theory.**

LAMÉ EQUATION (Generalized). The most general second-order (linear) differential equation with five regular singularities, one of them being the point at infinity, with preassigned exponents differing from each other by $\frac{1}{2}$ at each singularity, all other points of the complex plane being ordinary points. This equation is remarkable because of the large number of important equations (Legendre, Bessel, etc.) obtainable from it by confluence. Letting a_1, a_2, a_3, a_4 and ∞ be the singular points, with exponents $\alpha_1, \alpha_1 + \frac{1}{2}, \ldots \alpha_4, \alpha_4 + \frac{1}{2}, \mu_1, \mu_1 + \frac{1}{2}$, the equation has the form

$$\frac{d^2w}{dz^2} + P\frac{dw}{dz} + Qw = 0$$

where

$$P = \sum_{i=1}^{4} \frac{\frac{1}{2} - 2\alpha_i}{z - a_i},$$

$$Q = \sum \frac{\alpha_i(\alpha_i + \frac{1}{2})}{(z - a_i)^2} + \frac{Az^2 + 2Bz + C}{(z - a_1)\cdots(z - a_4)}.$$

A is expressible in terms of the α_i, and B, C are arbitrary constants. For example if the confluence takes the form $a_1 = a_2 = 0$, $a_3 = a_4 = \infty$, then choosing all $\alpha_i = 0$ and setting $z = \zeta^2$, with proper choice of B and C, we get

$$\zeta^2\frac{d^2w}{d\zeta^2} + \zeta\frac{dw}{d\zeta} + (\zeta^2 - n^2)w = 0,$$

which is Bessel's equation.

LAMELLA (Botany). The middle lamella is the compound layer composed of the primary walls and the cement-like intercellular substance which occurs between the primary walls of two cells. This composite layer is usually made up of pectic materials (colloids which have a great affinity for water), one of which is calcium pectate. The function of the middle lamella is to hold adjoining cells together. Sometimes, particularly in mature fruits, the middle lamella substance breaks down. As a result, the cells of the fruit separate easily, giving to the fruit a meal-like character.

The middle lamella is the common source of pectin which is added to concentrated fruit juices in the preparation of jellies. The term lamella is also applied to each of the concentric growth layers in large starch grains.

LAMELLAE. See **Bone.**

LAMELLA (Zoology). 1. A thin leaf or plate, such as a lamella of bone. 2. A flat plate formed by the fusion of ctenidial filaments in the bivalve mollusks. Two lamellae united by bridges of tissue form a gill through which water circulates under the influence of ciliary action in the persisting open spaces. This form of gill is the source of the name *Lamellibranchiata* applied to the class containing these animals.

LAMELLIBRANCHIATA. The bivalve mollusks, a class of the phylum *Mollusca* including the clams, mussels, oysters, scallops and related species. Many of these animals are valuable for food, and pearls and mother-of-pearl are produced by them. The class is also named *Pelecypoda*.

Bivalve mollusks differ from other members of the phylum in the following characters: (1) The body is bilaterally symmetrical and transversely compressed. (2) The mantle forms two lobes extending down along the sides of the body. In most species these lobes unite at the posterior end to form two passages, an upper excurrent and a lower incurrent siphon. Currents of water carry food and oxygen into the mantle cavity through the lower opening and a current bearing wastes passes out of the upper. (3) Each mantle fold secretes a valve of the shell formed of calcareous matter covered outside by a horny periostracum and inside by nacre, commonly called mother-of-pearl. The two valves of the shell are joined by a hinge ligament and the articulation is strengthened in some species by interlocking teeth. They are closed ventrally by the contraction of one or two adductor muscles. (4) The gills are thin plates on each side of the body in most species. They are formed of united ctenidial filaments. (5) The foot is a muscular wedge-shaped protuberance at the anterior end of the body. The head is rudimentary.

All bivalves are aquatic. Most species creep slowly by thrusting the foot into the muddy or sandy bottom but some propel themselves by jets of water squirted from the siphons or forced from the mantle cavity by rapidly closing the valves. The species vary from freshwater forms about $\frac{1}{8}$ inch long to giant marine shells more than a yard long.

The class is divided into four orders:

Order *Protobranchiata*. Gills in the form of small leaflets, two rows on each side of the body. Marine species.

Order *Filibranchiata*. Marine mussels, scallops, etc. Gills composed of filaments united only by ciliary junctions.

Order *Eulamellibranchiata*. Gill filaments united to form continuous plates. Freshwater clams or mussels, marine clams, oysters, shipworms, etc.

Order *Septibranchiata*. Gills replaced by a horizontal partition between the upper and lower divisions of the mantel chamber. A few marine species.

See also **Clam; Mollusca; Mussel; Oyster; Pearl**.

LAMINAR FLOW. A condition of fluid flow in a closed conduit in which the fluid particles or "streams" tend to move parallel to the flow axis and not mix. This behavior is characteristic of low flow rates and high viscosity fluid flows. As the flow rate increases (or viscosity significantly decreases), the streams continue to flow parallel until a velocity is reached where the streams waver and suddenly break into a diffused pattern. This point is called the *critical velocity*. See also **Turbulent Flow.**

Laminar flow is characterized by a parabolic flow profile where the maximum velocity at or near the center of the conduit is approximately twice the average velocity in the profile. Laminar flow often is referred to as *viscous flow*, streamline flows, and *low-Reynolds number flow*. Special attention must be paid to the constancy of coefficient of most flowmeters in the region of laminar flow. See also **Reynolds Number.**

Laminar Sublayer. When a fluid is in turbulent flow past a rigid surface, fluctuations of velocity in the direction normal to the surface are inhibited, and very close to the surface they may be negligible. Then the Reynolds shear stress is small compared with the viscous stresses, and it has been common to describe the region as a laminar sublayer. In fact, turbulent fluctuations of velocity in planes parallel to the wall are considerable in comparison with the mean velocity.

See also **Aerodynamics; Fluid and Fluid Flow.**

LAMP. See **Illumination.**

LAMPBLACK. See **Carbon black**

LAMPREYS (*Agnatha*). A jawless fish of the family *Petromyzontidae*, the lamprey appears much like an eel. However, it is *not* an eel. Normally, various species of lampreys occur on both sides of the Atlantic. Like its close relative, the hagfish, the lamprey is characterized by the primitive features of jawless fishes—no scales, no sympathetic nervous system, a cartilage skeleton, and single nostril. Since about the mid-1800s, lampreys have not been considered of commercial value. However, they were eaten during the middle ages. Probably the parasitic landlocked lamprey *Pretomyzon marinus* is best known because of the very extensive damage it has done to many freshwater fish species in the Great Lakes. The lamprey attaches itself to a host and literally sucks life-giving juices from it. In the saliva of the lamprey, there is an anticoagulating material which continually dilutes the blood of its victim. Once the host is drained of vital juices, the lamprey moves on to another victim. As an example of this damage, just a few decades ago, the Great Lakes yielded an annual catch of nearly 12 million pounds (5.4 million kilograms) of lake trout. Within a period of about 30 years, lake trout practically disappeared from the lakes. Initially, the lampreys migrated from their normal marine-water habitat to fresh water for spawning. Inasmuch as the Great Lakes are interconnected to the sea through various waterways, including the Welland Canal and the New York State Barge Canal, the lampreys ultimately invaded the lakes. During the latter 1970s, considerable progress was made toward specifically combating the lampreys without harming other species. Population of trout in the lakes is again rising.

LANGLEY. See **Solar Energy.**

LANGUAGE (Computer). A communications means for transmitting information between human operators and computers. The human programmer describes how the problem is to be solved using the computer language. A computer language consists of a well-defined set of characters and words, coupled with a series of rules (termed syntax) for combining them into computer instructions or statements. There is a wide variety of computer languages, particularly in terms of flexibility and ease of use. There are three levels in the hierarchy of computer languages: (1) machine languages; (2) procedure-oriented languages; and (3) problem-oriented languages.

Machine Language (1) A language designed for interpretation and use by a machine without translation. (2) A system for expressing information which is intelligible to a specific machine; e.g., a computer or class of computers. Such a language may include instructions which define and direct machine operations, and information to be recorded by or acted upon by these machine operations. (3) The set of instructions expressed in the number system basic to a computer, together with symbolic operation codes with absolute addresses, relative addresses, or symbolic addresses. In this case, it is known as an Assembler Language. See **Assembler (Computer System).**

Procedure-oriented Language. A machine-independent language which describes how the process of solving the problem is to be carried out. For example, FORTRAN, ALGOL, PL/I, and COBOL.

Problem-oriented Language. A language designed for convenience of program specification in a general problem area. The components of such a language may bear little resemblance to machine instructions and often incorporate terminology and functions unique to an application. Also known as Applications Language.

Other computer languages include:

Algorithmic language. An arithmetic language by which numerical procedures may be precisely presented to a computer in a standard form. The language is intended not only as a means of directly presenting any numerical procedure to any appropriate computer for which a compiler exists, but also as a means of communicating numerical procedures among individuals.

Artificial language. A language specifically designed for ease of communication in a particular area of endeavor, but one that is not yet "natural" to that area. This is contrasted with a natural language which has evolved through long usage.

Common machine language. A machine sensible information representation which is common to a related group of data processing machines.

Common business oriented language. A specific language by which business data processing procedures may be precisely described in a standard form. The language is intended not only as a means for directly presenting any business program to any appropriate computer for which a compiler exists, but also as a means of communicating such procedures among individuals.

Object language. A language which is the output of an automatic coding routine. Usually, object language and machine language are the same; however, a series of steps in an automatic coding system may involve the object language of one step serving as a source language for the next step and so forth.

Thomas J. Harrison, International Business Machines Corporation, Boca Raton, Florida.

LANGUR. See **Monkeys and Baboons.**

LANTERN FISHES. See **Iniomous Fishes.**

LANTERN FLY (*Insecta, Homoptera*). Any member of the family *Fulgoridae*, which differs from the related leaf hoppers and other families in having the antennae inserted at the sides of the head. The North American species are small but one giant Brazilian species has a wing spread of 6 inches. A large prominence on the head of this species was once said to be luminous, hence the name lantern fly has persisted although none of these insects is actually luminous.

LANTHANIDE CONTRACTION. The decreasing sequence of crystal radii of the tripositive rare-earth ions with increasing atomic number in the group of elements (57) lanthanum through (71) lutetium of the Lanthanide Series in the periodic table.

LANTHANIDE SERIES. The chemical elements with atomic numbers 58 to 71 inclusive, commencing with cerium (58) and through lutetium (71) frequently are termed collectively, the Lanthanide Series. Lanthanum, the anchor element of the series, appears in group 3b of the periodic table. Some authorities consider lanthanum a part of the series. Members of the series, along with lanthanum and yttrium, are described under **Rare-Earth Elements and Metals.** See also **Actinide Series.**

LANTHANUM. Chemical element symbol La, at. no. 57, at. wt. 138.91, periodic table group 3, homolog of the Lanthanide Series of elements, mp 918°C, bp 3464°C, density 6.146 g/cm^3 (20°C). Elemental lanthanum has a double close-packed hexagonal crystal structure at 25°C. The pure metallic lanthanum is silver-gray in color, but with a luster that remains only briefly upon exposure to air, rapidly oxidizing to a white powder. The oxide is hygroscopic and tends to spall, thus exposing fresh surfaces of the metal for oxidation. Thus, the metal must be handled in an inert atmosphere. Chips and powdered lanthanum are quite pyrophoric. Under required inert atmospheric conditions, the metal is easy to work with normal tools, paralleling tin in its workability. There are two natural isotopes ^{139}La and ^{138}La. The latter is mildly radioactive with a half-life of $10^{10}-10^{15}$ years. The element becomes a superconductor below 6K. There are 19 known artificial isotopes, all radioactive. Of the light (or cerium-group) rare-earth metals, lanthanum is the second most plentiful and ranks 57th in abundance of elements in the earth's crust, exceeding gold, tantalum, platinum, mercury, bismuth, and several other commonly-used elements. The element was first identified by C. G. Mosander in 1839. Electronic configuration

$$1s^22s^22p^63s^23p^63d^{10}4s^24p^64d^{10}5s^25p^65d^16s^2.$$

Ionic radius La^{3+} 1.061Å. Metallic radius 1.879Å. First ionization potential 5.571 eV; second, 11.06 eV. Oxidation potentials $La \rightarrow La^{3+} + 3e^-$, 237 V; $La + 3OH^- \rightarrow La(OH)_3 + 3e^-$, 2.76 V.

Other important physical properties of lanthanum are given under **Rare-Earth Elements and Metals.**

Much of the commercial lanthanum production uses bastnasite, a rare-earth fluorocarbonate found in Southern California, as the source. See also **Bastnasite.** The element is separated from other rare-earth elements in an ion-exchange process after acid leaching of bastnasite (or monazite) minerals. Pure lanthanum is obtained by (1) electrowinning from the oxide La_2O_3 in a molten fluoride electrolyte, (2) electrolysis of fused anhydrous $LaCl_3$, or (3) metallothermic reduction of LaF_3 by calcium in a reactor under an inert atmosphere.

Lanthanum metal dissolves readily in dilute mineral acids. The oxide is dissolved by concentrated mineral acids and acetic and formic acids. The metal is a component of mischmetal used for lighter "flints" and in the cores of carbon electrodes for high-intensity lighting. See also **Cerium.** Several of the best grades of optical glass require pure lanthanum oxide as an ingredient for lowering the dispersion of light and for improving the index of refraction. The oxide melts at 2,310°C. The oxide ranks eleventh among the most refractory metal oxides, but finds limited use because of its highly hygroscopic nature. The oxide also is used as a host matrix for fluorescent phosphors and in thermistors and capacitors and other elements of electronic circuitry. By far, the largest use of lanthanum (mixed with other rare-earths) is for molecular-sieve catalysts for cracking crude petroleum. This use has required over 10 million pounds of lanthanum rare-earth chloride annually.

As an alloying metal, lanthanum finds broad use. Although lacking mechanical strength, lanthanum has a high affinity for oxygen, sulfur, nitrogen, and hydrogen and thus makes an effective component for scavenging gases from molten metals. Cobalt-base alloys containing lanthanum have shown increased resistance to oxidation and hot corrosion. One of the intermetallic compounds of lanthanum $LaCo_5$ possesses excellent magnetic properties—well in excess of those of alnico and platinum cobalt permanent magnets. The intermetallic compound $LaNI_5$ shows exceptional properties for absorbing and desorbing large amounts of hydrogen at room temperature.

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.

LAPIS LAZULI. See **Lazurite.**

LAPLACE EQUATION. A second-order partial differential equation of elliptic type which, in vector form, is $\nabla^2\phi = 0$. It is the homogeneous case of Poisson's equation. Its solutions, the scalar quantity ϕ, occur in problems involving steady-state temperatures, gravitational and electric potentials, hydrodynamics of ideal fluids, and many other physical phenomena. The equation is usually solved by the method of separation of variables in a suitable curvilinear coordinate system and with boundary conditions imposed by physical requirements. Such solutions are called harmonic functions. In two dimensions, an analytic function of the complex variable must satisfy Laplace's equation. Thus, $w = u + iv$ is an analytic function of $z = x + iy$ for $u_{xx} + u_{yy} = 0$ and $v_{xx} + v_{yy} = 0$.

We shall here cite two familiar physical examples to which Laplace's equation applies; there are many others.

1. Consider a region of space in which there is an electric field (due to electric charges in the vicinity) but no free electricity, that is, no such space charge as exists in a vacuum tube in operation. At any point in this space there is an electric potential, which varies with the position of the point and is therefore a function of its coordinates. It is shown in electrostatic theory that if the potential satisfies Laplace's equation, it is possible to trace the lines of force and equipotential surfaces in the region, by means of special solutions of Laplace's equation, called harmonic functions, which satisfy the "boundary conditions" imposed by the arrangement of the neighboring charges. As to which form of the equation is to be used, this depends upon which system of coordinates is most conveniently adapted to the shape and arrangement of the charged bodies.

2. If specified parts of the surface of a solid thermal conductor (such as a block of copper) are kept at different specified constant temperatures, heat will flow from the warmer toward the colder boundaries within the conductor. When this flow has become steady, the temperature of the conductor takes on a definite constant value at each point, dependent upon the location of the point. If the temperature is designated by u, it now satisfies Laplace's equation, and the lines of flow and isothermal surfaces may be determined accordingly. A similar procedure is applied to potential in the steady flow of electricity through a metallic conductor.

LAPLACE THEOREM. This limit theorem, of which the Bernoulli theorem is a corollary, states that if there are n independent trials, in each of which the probability of an event is p, and if this event occurs k times, then

as $n \rightarrow \infty$, whatever the numbers z_1 and z_2. Roughly speaking, the theo-

$$P\left\{z_1 \leq \frac{k - np}{\sqrt{npq}} \leq z_2\right\} \rightarrow \frac{1}{\sqrt{2\pi}} \int_{z1}^{z2} e^{-1/2z^2}\, dz$$

rem states that the number of successes k in n trials is normally distributed for large n.

LAPLACE TRANSFORM. This class of integral transform has so many applications in modern technology that it justifies the more extended discussion in this special entry.

The behavior of ac electric circuits is readily expressed in terms of integro-differential equations, as is understood by recalling that the voltage drop across an inductance is a differential function of the current, whereas the voltage across a capacitance is an integral function of the current. Moreover, many types of nonelectrical systems—mechanical, acoustical, hydraulic—lend themselves conveniently to analysis by analogy to electrical systems, so providing a convenient means of generalizing and unifying the analysis of complex systems involving subsystems of various kinds. This fact, together with the adaptability of the Laplace transform method to the solution of differential and integral equations, and the analysis of discontinuous functions, has brought about its wide application to control problems.

The Laplace transform of a function $f(t)$ is defined as the function given by

$$\mathcal{L}[f(t)] = \int_0^\infty f(t)e^{-st}dt = F(s) \qquad (1)$$

Here the double equation means that the Laplace transform may be written as $\mathcal{L}[f(t)]$ or as $F(s)$, since the transformation process results in a new function $F(s)$ of a variable s instead of the original function f of time, t. The variable s may be real but is usually complex. The transformation is effected by integrating the function $f(t)$ by the use of the factor e^{-st}, where e is the natural logarithmic base (2.71828 . . .). The integral is, as shown, in the improper form, its upper value being considered to increase indefinitely, approaching ∞ as a limit.

Evaluating Laplace Transforms. One of the simplest discontinuous functions is the *unit function*, which has only two values, 1 for positive or zero values of t, and 0 for negative values of t; it is often expressed as $u(t)$. Thus for values of $t \geq 0$,

$$\mathcal{L}[u(t)] \quad \mathcal{L}(1) = \int_0^\infty 1e^{-st}dt = -\frac{1}{s}e^{-st}\Big]_0^\infty = -\frac{1}{s}0 - \left(-\frac{1}{s}1\right) = \frac{1}{s} \tag{2}$$

The *unit ramp function* is a linear and continuous function of t, thus expressed as $f(t)$ for positive or zero values of t, so that

$$\mathcal{L}[f(t)] = \int_0^\infty te^{-st}dt = -\frac{te^{-st}}{s}\Big]_0^\infty + \frac{1}{s}\int_0^\infty e^{-st}dt = 0 + \left[\frac{1}{s}\cdot\frac{-1}{s}e^{-st}\right]_0^\infty = \frac{1}{s^2} \tag{3}$$

The Laplace transform of a *power of t* is found by a change of the variable of integration, as from t to $w = st$, where s is real and positive:

$$\mathcal{L}(t^p) = \int_0^\infty w^p e^{-st}dt = \frac{1}{s^{p+1}}\int_0^\infty w^p e^{-w}\,dw \tag{4}$$

The last integral at the right is essentially the gamma function, which is given in tables, and which is a generalization of the factorial $n!$, that is defined only for integers and zero. The gamma function may be written as

$$\Gamma(p) = \int_0^\infty w^{p-1}e^{-w}$$

and when $p = n$, the value of the function reduces to $n!$. Therefore, we can write for Equation (5),

$$\mathcal{L}(t^p) = \frac{1}{s^{p+1}}\Gamma(p+1) \tag{5}$$

To find the Laplace transform of the *exponential function* e^{-ct} (for positive values of t) where c is a real number:

$$\mathcal{L}(e^{-ct}) = \int_0^\infty e^{-ct}e^{-st}dt = \int_0^\infty e^{-(s+c)t}dt = -\frac{1}{s+c}e^{-st}\Big]_0^\infty = \frac{1}{s+c} \tag{6}$$

To find the Laplace transform of the *trigonometric function* sin at,

$$\mathcal{L}(\sin at) = \int_0^\infty \sin at\; e^{-st}dt$$

where a is a real, positive number:

Now in the preceding examples, it was found that the evaluation of the Laplace transform of an exponential was quite simple because e^{-st} is also an exponential. So in the present example we replace sin at by its exponential expression $(e^{iat} - e^{-iat})/2i$, where $i = \sqrt{-1}$, giving

$$\mathcal{L}(\sin at) = \frac{1}{2i}\int_0^\infty (e^{iat} - e^{-iat})e^{-st}dt = \frac{1}{2i}\left(\frac{1}{s-ia} - \frac{1}{s+ia}\right) = \frac{a}{s^2+a^2} \tag{7}$$

A similar substitution is used to find the Laplace transform of cos at. Since *cos at* $= (e^{iat} + e^{-iat})/2i$, we have

$$\mathcal{L}(\cos at) = \frac{1}{2i}\int_0^\infty (e^{iat} + e^{-iat})e^{-st}dt = \frac{s}{s^2+a^2} \tag{8}$$

Properties of the Laplace Transform. An important property of the Laplace transform is its linearity, so that the Laplace transform of the sum of two functions is equal to the sum of the individual Laplace transforms

$$\mathcal{L}[f_1(t) \pm f_2(t)] = F_1(s) \pm F_2(s) \tag{9}$$

Here the original functions are denoted by $f_1(t)$ and $f_2(t)$, and their Laplace transforms by $F_1(s)$ and $F_2(s)$.

(I) The full statement of the *linearity* (also called *superposition*) *theorem* introduces two constants (a and b below):

$$\mathcal{L}[af_1(t) \pm bf_2(t)] = aF_1(s) + bF_2(s) \tag{10}$$

conveying the further information that constant factors of functions appear unchanged in their Laplace transforms.

Since by Equation (2) the Laplace transform of $t = 1$ was found to be $1/s$, then it follows directly that the Laplace transform of any constant is expressed by

$$\mathcal{L}(c) = c\mathcal{L}(1) = \frac{c}{s} \tag{11}$$

It also follows from the linearity theorem that the Laplace transform of any polynomial is equal to the sum of the Laplace transforms of its terms:

$$\mathcal{L}(a_0 + a_1t + a_2t^2 + \cdots + a_nt^n) = a_0\mathcal{L}(1) + a_1\mathcal{L}(t) + a_2\mathcal{L}(t^2) + \cdots + a_n\mathcal{L}(t^n) \tag{12}$$

The general expression for this series of terms is, therefore,

$$\sum^n a_i\frac{\Gamma(i+1)}{s^{i+1}} = \frac{a_0}{s} + \frac{a_1}{s^2} + \frac{2!a_2}{s^3} + \frac{3!a_3}{s^4}\cdots\frac{n!a_n}{s^{n+1}} \tag{12a}$$

Note that, as derived, Equation (12a) is valid only for positive and real values of s. In fact, in all operations with Laplace transforms constant attention must be paid to (1) whether the transform of a function exists at all, and, if it does, (2) for what values of the variables it has corresponding real values and for what values it becomes zero or indefinitely great.

(II) The *real differentiation theorem* asserts that if a function and its first derivative have Laplace transforms and if $\mathcal{L}[f(t)] = F(s)$, then

$$\mathcal{L}\frac{df(t)}{dt} = sF(s) - f(t \to 0+) \tag{13}$$

which states that the Laplace transform of the first time derivative of a function is equal to the product by s of the Laplace transform of the function itself minus the value of the function as t approaches 0 from the positive side.

(III) The *real integration theorem* asserts that if a function of t has a Laplace transform and if $\mathcal{L}[f(t)] = F(s)$, then

$$\mathcal{L}\left[\int f(t)\,dt\right] = \frac{1}{s}\left[F(s) + \int_{0+} f(t)\,dt\right] \tag{14}$$

which states that the Laplace transform of the integral of the function is equal to the product by $1/s$ of the Laplace transform of the function plus the value of the integral of the function as t approaches zero from the positive side.

The usefulness of these two theorems is due to the fact, as stated earlier, that so many of the circuit and control equations contain differentials or integrals or both.

(IV) There are two translation theorems used with Laplace transforms. The first of these is the *time-delay theorem*, which is written

$$\mathcal{L}[f(t - t_0)] = e^{-t_0 s}F(s) \qquad (15)$$

where t_0 is real and ≥ 0, provided $f(r) = 0$ for $t < 0$ and, in the case of the $0 +$ transform, $f(t)$ in addition contains no impulse functions at $t = 0$. A direct corollary of the time-delay theorem is the *unit step function*. In Equation (2) it was found that for values of t of unity, the Laplace transform was $1/s$. The unit step function is $u(t - t_0)$, where t has a value of 1 for all positive and zero values of the function, i.e., for all values of t greater than or equal to t_0, white t is zero for all its other values. This function expresses the commencement of a steady-state operation or process of unit value at time t_0.

Equations (2) and (15) give the unit step function directly, for by (2) the Laplace transform of unity is $1/s$, which corresponds to $F(s)$ in (15), giving

$$\mathcal{L}u(t - t_0) = e^{-t_0 s} \cdot \frac{1}{s} = \frac{e^{-t_0 s}}{s}$$

(V) The *time advance theorem* is written

$$\mathcal{L}[f(t + t_0)] = e^{et_0 s}F(s) \qquad (16)$$

Note the change of sign of t_0 on both sides of the equation from those in the time delay theorem of Equation (15). The same conditions apply: that t_0 is real and ≥ 0, provided $f(t) = 0$ for $t < t_0$ and, in the case of the $0 +$ transform, $f(t)$ in addition contains no impulse functions at $t = t_0$.

(VI) The *frequency shift theorem* is written

$$\mathcal{L}[e^{at}f(t)] = F(s - a) \qquad (17)$$

where a is any finite constant, real or complex.

(VII) The *final value theorem* assets that if the function $f(t)$ and its first derivative have Laplace transforms and if $\mathcal{L}[f(t)] = F(s)$, then

$$\lim_{s\to 0} sF(s) = \lim_{t\to\infty} f(t) \qquad (18)$$

provided that $sF(s)$ is analytic on the imaginary axis and on the right half-plane.

(VIII) The corresponding *initial value theorem* asserts that if a function and its first derivative have Laplace transforms, and if $\mathcal{L}\phi f(t)] = F(s)$, then

$$\lim_{s\to\infty} sF(s) = \lim_{t\to\infty} f(t) \qquad (19)$$

provided that the lim $sF(s)$ as s approaches infinity exists.

(IX) The last of these Laplace theorems that is widely used in elementary system calculations is the *convolution theorem*. The convolution of two functions $f_1(t)$ and $f_2(t)$, denoted by $f_1^* f_2 = f_2^* f_1$ is defined by the integral

$$f_1^* f_2 = \int_0^t f_1(t - \tau) f_2(\tau)\, d\tau$$

If $f_1 = t^n$ and $f_2 = t^m$ (n and m, two positive integers), it can be verified by direct calculation that

$$\mathcal{L}\{f_1^* f_2\} = F_1(s)F_2(s)$$

where $F_1(s)$ is the Laplace transform of $f_1(t)$ and $F_2(s)$ is the Laplace transform of $f_2(t)$. Thus, if $f_1 = t^2$ and $f_2 = t^3$, then

$$f_1^* f_2 = \int_0^t (t - \tau)^2 \cdot \tau^3\, dt = \frac{t^6}{60}$$

On the other hand, if

$$\mathcal{L}(t^2) = 2/s^3 \qquad \text{and} \qquad \mathcal{L}(t^3) = 6/s^4$$

then

$$\mathcal{L}(t^2) \cdot \mathcal{L}(t^3) = \frac{12}{s^7} = \mathcal{L}\left(\frac{t^6}{60}\right)$$

From this result it follows immediately, if $f_1(t)$ and $f_2(t)$ can be expanded in 2 convergent power series, that

$$(f_1^* f_2) = F_1(s)F_2(s) \qquad \text{(convolution theorem)} \qquad (20)$$

In fact, the theorem is true in more general cases. The convolution theorem can be used to derive additional Laplace transforms and also to solve so-called "integral equations," i.e., functional equations where the unknown function appears behind an integral sign.

The Inverse Laplace Transform. In solving problems by means of the Laplace transform, the inverse operation of finding functions which correspond to transforms is, of course, as important as the direct one of finding the transforms. For this purpose, it is convenient to have values of transform pairs, which are given in the accompanying table. The justification for the inverse operation rests, of course, upon the assumption of uniqueness, that for each transform there is only one function. This is true, although its proof is not given here.

TABLE OF LAPLACE FUNCTION-TRANSFORM PAIRS

	$f(t)$	$F(s)$
1.	1	$\frac{1}{s}$
2.	c	$\frac{c}{s}$
3.	t	$\frac{1}{s^2}$
4.	t^p	$\frac{\Gamma(p+1)}{s^{p+1}}$
5.	$\sin at$	$\frac{a}{s^2 + a^2}$
6.	$\cos at$	$\frac{s}{s^2 + a^2}$
7.	e^{-et}	$\frac{i}{s + c}$
8.	$e^{at}t^n$	$\frac{n!}{(s - a)^{n+1}}$
9.	$t \sin at$	$\frac{2as}{(s^2 + a^2)^2}$
10.	$t \cos at$	$\frac{s^2 - a^2}{(s^2 + a^2)^2}$

The symbol for the operation of inverse (Laplace) transformation is $\mathcal{L}^{-1}$.

There is one kind of inverse transform, however, that is not readily tabulated. It is the inverse transform of a fraction. Since fractions occur quite frequently in the problems solved by these methods, the following discussion of one case of finding their inverse is important.

Given a Laplace transform in the form of a rational algebraic fraction,

$$F(s) = \frac{A(s)}{B(s)} = \frac{a_m s^m + a_{m-1}s^{m-1} + \cdots + a_1 s + a_0}{b_n s^n + b_{n-1}s^{n-1} + \cdots + b_1 s + b_0} \qquad (21)$$

where the a_i and b_i are all real constants and n and m are positive integers. Where the roots of the denominator are all real or zero and different, a direct method is available to find $\mathcal{L}^{-1}$.

Let the roots of the denominator be $s_1, s_2, \ldots, S_n$. The $F(s)$ takes the form

$$F(s) = \frac{a_m s^m + a_{m-1}s^{m-1} + \cdots + a_1 s + a_0}{(s - s_1)(s - s_2)\cdots(s - s_n)}$$

This equation may be expressed as the sum of partial fractions

$$F(s) = \frac{k_1}{s - s_1} + \frac{k_2}{s - s_2} + \cdots + \frac{k_n}{s - s_n}$$

where the k_i are coefficients to be determined. The procedure is to multiply both sides of the equation by $(s - s_i)$, which is then equated to 0. Thus expressions for each coefficient are obtained of the general form

$$k_i = \left[(s - s_i)\frac{A(s)}{B(s)}\right]_{s=s_i}$$

where the term $A(s)/B(s)$ represents the transforms whose inverse is to be found.

Then using for each coefficient transform pair number 8 in the table of transforms, we have

$$\mathcal{L}^{-1}\left[\frac{k_i}{s - s_i}\right] = k_i e^{s_i t} \tag{22}$$

which can be found for each fraction, so that the inverse transform of $F(s)$ that is sought is the sum of all i values

$$\mathcal{L}^{-1}[F(s)] = \sum^{i} k_i e^{s_i t}, \quad 0 \le t \tag{23}$$

Applications of the Laplace Transform. (a) *Simple resistance-capacitance circuit.* A simple example of the application of the Laplace transform in solving equations is provided by the *RC*-circuit. Let us find, for example, the behavior of the instantaneous current $i(t)$ from the time that switch S in Fig. 1 is closed. The current produced by a given electromotive forces (E) varies inversely as the resistance ($i(t) = E/R$) or $E = i(t)R$, and the integrated current (with respect to time) varies directly as the capacitance ($\int i(t)\,dt = CE$) or ($E = (1/C)\int i(t)\,dt$) so that the equation for a circuit having both capacitance and resistance is

$$Ri(t)\,\frac{1}{C}\int i(t)\,dt = E \tag{24}$$

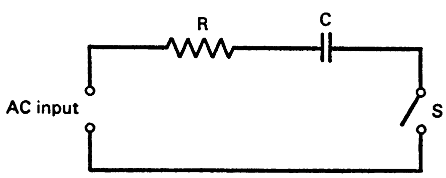


Fig. 1. Resistance-capacitance (*RC*) circuit.

Taking the Laplace transform of both sides, we have

$$\mathcal{L}\left[Ri(t) + \mathcal{L}\left[\frac{1}{C}\int i(t)\,dt\right]\right] = \mathcal{L}(E) \tag{25}$$

By theorem I, and the definition of the Laplace transform, $\mathcal{L}[Ri(t)] = RI(s)$, where $I(s)$ is the effective value of the current, by theorem II,

$$\mathcal{L}\left[\frac{1}{C}\int i(t)\,dt\right] = \frac{1}{sC}\left[I(s) + \int i\,dt(0+)\right]$$

Therefore the equation becomes

$$RI(s) + \frac{1}{sC}\left[I(s) + \int i\,dt(0+)\right] = \frac{E}{s} \tag{26}$$

by using transform pair number 2 from the table to show that $\mathcal{L}(E) = E/s$. If it is assumed that the initial condition of the circuit is with the switch open, then

$$\frac{1}{sC}\int i\,dt(0+) = 0$$

so that

$$RI(s) + \frac{1}{sC}I(s) = \frac{E}{s}$$

Transposing, we have

$$I(s) = \frac{E/s}{R + \dfrac{1}{sC}} = \frac{E}{s\left(R + \dfrac{1}{sC}\right)} = \frac{E/R}{s + \dfrac{1}{RC}}$$

Substitute P for RC, so that

$$I(s) = \frac{E}{R}\cdot\frac{1}{s + 1/P} \tag{27}$$

Taking the inverse transform, we have

$$\mathcal{L}^{-1}[I(s)] = i(t) = \frac{E}{R}\,\mathcal{L}^{-1}\frac{1}{s + 1/P}$$

Then by inverse transform number 7 in the table,

$$i(t) = \frac{E}{R}e^{-t/P} = \frac{E}{R}e^{-t/RC} \tag{28}$$

which is the instantaneous value of the current for given values of E, R, and C, at time t.

(b) *Generalized resistance-capacitance circuit.* In investigating the behavior of control and servo systems, an effective approach is from the point of view of the ac electrical circuit. Thus, voltage or electromotive force in the electrical system has as its analogs force in a mechanical system, and pressure in an acoustical or hydraulic system. Current in an electrical system is analogous to velocity in a mechanical system or an acoustical system, and to rate of flow in a hydraulic system. Electrical resistance has as its analogs friction in a mechanical system or a hydraulic system, and the real component of acoustic impedance in an acoustical system. These analogies can be extended to capacitance, inductance, impedance, etc.

Because of these analogies between systems, the use of generalized concepts in control problems is particularly valuable. Let us take the first step in this direction by using again the simple resistance-capacitance circuit of the previous application, but denoting the input voltage by $\theta_i(t)$ to differentiate it from the voltage available to an external load by connecting the latter across the condenser, the latter being called the output, $\theta_0(t)$. (See Fig. 2.) Then the electrical equations for these two voltages are

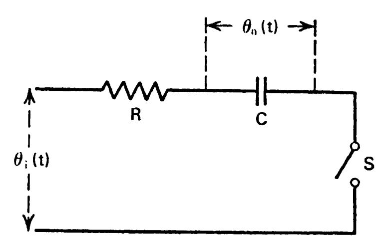


Fig. 2. Resistance-capacitance (*RC*) ac circuit showing input and output connections.

$$\theta_i(t)\ Ri(t) + \frac{1}{C}\int i(t)\, dt \tag{29}$$

$$\theta_0(t) = \frac{1}{C}\int i(t)\, dt \tag{30}$$

Assuming, as in the previous application, that there is no initial charge on C (switch open), then the Laplace transforms of $\theta_i(t)$ and $\theta_0(t)$ are

$$\mathcal{L}[\theta_i(t)] = \theta_i(s) = RI(s) + \frac{1}{sC} I(s) \tag{31}$$

$$\mathcal{L}[\theta_0(t)] = \theta_0(s) = \frac{1}{sC} I(s) \tag{32}$$

We now generalize these terms; $\theta_i(t)$ is called the driving function of the system; $\theta_0(t)$ is called the response function of the system; $\theta_i(s)$ is called the driving transform of the system; and $\theta_0(s)$ is its response transform. The ratio of the response transform to the driving transform, $\theta_0(s)/\theta_i(s)$ is called the transfer function, and is a characteristic of the system independent of its particular input or output: it is often denoted by $G(s)$.

Then we have, from the foregoing equations,

$$G(s) = \frac{\theta_0(s)}{\theta_i(s)} = \frac{\frac{1}{sC} I(s)}{RI(s) + \frac{1}{sC} I(s)} = \frac{1}{RCs + 1} \tag{33}$$

(c) *Unit step driving function.* The first Laplace transform evaluated in this section (Equation (2)) was that of the unit function $u(t)$ for the two values 1 and $t \geq 0$, and 0 for $t < 0$. Its Laplace transform was found to be $1/s$.

Therefore, if the driving function, $\theta_i(t)$, of a system is of the unit kind, as shown in Fig. 3, its driving transform is

$$\theta_i(s) = 1/s \tag{34}$$

and its response transform is, from Equation (33),

$$\theta_0(s) = G(s)\theta_i(s) = \frac{1}{RCs + 1} \cdot \frac{1}{s} + \frac{1}{s(RCs + 1)} \tag{35}$$

Then applying the partial fraction method expressed in Equations (22) and (23), we have

$$\theta_0(s) = \frac{k_1}{s} + \frac{k_2}{RCs + 1} = \frac{1}{s} + \frac{-RC}{RCs + 1} \tag{36}$$

Therefore,

$$\theta_0(s) = \frac{1}{s} - \frac{1}{s - 1/RC}$$

Then using transform pairs numbers 1 and 7 in the table,

$$\mathcal{L}^{-1}[\theta_0(s)] = \theta_0(t) = 1 - e^{-t/RC} \tag{37}$$

Figure 3 shows how the value of $\theta_0(t)$ approaches the unit value asymptotically as t increases, as is evident from the equation.

(d) *Unit ramp driving function.* The second Laplace transform evaluated in this treatment, Equation (3), was that of the unit ramp function $f(t)$ for values of $t \geq 0$. Its Laplace transform was found to be $1/s^2$.

Therefore, if a driving function $\theta_i(t)$ is of the Heaviside unit ramp kind, as shown in Fig. 4, its driving transform is

$$\theta_i(s) = 1/s^2$$

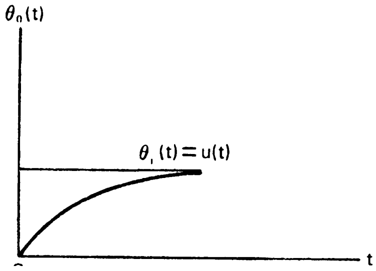


Fig. 3. Response of RC circuit to unit input.

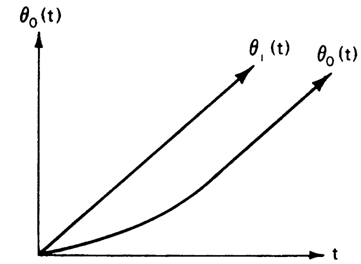


Fig. 4. Response of RC circuit to Heaviside unit ramp input.

and its response transform is, from Equation (35),

$$\theta_0(s) = \frac{1}{s^2(RCs + 1)} \tag{38}$$

Then applying the partial fraction method expressed in Equations (22) and (23), we have

$$\begin{aligned}\theta_0(s) &= \frac{k_1}{s^2} + \frac{k_2}{s} + \frac{k_3}{RC + 1}\\ &= \frac{1}{s^2} - \frac{RC}{s} + \frac{(RC)^2}{RCs + 1}\\ &= \frac{1}{s^2} - \frac{RC}{s} + \frac{RC}{s + 1/RC}\end{aligned}$$

Then by use of transform pairs numbers 1, 2, and 7 in the table, we have so

$$\mathcal{L}^{-1}\theta_0(s) = \mathcal{L}^{-1}\left[\frac{1}{s^2} - \frac{RC}{s} + \frac{RC}{s + 1/RC}\right]$$

$$\mathcal{L}^{-1}\theta_0(s) = \theta_0(t) = t - RC(1 - e^{-t/RC}) \tag{39}$$

Figure 4 shows how the value of $\theta_0(t)$ lags the unit ramp function as t increases. See **Time Constant.**

LAPLACIAN. The vector operator

$$div \text{ grad} = \nabla \cdot \nabla = \nabla^2$$

In rectangular coordinates its components are $\partial^2/\partial x^2$, $\partial^2/\partial y^2$, $\partial^2/\partial z^2$; in spherical polar coordinates

$$\frac{1}{r^2}\frac{\partial}{\partial r}\left(r^2\frac{\partial}{\partial r}\right), \frac{1}{r^2 \sin\theta}\frac{\partial}{\partial\theta}\left(\sin\theta\frac{\partial}{\partial\theta}\right), \frac{1}{r^2\sin^2\theta}\frac{\partial^2}{\partial\phi^2}$$

and in cylindrical coordinates

$$\frac{1}{\rho}\frac{\partial}{\partial\rho}\left(\rho\frac{\partial}{\partial\rho}\right), \frac{1}{\rho^2}\frac{\partial^2}{\partial\phi^2}, \frac{\partial^2}{\partial z^2}.$$

See **Del; Divergence (Mathematics);** and **Gradient (Mathematics).**

LAPPING. The process of producing an extremely accurate, highly finished surface by means of a lap, which is a block charged with abrasive. Lapping reduces the possibility of wear on close-fitting running parts or on the surfaces of measuring equipment, by reducing the minute ridges and serrations left by machining and grinding operations to a more uniform bearing surface. Lapping may be done by hand or by machine. If a part is to be hand lapped to a final accurate dimension, a lap or mating form is made from a metal somewhat softer than the part to be finished. The surface of the lap is charged with a fine abrasive, or a small amount of abrasive mixed with grease, oil, or alcohol. A flat lap has a carefully trued surface with a series of grooves in it. The lapping compound is smeared on the face of the lap and the work is rubbed over the face along an ever-changing path. The grooves in the face of the lap act as channels for any excess abrasive and oil. Very little pressure is used in order to eliminate the danger of scoring the work or stripping the lap. Hand lapping requires skill and time. The amount of material removed by lapping should not exceed 0.0002 to 0.0005 inch (0.005 to 0.013 millimeter).

LAPSE. See **Atmosphere (Earth).**

LAPWING. See **Waders, Shorebirds, and Gulls.**

LARCH TREES. Members of the family *Pinaceae* (pine family), these trees are of the genus *Larix*. The larches are different among the conifers in that they are deciduous, turning a beautiful yellow color in the fall. They prefer a lot of sunlight and do well in most soils except dry, yellow, chalky soils. The principal species, in addition to those listed on accompanying table, include:

Dahurian larch	*Larix gmelini*
Dunkeld hybrid larch	*L. eurolepsis*
Japanese larch	*L. kaempferi*
Red larch	See Japanese larch
Siberian larch	See Dahurian larch
Siberian larch (red)	*L. olgensis*
Sikkim larch	*L. griffithiana*

Forest ranger measuring circumference of an eastern larch located at Jay, Maine. (*Maine Bureau of Forestry.*)

The term *tamarack* frequently refers to the eastern larch, but is also used in connection with certain other larches, including the western larch.

The eastern larch is found through Alaska, Canada, south to Minnesota and through the northern parts of the midwestern United States, and south and eastward through West Virginia, Pennsylvania, and New York. In Canada, the tree follows the Mackenzie River northward into Yukon-Northwest Territory. The tree bark is thin, bright red-brown. The wood weighs about 38 pounds per cubic foot (609 kilograms per cubic meter). It is strong, durable, hard, and close-grained. In the spring, the flowers are of bright yellow. The needles are bright green from $\frac{1}{2}$ to 1 inch (1.3 to 2.5 centimeters) long. The cone is winged, chestnut brown, and from $\frac{1}{2}$ to $\frac{3}{4}$ inch (1.3 to 1.9 centimeters) long. The eastern larch or tamarack of the far north does not grow so large as those found in more southern climes, but it grows very straight and is ideal for telephone poles and railroad ties. However, it is not as extensively used for timber as the western larch. Other names sometimes used for the tree are American larch, black larch, and eastern hackmatack. See accompanying figure and table.

The western larch is found principally from British Columbia to Oregon, Montana, and into Idaho. It grows at elevations up to about 7000 feet (2100 meters). The bark is thick (3 to 6 inches) (7.6 to 15.2 centimeters) and is reddish-brown with deep wide ridges. The twig is stout and brittle. The leaf is pale green, 1e1/2 to 2 inches (3.8 to 5 centimeters) in length and grows in clusters of 20 to 30. The wood is heavy, durable, strong, and close-grained. The heart wood is red-brown in color; the sap wood is thin and nearly white. In the green condition, the moisture content of western larch is 58%, with a weight of 48 pounds per cubic foot (769 kilograms per cubic meter); when air-dried to 12% moisture content, the weight is 36 pounds per cubic foot (577 kilograms

Record Larches in the United States[1]

	Circumference[2]		Height		Spread		
Specimen	Inches	Centimeters	Feet	Meters	Feet	Meters	Location
European larch (1991) (*larix decidua*)	171	434	83	25.2	48	14.6	Connecticut
Subalpine larch (1986) (*Larix lyallii*)	249	632	95	28.9	77	23.5	Washington
Tamarack (Eastern larch) (1983) (*Larix laricina*)	143	363	92	28.0	31	9.5	Maine
Western larch (1980) (*larix occidentialis*)	233	592	175	53.3	37	11.3	Montana

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

per cubic meter), and 1,000 board-feet (2.36 cubic meters) weigh about 3,000 pounds (1360 kilograms). The compressive or crushing strength parallel to the grain of the green wood is 3,990 pounds per square inch (27.5 MPa); of the air-dried wood, 8,110 pounds per square inch (56 MPa). The tensile strength perpendicular to the green of green wood is 330 pounds per square inch (2.3 MPa); of the air-dried wood, 430 pounds per square inch (3 MPa). The wood is valuable for many uses, including interior finished products, furniture, and cabinetry, but its major value lies in its use for heavy-duty construction, rough timbers for mining, railroad ties, and telephone poles. The tree usually grows very straight. In some areas, the tree has been extensively cut and is now under government protection. The tree frequently is planted for ornamental purposes where the size is appropriate. Douglas fir timber frequently is shipped along with the western larch timbers. Other commercial names for the wood include mountain larch, Montana larch, and hackmatack.

As noted from the list of species, larches also occur widely in parts of Europe and Asia. The European larch is an important source of wood in Russia, and other eastern European countries.

See also **Conifers.**

LARGEMOUTH BLACK BASS. See **Sunfishes.**

LARGE NUMBERS (Law of). See **Law of Large Numbers.**

LARK (*Aves, Passeriformes*). Song birds of many species, confined to the northern hemisphere. The skylarks of Europe and Asia are the most famous members of the group because of the quality of their song. The common European species, *Alauda arvensis*, is established in Oregon and North America also has a native species, the horned lark, *Otocoris alpestris*. The meadowlark is more closely related to the blackbirds and orioles. See **Meadowlark.**

LARMOR'S THEOREM. This theorem concerns the motion which a charged particle or system of charged particles subject to a central force directed towards a common point experiences when under the influence of a small uniform magnetic field. If a coordinate system is chosen which rotates about the direction of the magnetic field with an angular velocity,

$$\omega = -\frac{q}{2mc}H$$

where ω is the angular velocity, q is electric charge in electrostatic units, m is mass, c is the velocity of light, H is magnetic field strength in electromagnetic units, then the motion in this system of coordinates is the same as the motion referred to a coordinate system fixed in space without a magnetic field. Application of this principle arises almost exclusively in the description of the motion of atoms and electrons in magnetic fields.

LARVA. An immature form of animals that undergoes metamorphosis between emergence from the egg and the attainment of adult life. The larva is often very different from the adult.

The larvae of many invertebrates of sessile habit, such as the sponges and some coelentrates, are ciliated (cilia) organisms which swim about for a time before attaching themselves to the permanent support where they are to develop. In the two phyla mentioned the larva is little more than a ciliated gastrula. It is filled with solid endoderm in the coelenterates and is called a planula.

The flukes also begin life as a ciliated larva known as a miracidium. This form gives rise to more complex larvae called rediae and these in turn produce tailed larvae called cercariae. The cercaria is transformed into the adult. In the same phylum the tapeworms hatch as 6-spined hexacanth larvae and pass through a bladder-worm stage before becoming adults.

Roundworms of many species pass through one or more larval stages in which they are worm-like but differ in habits and in some structural details from the adults.

Some of the segmented worms, mollusks, echinoderms, and chordates, hatch as complex larvae with localized zones of cilia for locomotion. The trochophore or trochosphere larva of annelids and mollusks has a ciliated alimentary tract with mouth and anus 90 degrees apart and a belt of cilia around the middle of the body. Larvae of echinoderms also have a bent alimentary tract but the cilia are arranged in one or more bands, sometimes of intricate form. These larvae bear various names: bipinnaria, auricularia, pluteus, of the starfishes, sea cucumbers and sea urchins and brittle stars, respectively, or collectively the dipleurula. The bipinnaria resembles the tornaria larva of Balanoglossus.

Among the arthropods the high development of metamorphosis is accompanied by great diversity of larval forms. In the class Crustacea these forms seem to represent previous evolutionary stages and are named after groups of the class whose adults they resemble. Among them are the nauplius, cypris, and cyclops larvae and many others. A single individual may pass through several of these stages in the course of its development. Insects present an entirely different type of metamorphosis in which the larval characteristics appear to have been acquired later in the course of evolution than those of the adult, as an adaptation to special conditions (Cenogenesis). In species with complete metamorphosis, only the first immature stage is called the larva. In this stage, butterflies and moths are called caterpillars; some of the beetles, grubs; and many flies, maggots. Species with less complex metamorphosis are called nymphs or naiads during development (in gradual and incomplete metamorphosis respectively).

See also **Metamorphosis;** and **Pupa.**

LARYNGITIS. See **Larynx.**

LARYNX. Sometimes called voice box, the larynx is a cartilaginous organ of the throat which contains the vocal cords and produces most of the sound in phonation. The larynx is shaped fundamentally like a tube, wide at the top and narrow at its lower portion. Generally, it appears somewhat like a three-cornered tube with a prominent ridge on the front side. It is made up of several pieces of firm elastic tissue (cartilage), which are held together by muscles and ligaments. The thyroid cartilage is the largest cartilage of the larynx and consists of two plates standing on end, which meet in the front of the neck and form a ridge (Adam's apple).

The interior of the larynx extends from the pharynx above to the windpipe (trachea) below. The inner tube of the larynx is divided horizontally into two parts by the projection of the muscular *vocal folds*, which contain the two *vocal cords*. The cords are really folds in the lining of the larynx rather than true cords. These cords produce the sound which is converted into speech by the movements of the mouth and tongue. When the vocal cords are tightened, the air being exhaled causes the cords to vibrate, and sounds are produced. The tighter the cords, the Higher the tone. These sounds are made into words by the tongue, teeth, and lips. The degree to which a person can tighten and relax his vocal cords determines his tone range in singing and speaking. When the vocal cords are fully relaxed, no sound is made.

Laryngitis is an inflammation of the mucous membrane of the voice box. Common microbial flora found in the larynx include *Haemophilus influenzae*, influenza virus, parainfluenza virus, adenovirus, coxsackievirus, rhinovirus, and respiratory syncytial virus. uncommonly present in the larynx are pneumococci, Group A streptococci, *Mycobacterium tuberculosis*, and *Mycoplasma pneumoniae*. Viral laryngitis in children is commonly called *croup*.

Cancer of the larynx makes up only about 2% of all human cancers. It occurs most often in men between ages 40 and 60 years, although all age groups and both sexes may be affected. The main symptom is hoarseness. Any voice change that persists more than two weeks should be investigated by a physician. There may be a tickling sensation in the throat, or simply a discomfort of the throat. There may be difficulty in swallowing and pain on speaking. Other ensuing developments may be a cough, shortness of breath, wheezing, and halitosis. Lymph nodes in the neck may become enlarged. Occasionally the patient may expectorate blood.

In the United States, new cases of laryngeal cancer number about 12,000 per year, of which 3,700 are ultimately fatal. Epidemiologic studies from many parts of the world have demonstrated an increase in the relative risk of laryngeal cancer in smokers as compared with nonsmokers, ranging from 2.0 to 27.5, with a strong dose-response relation. Laryngeal cancer, in accordance with recent statistics on smoking, may be increasing among women and somewhat decreasing among men. As with lung cancer, the time necessary after smoking cessation for the risk levels for laryngeal cancer to approach those in nonsmokers is about 10 to 15 years.

Radiation and/or surgical therapy may be used in connection with laryngeal cancers. Surgery is usually indicated if the disease involves a large portion of the larynx. When the larynx is removed, the patient's windpipe is attached to the skin of the neck. Thus, the patient no longer breathes through the nose, but rather through a hole in the neck. The patient must take precaution to avoid breathing in foreign matter because of the lack of protection normally afforded by the nose. As soon as healing permits, the patient will normally commence speech lessons to learn to speak by application of one of two natural procedures, the esophageal method or the pharyngeal method.

Where sustained effort to learn either of these methods does not lead to success, the patient may use an artificial larynx, a device which is held against the side of the throat. When activated, the device transmits vibrations to the pharynx and mouth, providing intelligible but unnatural sounds.

Alternate Treatment. As pointed out in a 1991 study by the Department of Veterans Affairs, a Laryngeal Cancer Study Group reports, "Because laryngectomy results in substantial functional morbidity, including the loss of the natural voice, alterations in deglutition (swallowing), and the creation of a permanent tracheostoma (hole) in the neck, alternative forms of treatment have been developed. In selected patients with moderately advanced cancers, partial laryngeal resections that spare vocal function or primary radiation therapy achieve survival rates comparable to those obtained with total laryngectomy, permitting preservation of the laryns in 40 to 70 percent of patients. For patients with more advanced disease, however, treatment by radiation alone, with salvage by surgery, if necessary, results in lower survival rates."

The Veterans Affairs Group made a study of 352 patients, who were randomly assigned to receive either three cycles of chemotherapy (cisplatin and fluorouracil) and radiation therapy or surgery and radiation therapy. After two cycles of chemotherapy, the clinical tumor response was complete in 31% of patients and partial in 54% of patients. The report concludes: "These preliminary results suggest a new role for chemotherapy in patients with advanced laryngeal cancer and indicate that a treatment strategy involving induction chemotherapy and definitive radiation therapy can be effective in preserving the larynx in a high percentage of patients, without compromising overall survival."

Injuries. Freak, uncommon accidents can affect the functioning of the larynx and related organs. In 1991, T. C. James, T. C. Li (Mayo Clinic), and D. Gunderson (Park Nicollet Medical Center, Minneapolis) report that a young man suffering from asthma awoke in the middle of the night and reached for a metered-dose inhaler that he had been using. Unfortunately, still groggy from sleep, the man did not remove the cap of the inhaler and aspirated it into the larynx. He suffered some vocal cord damage, but recovered within a period of about 2 months. In another instance, a middle-aged man had acute respiratory distress after using an inhaler. A laryngoscopy revealed that he had inhaled a loose coin that he kept in his pocket along with the inhaler. Radiology revealed a coin in the right mainstem bronchus. This was removed, and the patient recovered without sequelae. This latter case was reported by Carol Shultz (Henry Ford Hospital) and associates in 1991.

Obviously, persons who are given inhalers should be instructed to check the mouthpiece before each use and to cap the device after each use.

Myasthenic laryngitis signifies a weakness and exhaustion of the muscles in the larynx resulting from overuse of the voice box. Resting the voice part of each day is important, as well as refraining from shouting, talking in a loud voice, and all unnecessary uses of the voice. Usually the voice returns to normal when these precautions are observed.

Additional Reading

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Schultz, C. H., Hargarten, S. W., and J. Babbitt: "Inhalation of a Coin and a Capsule from Metered-Dose Inhalers," *N. Eng. J. Med.*, 431 (August 8, 1991).

Veterans Affairs Laryngeal Cancer Study Group: "Induction Chemotherapy Plus Radiation Compared with Surgery Plus Radiation in Patients with Advanced Laryngeal Cancer," *N. Eng. J. Med.*, 1685 (June 13, 1991).

LASER. An acronym for *light amplification by stimulated emission of radiation.* The device is identical in theory of operation to the maser except that it operates at frequencies in the optical region of the electro-magnetic spectrum, rather than in the microwave. See **Maser.** By common usage, these devices are all called lasers, although more precise terminology would aptly use such terms as ultraviolent maser, optical maser, infrared maser, etc. Although the original microwave maser offers an extremely stable frequency source, its main use has been as an amplifier with very low noise output. In contrast, one of the principal attributes of the laser is its ability to produce a single frequency at high intensity in the optical region. Not only may the output be a single monochromative wave, but the wave may be coherent, or in phase, over the whole face of the radiator. In this mode of operation, the laser is an oscillator whose output depends upon the selective amplification of one of the single-frequency modes of the resonant cavity containing the active laser medium.

The decade of the 1980s represented the period when laser technology received profound acceptance, to the point where lasers, like semiconductors and digital computers before them, are now considered important components (hardware) of larger systems. The development of laser technology has continued apace during the early 1990s, with numerous new and practical applications occurring at an accelerated rate. lasers now range from microlasers that measure but a few millionths of a meter, for use in optical communications and information processing, to the building-sized x-ray research laser that can deliver 100,000 joules of energy in less than one-billionth of a second—that is, a 10^{14}-watt pulse. see Fig. 1.

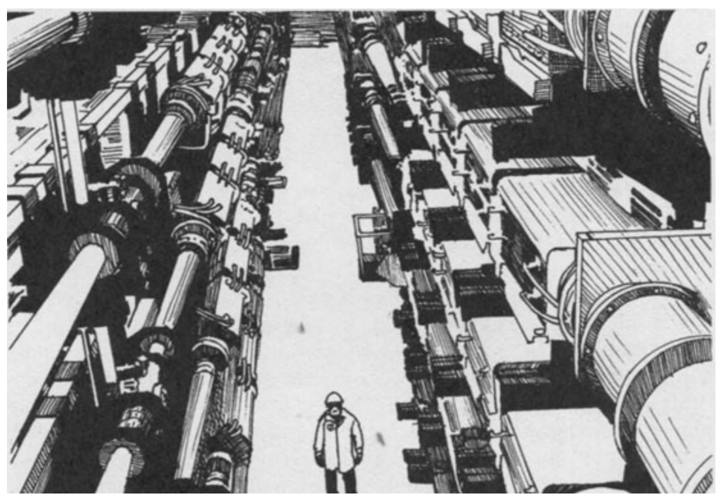

Fig. 1. A very powerful optical laser installed at the Lawrence Livermore National Laboratory. Known as NOVA, two of its ten beams serve as the energy source for x-ray lasers. The latter need about 1000 times the pump energy of optical lasers, delivered about 10,000 times faster. Note size of person in the forefront of this sketch.

Historical Perspective

The fundamental theoretical concepts of the maser and laser date back many years to the early workers in quantum mechanics who appreciated that an incident electromagnetic beam of an appropriate reasonant frequency passing through a medium might stimulate molecules in an upper quantum energy state to return to a lower quantum energy state and thus reinforce the primary beam by negative absorption. As early as 1940, Fabrikant (Russia) suggested that experiments be made to prove negative absorption. Fabrikant was the first scientist to introduce the term "collisions of the second kind" (later to prove of importance in laser patent litigation) and by which he meant a collision in which some of the kinetic energy of motion of colliding particles is converted to internal energy (or a change in energy state of at least one of the colliding particles).

It is also interesting to note that as early as 1950, researchers Lamb and Retherford (Columbia University), observed that, if an upper quantum energy state could be caused to be more highly populated (as compared with a lower quantum energy state), the result would be a net induced emission to an incident beam. They further indicated that such a population inversion would probably occur between the $2p$ and $2s$ levels in hydrogen. There shortly followed experiments by Purcell and Pound (Harvard University) who used magnetic techniques to invert the population of a pair of nuclear spin states in lithium fluoride and thus were the first researchers known to directly observe negative absorption of an applied pulse, a phenomenon which they called negative temperature.

Townes (Columbia University) is génerally accredited with first recognizing that stimulated emission could be utilized in the making of practical hardware. In 1951, Townes described an approach wherein an ammonia beam would be divided into two portions along the lines of experimentation carried out in Germany, wherein a quadrupolar focusing technique was used for separating a beam of molecules into two portions, one of which contained molecules predominantly in the upper of two energy states. Townes proposed to pass the high-energy portion through a cavity resonant at the frequency corresponding to the energy separation of the two states and so reported this proposal. Instead of a millimeter wave generator, a microwave oscillator was used. The latter was given the name *maser*. This led to the award of U.S. Patent No. 2,879,439, which covered the use of stimulated emission for the amplification and/ or generation of oscillatory electromagnetic energy. This patent subsequently was licensed to laser manufacturers.

Following the development of the microwave maser, Schawlow (AT&T Bell Laboratories) and Townes in 1958 proposed that optical maser action could be obtained by placing an active medium in an optical cavity. The medium would be a gas or a solid which was excited electrically or by light in such a manner that any optical wave present would be amplified as it moved through the material. This work led to the ultimate awarding of the basic laser patent in March 1960 (U.S. Patent No. 2,929,922). During the 1960s, laser research was carried on at a rapid pace at AT&T Bell Laboratories, among others. Considering the work at AT&T Bell Laboratories, in 1960 a laser capable of emitting a continuous beam of coherent light (using helium-neon gas) was developed; in 1961, the continuous-wave solid-state laser (neodynmium-doped calcium tungstate) was developed. As a refinement of the helium-neon laser, in 1962 the basic visible light helium-neon laser was developed, of which several hundred thousand are in use as of the early 1980s. In 1964, the carbon dioxide laser (highest continuous-wave power output system known to date) was developed. Other developments during 1964 included the neodymium-doped yttrium aluminum garnet laser, the continuously operating argon ion laser, the tunable optical parametric Oscillator, and the synchronous mode-locking technique, a basic means for generating short and ultrashort pulses. In 1967, the continuous wave helium-cadmium laser (utilizing the Penning ionization effect for high efficiency) was developed. These lasers find use in high-speed graphics and biological and medical applications. In 1969, the magnetically tunable spin-flip Raman infrared laser, used in high-resolution spectroscopy, as well as in pollution detection in both the atmosphere and the stratosphere, was developed. Laser developments continued and, in 1970, semiconductor heterostructure lasers capable of continuous operation at room temperature were introduced. The distributed feedback laser, a mirror-free laser structure compatible with integrated optics, was introduced, in 1971. These were followed by the tunable, continuous-wave color-center laser (1973), techniques for creating optical pulses of less than one-trillionth second duration (1974), and, in the late 1970s, long-life semiconductor lasers for lightwave communications.

One of the first uses of a laser in an astronomical setting was the placement of a laser retroreflector in the Sea of Tranquility on the moon during the landing of Apollo 11 on July 20, 1969. The retroreflector is passive, requiring no power. Two additional reflectors were mounted, to form a triangular arrangement, by Apollo 14 and 15 at later dates. The information continues to be checked periodically, when the atmosphere is clear by the McDonald Observatory in Texas, the Lure Observatory in Hawaii, and the Calhern Observatory in southern France. The reflectors are pulsed with 259-million joule neodymium-YAG (yttrium-aluminum-garnet) lasers. Because the retroreflectors were left on featureless land (not mountainous), telescope pointing is described as being akin to targeting a moving dime with a rifle 2 miles (3.7 km) away. The ranging experiments have contributed much to the knowledge of the moon's orbits and other geometric data. It has been noted that the moon recedes from Earth at a rate of about 1.5 inches (3.8 cm) per year. Minor variations in the moon's rotation also have been noted. These measurements indicate variations as small as 1/1000 of an arc-second over the course of a year. See also **Moon (Earth's).**

Laser Principles

The basic concept of the laser may be described in general terms as follows. Optical maser action can be obtained by placing an active medium in an optical cavity. The medium may be a gas, solid, or liquid which can be excited electrically, by light, chemically, or thermally in such a manner that any optical wave present will be amplified as it moves through the material. One of the first cavities proposed, for example, is a Fabry-Perot resonator—two plane, parallel reflecting plates with a small transmission through which the radiation can escape. Upon excitation of the material, light is emitted with a band of frequencies determined by the particular material. In addition, the direction of emission is nominally random. In the presence of the cavity, some of the waves escape after several back-and-forth reflections from the parallel plates, "walking off" the edge of the reflectors, so to speak. Those waves which travel normal to the walls remain in the cavity and are amplified, provided that they reinforce each other after each round-trip reflection at the two surfaces. this reinforcement or resonance is only satisfied if the spacing of the plates is an integral multiple of one-half the wavelength in the medium. thus, after a short time, only that frequency which satisfies the resonant condition and those waves traveling normal to the reflector will build up an appreciable intensity. the resultant light which is partially transmitted through one of the reflectors will thus be a single frequency, or several discrete frequencies if there is more than one cavity resonance within the band of frequencies emitted by the laser material. In addition, the wave front will be in phase across the surface of the reflector since waves striking the surface at normal incidence are amplified most strongly. The resultant beam will then be diffraction limited, i.e., the beam will spread by an angle in radians given approximately by the ratio of the wavelength to the diameter of the beam. In actual practice, single-mode operation is obtained only under special conditions. Several frequency modes may be present because of the multiple resonances of the cavity and numerous "off-axis" modes may be found which correspond to resonant waves which travel at small angles from the normal to reflectors. These waves "walk off" so slowly that they still are amplified appreciably. Refinements of the simple cavity may consist of concave reflectors which decrease the diffraction losses, or several parallel reflectors which limit the oscillation to a frequency common to each pair in the set, among other approaches.

Basic Requirements of Lasing Media. The key to successful laser operation is the active medium which amplifies the wave. Qualitatively, a material which fluoresces or exhibits luminescence is an obvious candidate. In fluorescence, electrons are excited to an upper-energy state by short-wavelength light, such as ultra-violet, while luminescence is produced by passing an electron current through the medium, such as in a gaseous discharge. In either process, stimulated emission can occur only if more electrons are produced in the upper-energy state than in the lower or terminal state for the radiating transition. In this case, an incident photon will stimulate further transitions and amplification will result. If the final state were more heavily populated, then the photon would cause more upward or absorbing transitions and the net effect would be absorption.

Laser source requirements vary widely among specific applications. Practical systems require a large range of wavelengths, output powers, spatial and temporal beam characteristics, among other features. In many applications, the following factors apply: (1) an optimum wavelength exists; (2) a specific minimum power level is required; (3) capital and operating costs of the laser must be minimized; (4) size and weight constraints must be met; (5) the laser should be capable of operating for extended periods with little maintenance; (6) the laser output should have a specific temporal and spatial characteristics; and (7) operator safety must be considered in design and use of the overall laser system.

Wavelength characteristics of available laser types are given in Table 1.

TABLE 1. WAVELENGTH RANGE OF SOME LASERS

Type of Laser	Wavelength (Micrometers)
Solid state	1.06
Ion	0.514, 0.488
Carbon dioxide	10.6
Diode	0.65–1.8
Dye	Visible (tunable)
Helium-neon	0.63
Rare gas halide	0.35, 0.25, 0.19
Helium-cadmium	0.422

Types of Lasers

From the foregoing list of media requirements, it is obvious that a majority of materials are not candidates for making an effective laser. When the laser was first conceived, the possibility of finding numerous materials, as has turned out to be the case, initially was not envisioned by most researchers in the field.

Early Ruby Laser. It is recorded that the first optical laser was demonstrated by Maiman (Hughes Aircraft Research Laboratories) in 1960. Maiman used a ruby, which is a single crystal of aluminum oxide doped with chromium impurities. By applying semitransparent reflective coatings on the ends of a rod about 2 inches (5 centimeters) long, Maiman made the cavity and the crystal an integral unit. See Fig. 2. Exposure to an intense exciting light from a xenon flashtube was found to invert the population between the red-emitting level and the ground or lowest-energy state of the electrons. The result was a burst of intense red light emanating in a beam through the end reflectors. This was a powerful laser.

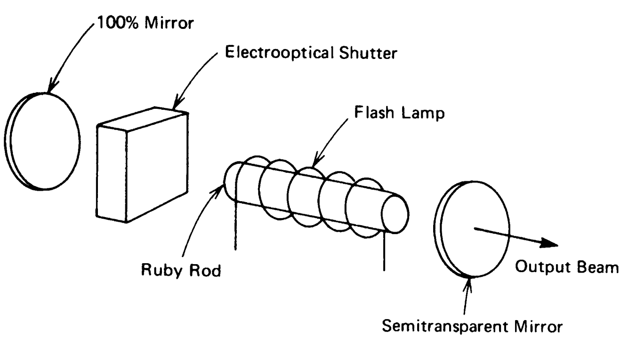

Fig. 2. Schematic diagram of early giant pulsed ruby laser.

Because of off-axis modes and multiple resonances, the output is not a single frequency, single plane-wave mode, but generally consists of the order of 100 separate modes. The beam is still quite narrow, being of the order of 1 milliradian or 0.05 degree. As a comparison with conventional light sources, the energy radiated from 1 square centimeter of the brightest flash lamp is less than 10 kW and is distributed over the entire visible spectrum. In addition, the radiation is incoherent and is spread out uniformly in all angles from the source. Thus, the directivity and spectral purity of the laser source are many orders of magnitude superior to that of an incandescent source. The ruby laser suffers from low efficiency, about 1%, and except with elaborate cooling systems, only operates on a pulsed basis. The ruby laser was first reported in the literature (Maiman, 1960) and, in 1967, Mainman was awarded U.S. Patent No. 3,353,115 for the optically pumped ruby laser.

Other crystalline or glass systems with impurity ions have been developed, which yield wavelengths from the ultraviolet to approximately 3 micrometers wavelength in the infrared. Some, such as neodymium-doped yttrium aluminum garnet (YAG), operate in a continuous mode at the one-watt level, while peak powers have reached values as high as 10^{14} W, or greater, in pulses of the order of 10^{-12} second. These ultrahigh powers are obtained in neodymium-doped glass systems, using several stages of amplification and novel pulse-forming techniques.

Gas Lasers. In 1961, Javian, Bennet, and Herriott demonstrated laser action in a gaseous discharge of helium and neon. The parallel-plate reflector cavity was used, but with much greater spacing. Later, concave mirrors were used to decrease the loss of energy out the sides of the cavity. See Fig. 3. The gas laser operated continuously and delivered power up to about one watt. Pulsing the gas discharge yielded peak power as high as 100 watts. The first laser radiated at 1.15 micrometers in the infrared, while further development with different gases yielded output from the ultraviolet to 330 micrometers or 0.33 millimeters in the far infrared. In contrast to the ruby laser, the gaseous laser beam may be diffraction limited and the frequency is pure, i.e., oscillation may be limited to one mode. By careful design, the frequency may be stabilized to within a few thousand cycles per second, or approximately one part in 10^{13} or better. Although the original gas laser utilized electrical excitation of electronic transitions, later versions use vibrational transitions in molecules, such as carbon dioxide, and the excitation may be electrical, chemical, or thermal. The helium-neon laser is commonly used in supermarket bar-code scanners.

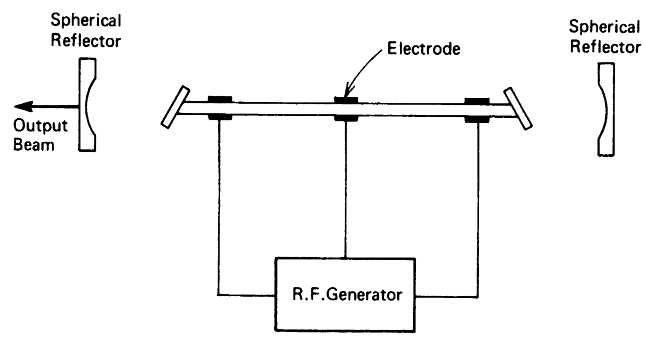

Fig. 3. Schematic diagram of early helium-neon laser with external spherical reflectors. Curvature of reflectors is exaggerated.

In the chemical laser, atomic species such as hydrogen and fluorine can be reacted to produce molecules in an excited vibrational state which, in turn, yields amplification or oscillation. Electrically excited lasers, particularly those using carbon dioxide at 10 micrometers, can operate at atmospheric pressure, using spark discharges or pre-ionization voltages in the 100-kV range. The high pressure and the powerful electrical excitation result in peak powers in the 10 to 100 MW region. For continuous laser operation, the gas may be circulated rapidly to avoid excessive heating.

In a gas dynamic laser, an appropriate fuel is burned to produce carbon dioxide and nitrogen at high temperature and pressure. When released through a nozzle into the optical resonator region, the gas cools rapidly in terms of its kinetic or translational energy, but the population of the vibrational energy levels of the carbon dioxide molecules becomes inverted since the lower level of the laser transition relaxes much more rapidly. In addition, the vibrationally excited nitrogen molecules are in near resonance with the upper laser state of the carbon dioxide and transfer energy with high efficiency to maintain the inversion. Lasers of this type are capable of producing continuous power at a relatively high level.

Semiconductor Laser. In the semiconductor laser, a solid (semiconductor) material is used. The electron current flowing across a junction between *p*- aNd *n*-type material produces extra electrons in the conduction band. These radiate upon making a transition back to the valence-band or lower-energy states. If the junction current is large enough, there will be more electrons near the edge of the conduction band than there are at the edge of the valence band and a population inversion may occur. To utilize this effect, the semiconductor crystal is polished with two parallel faces perpendicular to the junction plane. The amplified waves may then propagate along the plane of the junction and are reflected back and forth at the surfaces. See Fig. 4.

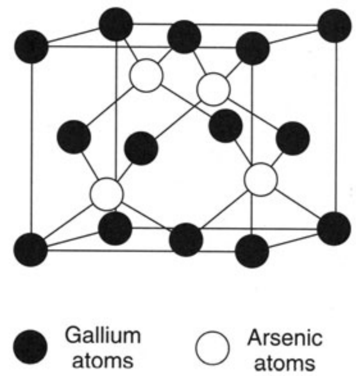

Fig. 5. Schematic diagram of a gallium arsenide (GaAs) crystal, which is commonly referred to as a zincblende structure. The structure consists of a face-centered cubic lattice of gallium atoms with arsenic atoms positioned on the body diagonals. The arsenic atoms also lie on a face-centered cubic lattice displaced relative to the fallium lattice by one-fourth the body diagonal of the cube.

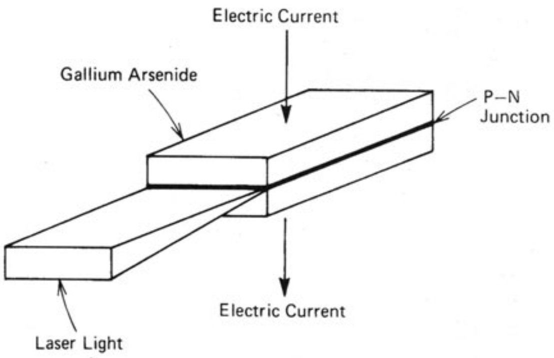

Fig. 4. Schematic diagram of early gallium arsenide laser.

A major advantage of the gallium arsenide (GaAs) laser is that it has the electron distribution of a semiconductor. The main difference between electrons in semiconductors and electrons in other laser media is that in semiconductors all of the electrons occupy and thus share the entire crystal volume. Although all semiconductors possess this property, not all of them can be used as lasers. See Fig. 5.

The gain in the material is high enough so that the reflection at the semiconductor-air interface is sufficient to produce oscillation without special reflective coatings. The first such device used gallium arsenide and radiated at 8400 angstroms, or just beyond the visible region in the infrared. The efficiency of this laser, developed in 1962, was high (about 40%) and the power source was low-voltage direct current.

The compactness and efficiency of the semiconductor laser make it particularly attractive for systems use. Other substances used have included indium arsenide, indium phosphide, indium antimonide, and alloys such as gallium-arsenide-phosphide. These lasers may be tuned over several percent of their normal frequency of operation by varying the current flow through the device. The tuning results from the variations in temperature with current which, in turn, changes the index of refraction and the resultant resonant frequency of the cavity.

Free Electron Laser. Lasers of this type depart markedly from those lasers that are considered conventional or traditional. Free electron lasers employ an electron beam and a magnetic field. A "free" electron may be defined as an electron that is not bound into atoms or molecules. Traditional lasers use bound electrons. Thus, the conventional laser is limited to producing light (radiation) that is consistent with those frequencies that are specific to the vibration of a given atom or molecule. Carbon dioxide and helium lasers, for example, are so inhibited.

Free electrons are caused to vibrate by passing them through an alternating magnetic field. By changing (tuning) the apparatus, a broader range of frequencies is obtainable. Coherent radiation can range from the far-infrared to the far-ultraviolet regions of the spectrum. See Fig. 6.

Currently, free electron lasers are large and costly, but are ideal for certain kinds of research. Developers believe that ultimately the free electron laser will find numerous applications beyond research. Kim and Sessler (reference listed) suggest that applications may include surgery, fixing polymers, pharmaceutical manufacture, and lithography. It is predicted that the free electron laser will continue to expand in research usage, including condensed matter studies, nonlinear plasma studies, nonlinear quantum electrodynamics, nonlinear optics, and nonlinear microwaves, as well as microscopy, DNA studies, and cell response research in biology. One study is in progress to determine the feasibility of adapting the free electron laser to perform precision radar measurements in space. Researchers will study the potential of the laser as a compact space radar transmitter for discriminating objects in space, as would be required, for example, in connection with the Strategic Defense Initiative (SDI) program. The program will take advantage of the electron laser's inherent tunability, high power and efficiency, and ability to operate in frequency bands of 100 GHz and higher. The program's ultimate goal is a space-based, multiband, adaptive laser capable of operating efficiently at randomly chosen, stable frequencies.

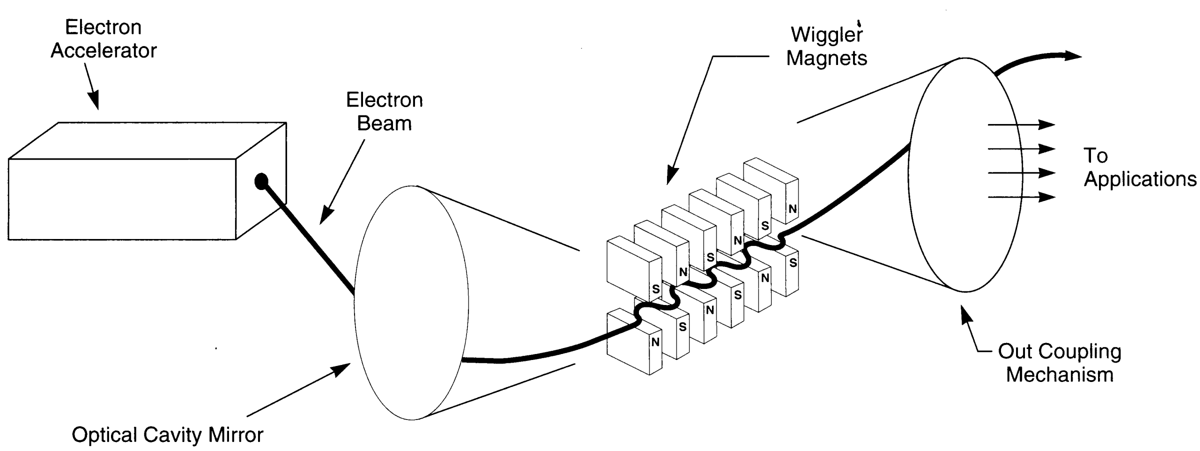

Fig. 6. Schematic diagram of a free-electron laser. Beam of accelerated electrons passes through a field of alternating magnetism (wiggler magnets). The coherent light is generated and contained in an optical cavity defined by mirrors. (*Kim and Sessler.*)

Researchers have found that electrostatic accelerators are well suited for the far-infrared spectrum. An early device of this type was built at the University of California at Santa Barbara for the main objective of free-electron research and studies in solid-state physics and biophysics. The accelerator has an operating range of 2–6 MV, corresponding to wavelengths in the range from 100–800 micrometers. Pulse duration is from 3–30 microseconds.

Solid-State Laser Development. Over the years, the progress of solid-state lasers has depended heavily on the improvement of old and the establishment of new pump sources. The helical lamp used by the early ruby laser was replaced by the linear flash lamp and arc discharge lamps. The next step was that of using a diode laser to pump another solid-state laser. This latter approach is advantageous because the diode laser emits optical radiation into a narrow spectral band. If the emission of the wavelength of the diode laser lies within the absorption band of the ion-doped solid-state laser medium, diode laser optical pumping can be very efficient and accompanied by little excess heat generation. By contrast, flash lamp pumping is limited by the broad emission spectrum and excess heat production.

As pointed out by R. L. Byer (Stanford University), "The diode laser is essentially a continuous wave device with low energy storage capability, whereas the solid-state laser can store energy in the long-lived metastable ion levels." Stored energy can be extracted by Q-switching (rapid switching) to provide peak power levels that are orders of magnitude greater than obtainable from the diode laser per se. Important, too, is the fact that a solid-state laser can collect output from several diode lasers and thus furnish greater average power than is obtainable from a single diode laser. Furthermore, the line width of the diode laser-pumped solid-state laser is many times less than that of the diode laser source. Finally, the solid-state laser source emits optical radiation in a diffraction-limited spiral beam that is easily focused into a fiber or small space.

As early as 1982, a diode laser-pumped miniature Nd:YAG laser with a linewidth of less than 10 kHz was demonstrated. The research in this area continued apace at Standford University and by a number of commercial electronics firms, with emphasis placed on the development of three-level lasers, Q-switched and mode-locked operation, single-frequency operation (monolithic nonplanar ring oscillator), visible radiation by harmonic generation, and array-pumped solid-state lasers. See Fig. 7.

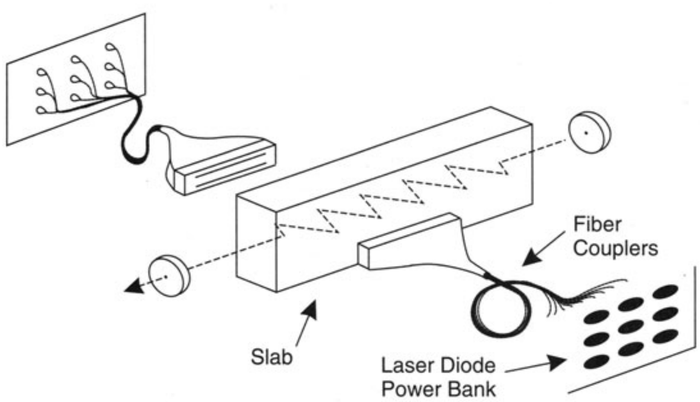

Fig. 7. Diagram of high-average power slab laser oscillator pumped by an array of diode lasers. Such an arrangement offers lower cost, ease of power scaling, and long-term reliability. (*After Byer.*)

Laser Microminiaturization—Target Optical Computer. For many years, scientists have accepted the fact that optical computers would not become a reality until optical components of micro size and exceptional performance equivalent to the already existing electronic switches and circuits could be developed. Thus, the optical computer became a major driving force toward the development of optical components. The problem was extraordinarily complex because size reduction in terms of several orders of magnitude were mandatory.

As early as 1989, researchers at AT&T Bell Laboratories fabricated more than a million micron-size lasers (microlasers) on a single semiconductor chip, about 7 mm wide by 8 mm long. Individual devices ranged in size from 1 to 5 microns. Thus, these devices were two orders of magnitude smaller than conventional diode lasers. Researchers feel that it may be feasible to manufacture such devices that measure only between $\frac{1}{2}$ and $\frac{1}{4}$ micron. Traditional devices that measure a few microns wide by several hundred microns long have been well established for use in compact-disk players and fiber-optic communications. Thus, the *much, much smaller* devices coming out of research comprise a major step toward achieving the needs of optical computing. Much competitive research continues during the early 1990s because the market demand for microminiature lasers can be immense.

As pointed out by Jewell and associates (Bellcore), "The principles of operation underlying a diode laser are the same as those for any laser. Atoms in a part of the laser called the *amplifying medium*—typically a solid, liquid, or gas—are pumped, or energized, either electrically or with a source of electromagnetic radiation. When a light wave of a specific wavelength traveling through the amplifying medium encounters a pumped atom, it can induce the atom to release its energy in the form of a light wave at the same wavelength. The process is coherent, which is to say that the crests and troughs of the waves match up, and the intensity of the light increases. Mirrors on each end of the amplifying medium form a cavity, and they force the light to bounce back and forth many times through the medium, maximizing the increase in intensity."

The differences in construction of a microlaser from a conventional gas laser or conventional diode laser are shown in Fig. 8.

To date, manufacture of the very tiny microlasers depend upon critical production techniques. For example, molecular beam epitaxy allows the basic material of each laser to be built up from layers of semiconducting materials. A typical microlaser may comprise 500 or more individual layers.

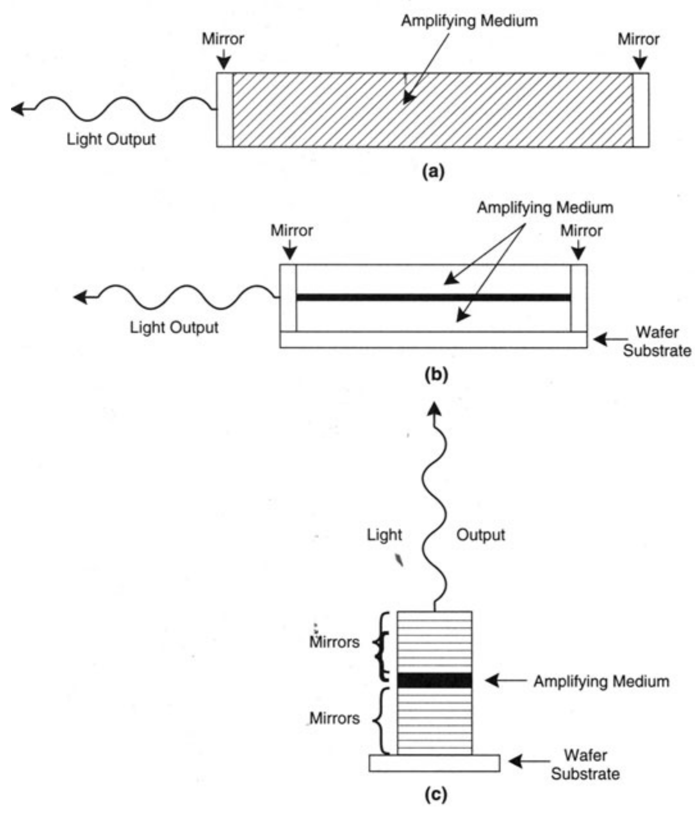

Fig. 8. Comparison of constructional configurations of (a) conventional helium-neon laser, (b) traditional diode laser, and (c) microlaser. Several orders of magnitude in size difference cannot be depicted here. The helium-neon laser ranges from 100 to 1,000 times longer than a traditional diode laser. The latter is some 100 times longer than a microlaser. (*After Jewell, Harbison, and Scherer.*)

It was reported in late 1991 that a semiconductor laser that emits a blue-green light had been developed (3M Company). Manufacturers of optical disks and other consumer electronic devices have been seeking a laser that would produce light in this range of the spectrum. The detailed development procedures required are beyond the scope of this article. However, it has been reported that the layered device is comprised of a gold electrode, a *p*-type zinc selenide, a quantum well (cadium zinc selenide), an *n*-type zinc selenide, a substrate (gallium arsenide), and an indium electrode.

Liquid Lasers. Lasers operating in liquid media utilize rare earth ions in such organic hosts as chelates. Laser action is obtained in liquids using a flash tube or another laser as the pump. Early versions used rare earths in an organic liquid, while organic dyes, introduced somewhat later, were found to be more efficient, but required a separate laser for the exciting radiation. The dye laser has the special attraction that one laser may be tuned over a significant fraction of the visible spectrum by using a reflection grating as one of the cavity mirrors.

Another type of liquid laser utilizes a different principle and depends upon stimulated Raman scattering. Raman laser action was discovered by Woodbury in 1962, using a ruby laser and nitrobenzene. Here the laser excites the nitrobenzene, which in turn shows amplification at a frequency displaced from the ruby line by the vibrational frequency of the molecule. There is not true inverted population in this case. The incident photon is scattered by the molecule which absorbs an amount of energy determined by its vibrational energy. The molecule is left in an excited state and the scattered photon is frequency-shifted by the energy loss. The process may be stimulated inasmuch as the rate at which the scattered photons are produced is proportional to the number of photons already present in the cavity at the scattering wavelength. As in the normal stimulated emission case, the frequency and phase of the output wave are identical with the wave which stimulates the scattering.

The Raman laser normally operates using the Stokes line, or the wavelength corresponding to the loss of one vibrational quantum. Other modes of operation utilize the second or third Stokes lines corresponding to double or triple vibrational absorptions. Similarly, higher-order effects in the medium may produce a series of anti-Stokes lines which correspond to vibrational energy being added to the initial energy of the photons from the driving laser. The wavelength range of Raman lasers using different liquids is from the visible to the near-infrared.

X-Ray and Other Very High-Power Lasers. Almost since the inception of laser technology, the laser has been of interest to the military in connection with a variety of applications—weapons, radars, illuminators, rangers, etc. Among the highly sophisticated laser applications long envisioned by the military is the use of a laser to propel a spacecraft. In early concepts, a laser beam produced at ground level would vaporize an appropriate fuel, which could be water, and the supersonic jet caused by vaporization would be sufficient to place the vehicle into orbit without any chemical energy being expended by the craft itself. Within the last few years, a major interest of the military has been directed toward the development of an x-ray laser for possible application in connection with the SDI (Strategic Defense Initiative) program. There are, of course, also strictly scientific interests in tunable coherent x-rays.

The continuation of these programs and the future release of technical details will be determined by the Clinton Administration (1993) and the U. S. Congress pertaining to developments within the United States.

It is generally understood that for a laser to serve as a weapon it must have appropriate wavelength and brightness characteristics. The wavelengths produced by most lasers are absorbed by the atmosphere. Laser wavelengths between 0.3 and 1 micrometer are generally the most easily transmitted. Investigators have estimated that if a laser firing over a 3000 km engagement distance is to burn through a missile skin in 1 second, it must deliver 10,000 joules of energy per centimeter. (These requirements correspond to a brightness of 10^{21} watts per steradian, or unit of solid angle.) It is further estimated that a laser having the desired wavelength and brightness would require a beam power of approximately 100 MW. By comparison, a typical nuclear power plant has an output of about 1000 MW.

European Superlaser. In mid-1990, five European countries agreed to fund ($200 to $500 million) construction of a European High Performance Laser Facility. Sponsored by France, Germany, Italy, Spain, and the United Kingdom, the new laser will be three to four times more powerful than the Lawrence Livermore National Laboratory's NOVA, currently the world leader. The program will progress in two principal stages, commencing with two intermediate power lasers—one neodymium glass and one KrF laser, built side by side. The major goal is to free European laser scientists from the dependence on high-powered machines in the United States and Japan. A superlaser of this type can be used to investigate some fundmental problems of physics. The intense pulse of the proposed laser would create conditions even hotter than the core of a burning star.

As of the late 1980s and early 1990s, four kinds of laser have been under development. (1) *Chemical lasers* utilize chemical reactions between two gases to generate radiation. This technology is probably the most mature of the four kinds of lasers. It has been reported that the brightest of the chemical lasers is the MIRACL (mid-infrared advanced chemical laser), which in a demonstration at the White Sands Missile Test Range (New Mexico) destroyed a mock-up of a missile standing about a half-mile distant from the laser. The MIRACL was estimated to have a brightness of about 10^{17} per steradian, short of the SDI goal by a factor of about 10,000. (2) The *Excimer* (meaning excited dimer) consists of an unstable compound made up of two molecules. An electric discharge excites the molecules to form the dimer and, and in breaking down, the dimer emits radiation. In some way, this radiation triggers a cascade of reactions that produce a laser beam. Radiation in the beam is estimated between 0.2 and 0.4 micrometer. A krypton-fluoride laser, tested at the Los Alamos National Laboratory, is estimated to operate at a wavelength of about 0.25 micrometer, delivering 10,000 joules of energy in a 380 nanosecond pulse. (One nanosecond = one billionth of a second.) It is reported that while the energy produced meets SDI goals, the pulse duration is off by a factor of about 3 million. (3) In the *free-electron laser*, a beam of electrons passes by a series of so-called wiggler magnets, which cause the electrons to vibrate and emit radiation. The wavelength can be tuned, theoretically, to any value between about 0.1 and 20 micrometers. The smaller the wavelength, the greater the energy. It is reported that a free-electron laser operated at wavelengths down to 10 micrometers in tests at Los Alamos. Current research targets a 1-micrometer radiation of 100 microsecond pulses, containing 30 kW of power. Thus far, both excimer and free-electron lasers are relatively poor converters of electric energy into beam energy, requiring massive power supplies—hence difficulties in locating the needed equipment in space. See prior description of free-electron laser in this article. (4) The *x-ray laser* also appears to be plagued with heavy and costly support equipment. Essentially, the x-ray laser consists of a nuclear explosive surrounded by an array (cylindrical configuration) of metal fibers. The emission of x-rays during the nuclear explosion stimulates the emission of a beam of x-rays from the fibers. This occurs within a microsecond prior to immolation of the device per se. For obvious reasons, further details remain sparse. It has been reported that to make an effective weapon, a particle beam would require energy of 250 MeV. If one assumes an acceleration gradient of 10 MeV per meter, it follows that the structure must be 25 meters long. When accounting is made of the mass of the power supply and its fuel, the weight of the weapon is found to be 50 to 100 tons. (Current typical payloads weigh a comparatively few tons.) Much additional work of a guarded nature continues.

Soft X-Ray Laser. A modern 1- to 2-billion-eV synchrotron radiation facility, based on high-brightness-electron beams and magnetic undulators, would generate coherent, laserlike soft x-rays of wavelengths as short as 10 angstroms. As reported by Attwood, Halbech, and Kim (see reference), this radiation would be broadly tunable and subject to full polarization control. Radiation with these properties could be used for phase- and element-sensitive microprobing of biological assemblies and material interfaces as well as research on the production of electronic microstructures with features smaller than 1000 angstroms. These short-wavelength capabilities, which extend to the K-absorption edges of C, N, and O, are neither available nor projected for laboratory XUV (soft x-ray and ultraviolet radiation) lasers. Higher-energy storage rings (5 to 6 million eV) would generate significantly less coherent radiation and would be further compromised by additional x-ray thermal loading of optical components. Synchrotron radiation is discussed further in the article on **Particles (Subatomic).**

To extend scientific and technological opportunities, authorities suggest that a bright source of tunable, partially coherent, XUV radiation is needed. Coherence, in the limited sense used here, refers to the ability to form interference patterns when wave fronts are separated and recombined. The availability of a tunable source of coherent soft x-rays, combined with other developments in x-ray optical techniques, would make it possible to construct an x-ray microprobe of sufficient intensity to permit fundamentally new, phase-sensitive experimentation in a number of scientific and technological fields. Various imaging and scattering techniques would be enhanced by the greatly increased photon flux available to study small samples, as well as providing the capability of tuning the radiation to the wavelength of interest. For example, with soft x-rays well matched to the absorption edges of elements, such as carbon (284), nitrogen (400), and oxygen (532), as well as other elements of relatively low atomic numbers (Na, P, S, K, and Ca), it should be possible to study elemental distributions and motion within biological specimens without the need for dehydration, fixing, or staining (according to Attwood, et al.). Three-dimensional imaging, made possible by combining partially coherent undulator radiation and x-ray microholographic techniques, would complement the information available from electron microscopes.

Extensive development and experimental designs of soft x-ray lasers have been underway at the Lawrence Livermore National Laboratory and the Princeton Plasma Physics Laboratory, among other research institutions.

Researchers Suckewer and Skinner (Princeton University Plasma Physics Laboratory), in an early 1990 paper, observe, "Most of what is known about the internal structure of cells has been learned by the development and application of the techniques of electron microscopy. This knowledge rests on the premise that the intensive procedures necessary to prepare a specimen for electron microscopy do not significantly influence the structure, form, and high-resolution detail observed. Nonetheless, unanswered questions remain about the fidelity of the image of a cell that has been fixed, stained with heavy metals, and sectioned to the original living cell. X-ray microscopy offers a new way to look at unaltered cells in their natural state." X-ray laser microscopy can offer numerous advantages in this regard.

In summarizing the current status of soft x-ray laser research, the aforementioned researchers comment, "The general impact soft x-ray lasers will have in science and technology will depend on improvements in their performance and cost. It is necessary for their successful commercialization that these devices operate routinely at high gain-lengths ($GL > 4$), with the use of a low-cost driver laser, and this needs more system development and engineering. Most applications of visible-wavelength lasers are based on the fact that the brightness of these lasers is several orders of magnitude greater than that of conventional spontaneous emission sources, and this is achieved principally by the laser cavity mirrors. This technology is significantly more difficult in the x-ray region because of intrinsic limitations of x-ray absorption in materials and present limits in the soft x-ray laser pulse lengths. Nevertheless, a 'revolution' in x-ray optics is under way and the precedent of visible-wavelength lasers illustrates the potential benefits awaiting the creative inventor of applications of this technology to novel fields."

Laser Applictions

During the early phases of laser development (late 1950s to early 1970s), there was a high tempo of research activity and confidence in the ultimate potential for practical applications. Some of the early suggestions for laser use included instrumental applications in metrology and spectroscopy, as well as working tools for industry, such as cutting, welding, and annealing. But during that period it was also observed by some researchers that the laser was an invention for nonexistent needs. The decade of the 1980s removed all such doubts, and as science enters into the last decade of this century, the laser has become established as an essential component in numerous laboratory research programs and industrial and medical applications. Just a cursory inspection of the literature during the late 1980s and early 1990s is indicative of the wide scientific interest in the laser.

A number of representative uses are described here; many other uses are covered elsewhere in this encyclopedia. Check the alphabetical index. The use of lasers in fiber optic communications is described in the article on **Telephony.** Among medical applications for laser technology, photocoagulation procedures appear to predominate. Among other medical entries in this book, check the article on **Vision and the Eye.** Laser fusion is described under **Fusion Power.** Lasers are also described in the article on **Light.** The use of lasers in metrology is further described under **Interferometer** and in surveying and leveling large plots of land under **Irrigation.**

Atomic Cooling and Trapping. During the last few years, the ability to control the position and velocity of isolated atoms and microscopic particles has progressed markedly. As pointed out by S. Chu (Standford University) in a late 1991 paper, "Light can exert forces on an atom because photons carry momentum. The exchange of photon momentum with an atom can occur *incoherently,* as in the absorption and reemission of photons, or *coherently,* as in the redistribution (or lensing) of the incident field by the atom."

Coherent interaction is called the *dipole force*. The incoherent interaction that alters the momentum of an atom is called the *scattering force*.

Successful atom manipulation, however, often depends more upon cooling the atoms than upon exciting the aforementioned forces. Dramatic cooling of atoms to extremely low temperatures is accomplished by employing counterpropagating laser beams, arranged along x, y, and z axes—in essence, creating three-dimensional cooling. As pointed out by Chu, "Because the cooling force is viscous (linearly proportional to the velocities of the atom for low velocities), we named the laser beams that generate the drag force, 'optical molasses'."

In 1991, a research team (Ecole Normale Superieure, Paris) reported the cooling of a sample of cesium atoms to 2.5 μK. At about the same time, a research group (Joint Institute for Laboratory Astrophysics, Boulder, Colorado) reported the achievement of 5 μK. The aforementioned "optical molasses" technique was used in both cases.

Laser cooling, trapping, and related techniques are finding numerous research and practical applications. For example, practical laser-cooled atomic clocks are now possible, constituting a major improvement in accuracy over present atomic clocks. As mentioned by Chu, "A cesium time standard based on a sealed design for which the cooling, manipulation, and detection of the atoms are all done with diode lasers should exceed the stability of the best present-day time standards."

In an excellent paper, Chu (see reference) observes, "Perhaps the most exciting applications in the field of laser cooling and trapping will come out of the ability to study problems in polymer physics and biology on a single molecular basis. Normally one examines the behavior of a large number of molecules, and the fundamental chemistry of the molecules must be inferred from the average behavior of the entire ensemble. On the other hand, the processes that govern the behavior of a single molecule are important: for example, the nucleus of a cell has a single molecular copy of its genetic blueprint, and its chemistry depends in part on the chemistry of single molecules."

In a 1990 paper, Zewail (see reference) describes how atoms can collide, interact, and give birth to molecules in less than a trillionth of a second. As an example of how high-speed imagery has improved over the years, he compares photos of a galloping horse (10 meters per second) taken in 1887 with quantitative observations (made in 5 trillionths of a second) of hydrogen iodide colliding with carbon dioxide to form carbon monoxide, hydroxide, and iodine.

Lasers as Mini-Manipulators. As scientists continue to probe the very minute aspects of natural organs and substances (nanotechnology), small lasers have been found to possess "manipulative" abilities of a kind not envisioned in the early years of laser technology. Scientists (Massachusetts Institute of Technology) in 1990 reported of how lasers can be used effectively as manipulators at the microscopic level. In a study of "mechanoenzymes," which are responsible for the rotary motions of flagella, laser light was used to lift up, move, and position microscopic objects with the "pressure" of the laser light itself, a phenomenon that has been described by Amato as "akin to a blast of air levitating a plastic ball."

Laser mini-tweezers also have been used to clip off regions of chromosomes, moving organelles around inside cells, pushing molecules tiny distances within crystals, and, when used as tiny scalpels or scissors, to catch, trap, puncture, and splice subcellular structures. Recently, measurements have been made of the elastic properties of DNA.

Also, it has been found that bacteria can be moved around in a water solution without apparent damage to the organism. Medical applications are described later.

Laser Spectroscopy. In the early years of laser technology, spectroscopy was one of its major uses, an application that has expanded markedly during the past few decades. The techniques of laser spectroscopy parallel those of microwave or radio-frequency spectroscopy, but because lasers are imbued with high spectral purity, they permit grossly improved resolution of fine detail. Early lasers were limited to molecular lines that were coincident with the laser wavelengths. Then lasers using fluorescent organic dyes appeared. These instruments had relatively wide emission bands, offering a tuning capability. Both continuous-wave and pulsing dye lasers have been widely used in most of the visible and near-visible ranges of the spectrum. During the interim, much progress has been made, particularly in providing tunability to lasers. For example, a methyl fluoride molecular gas laser is continuously tunable over broad portions of the far infrared, a region that previously had been difficult. A highly schematic diagram of the operating principle used in early laser-probe emission spectrography is given in Fig. 9.

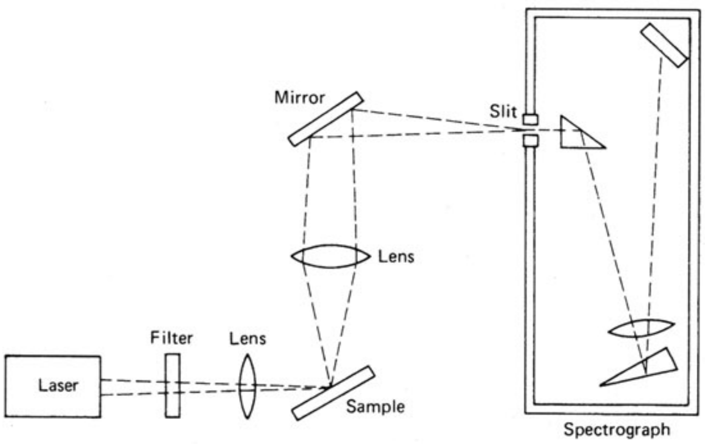

Fig. 9. Operating principle of laser-probe emission spectrography.

A particularly interesting development is that of the so-called "atomic fountain." As noted by Chu in 1991, "The precision of a spectroscopic measurement depends on both the high Q (Q = quality factor of the resonance defined by $Q = V/\Delta V$) and the signal-to-noise ratio of the signal. Thus, it is important to create a high-flux source of cold atoms. Also, many applications would benefit from a continuous beam of atoms instead of the pulsed sources." An extreme limit of a slow beam is an "atomic fountain," which first was envisioned by Zacharias in the early 1950s. A group of Stanford University scientists has constructed an atomic fountain by first trapping atoms from a thermal beam in a magneto-optic trap and then pushing the atoms upward with a pulse of light from a continuous-wave laser. See Fig. 10.

One reason for the rapid advancement of chemical reaction dynamics research has been the availability of tunable laser sources that operate throughout the infrared, visible, ultraviolet, and vacuum ultraviolet regions of the spectrum. Techniques have been developed to probe almost any kind of atomic or molecular state, quite often with sensitivities approaching number densities of 10^5 cm^{-3}, and, in special situations, with detection sensitivity for single atoms. As stressed by Leone, the types of lasers range from solid-state tunable diode lasers in the infrared to liquid-phase organic dye lasers in the visible. By using nonlinear optical techniques, the outputs from high-power, pulsed visible dye lasers can be summed and mixed to yield useful tunable ultraviolet and vacuum ultraviolet light, with wavelengths as short as 100 nm.

In a detailed reference, Grant and Cooks (Purdue University) explain in considerable detail the combining of the latest advances in mass spectrometry with laser spectrometers. This technique is contributing in a major way to studies of chemical dynamics, cluster structures, and reactivity, and to the elucidation of the properties of highly excited molecules and ions.

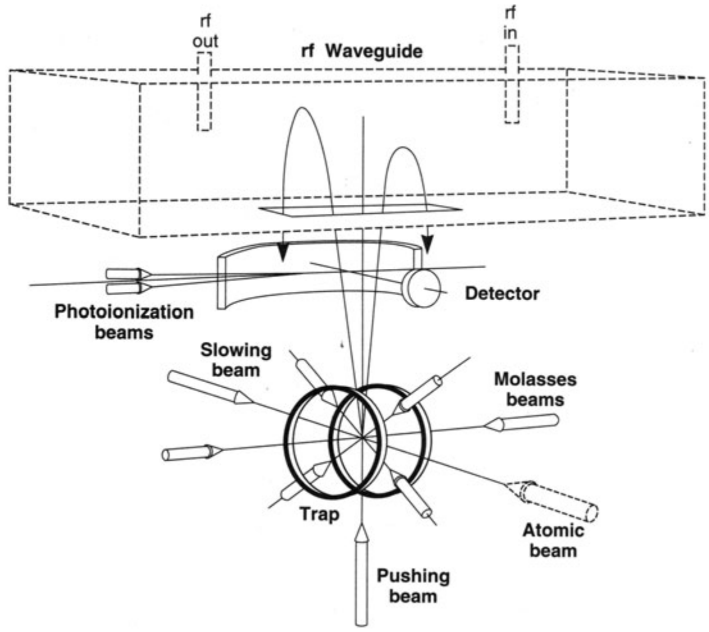

Fig. 10. The atomic fountain makes it possible to determine precisely the energy states of atoms. Upon injection, the atoms in question are slowed down by a laser beam. Then, the atoms are captured and cooled by means of a magnetic field and several light beams. The cooled atoms follow a ballistic trajectory through a radio frequency (rf) waveguide and a resonant photoionization detection region. (*After Kasevich, Riis, Chu, and DeVoe.*)

Laser Remote Sensing of Atmospheric Properties

LIDAR (an acronym for light detecting and ranging) is analogous to radar. In lidar, the projection of a short laser pulse is followed by reception of a portion of the radiation reflected from a distant target or from atmospheric constituents, such as molecules, aerosols, clouds, or dust. As explained by Killinger and Menyuk (see reference), the incident laser radiation interacts with the aforementioned constituents to cause alteration in the intensity and wavelength in accordance with the strength of the optical interaction and the concentration of the interacting species in the atmosphere. Information on both composition and physical state of the atmosphere can be deduced from lidar data. The range of the interacting species can be determined from the temporal delay of the backscattered radiation. See Fig. 11.

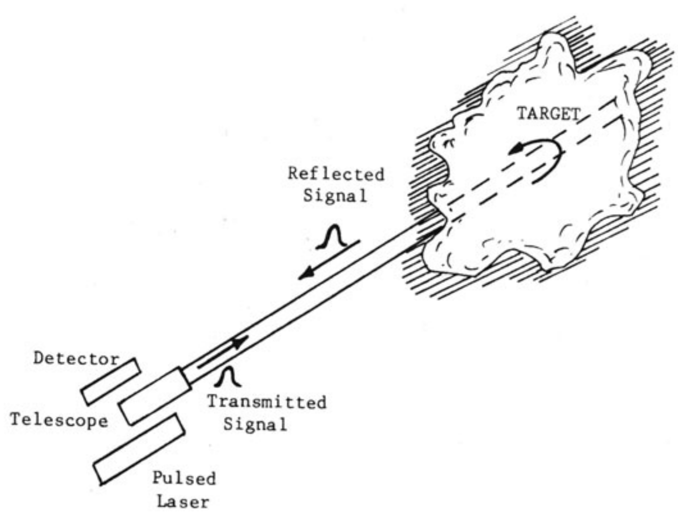

Fig. 11. Basic components of lidar system used for remote sensing of the atmosphere. Backscattered information sometimes will contain spectral information useful for determining composition and physical characteristics of the cloud or of the intervening atmosphere.

Among specific uses of lidar have been: (1) measurement of movement and concentration of urban air pollution; (2) determinations of chemical emissions from and in the vicinity of industrial plants; (3) determination of atmospheric trace chemicals in the atmosphere; (4) measurement of the velocity and direction of winds near storms and airports, including windshear and gust fronts; (5) determination of the global circulation of volcanic ash emitted into the atmosphere, relatively recent examples including Mount Pinatubo and Kilauea; among several other applications.

Shortly after the discovery of lasers (early 1960s), Fiocco and Smullin bounded a laser beam off the moon (1962). These researchers also investigated the turbid layers in the upper atmosphere. As early as 1963, Ligda used a ruby laser to obtain the first lidar measurements of cloud heights and tropospheric aerosols. In 1964, Schotland used a temperature-tuned ruby laser to detect water vapor in the atmosphere. Lidar, in recent years, has been greatly improved because of the availability of several kinds of laser sources and improvements made in optical instrumentation and data processing.

As summarized by Killinger and Kenyuk, the future of laser remote sensing is promising and will depend upon several factors, including: (1) development of practical, eye-safe laser sources that cover certain spectral gaps where lidar is currently weak; (2) a further simplification of lidar systems, including lowering size and cost of equipment needed; and (3) more experience to be gained from promising new applications. Among these new applications are: (a) detection of methane gas leaks in coal mines, using a diode laser lidar system; (b) detection of methane and natural gas leaks in industrial plants, using a laser coupled to a low-loss optical fiber network; (c) measurement of global wind fields through the use of Doppler lidar systems mounted in a satellite as a means for improving weather forecasting; and (d) the planned use of lidar on the NASA space-borne Earth Observing System for measurements of global temperature, water vapor, and pressure.

Classification of Lidar System. Lidar systems can be classified on the basis of particular optical interactions which they utilize. Classes of lidar include:

1. *Atmospheric backscatter lidar*, wherein the lidar system transmits one laser wavelength and detects changes in the backscatter due to the aerosols or dust in the atmosphere. This is the most common type of lidar and consists of a nontunable, high-power, pulsed laser. Atmospheric constituents having comparatively large optical scattering cross sections are relatively easy to detect. These systems are used in tracking turbid effluent and gas plumes from factories as well as for mapping rain, snow, ice crystals, and dense clouds in the atmosphere. This type of system was used for checking volcanic ash in the atmosphere.

2. *Differential-absorption lidar* (*DIAL*), a system which measures the concentration of a molecular species in the atmosphere. This is accomplished by transmitting two wavelengths, only one of which is absorbed. The difference in the intensity of the returns at the two wavelengths is measured. Backscatter in DIAL may come from a hard target or aerosols and dust. One wavelength will be absorbed by the target molecules; the other wavelength will not be absorbed. Many DIAL studies have been carried out in the infrared (IR) range, where almost all molecules of interest have extensive absorption bands. Molecules so far studied include SO_2, NH_3, O_3, CO, CO_2, HCl, NO, N_2H_4, N_2O, and SF_6.

3. *Fluorescence lidar* uses two wavelengths (as in DIAL) plus spectrometric techniques for separating the wavelength-shifted fluorescence signal from the strong Rayleigh backscatter in the atmosphere. The laser is tuned to an absorption line of the species to be measured. Reradiated fluorescence is detected by selective spectral filtering of the returned radiation. The fluorescence radiation may be at the same wavelength as the excitation wavelength, or it may have a longer wavelength because of the red-shift. The backscatter coefficient for fluorescence is greater in the ultraviolet (UV) than in the IR—this due to combined effects of absorption cross section, which is greater in the UV than in the IR. For some applications, fluorescence lidar is limited for remote sensing because of detector sensitivity coupled with solar background radiation. The latter tends to confine fluorescence measurements to nighttime studies and to wavelengths shorter than 1 micrometer, where photomultiplier detection can be used. Nevertheless, some investigators have been quite successful in using the method, particularly in the study of alkali metal (Na, K, Li, and Ca) profiles at altitudes of 80 to 100 km. The method also has been useful for studying the hydroxyl free radical (OH). This radical is of principal interest because of the catalytic role which it exerts in atmospheric chemistry. The OH radical, along with chlorine and nitrogen oxides, is involved in the ozone destruction cycle.

4. *Raman lidar*, a method that is limited by the small optical interaction strength for Raman scattering. High-energy pulsed lasers are employed in this method. The method is limited to the UV or visible regions to permit the use of sensitive photomultiplier tubes for detection. Raman lidar has been used effectively for species that either are at close range or in high concentration, such as N_2, O_2, and H_2O. As mentioned by Killinger and Menyuk, this method does have some attractive features, the most noteworthy of which is that the laser wavelength need not be tuned across an absorption line because the spectral information is given by the frequency shift of the emission (independent of laser wavelength).

5. *Doppler lidar*, a method that detects only a very narrow spectral range ($\sim 10^{-5}$nm) that encompasses the Doppler-shifted backscatter lidar return. Doppler shifts in the return lidar signals have been used to measure wind velocities and to differentiate between molecular and aerosol returns in the atmosphere. Optical heterodyne techniques are used to detect the shifts which are very small—for example, a fractional change in frequency of about 10^{-8} for a velocity of 1 m/sec at a wavelength of 10 micrometers. Carbon dioxide lasers which provide high power and stable single-frequency operations are commonly used in Doppler lidar systems. These are in the 10-micrometer range. The system has provided information on boundary layer flow near storm gust fronts and windshears near airports. They also have been used to measure aircraft vortices and clear air turbulence. See Fig. 12.

Doppler broadening effects also have been used to separate backscattered lidar signals into molecular and aeosol components. Characteristics of some lidar systems are summarized in Table 2.

Laser Techniques in High-Pressure Geophysics

The experimental establishment of high pressures and temperatures in the laboratory comparable to conditions found in the Earth's interior is described in the article on the **Diamond Anvil High Pressure Cell.** Laser techniques used in conjunction with the diamond cell make it possible to study the high-pressure properties of material which here-

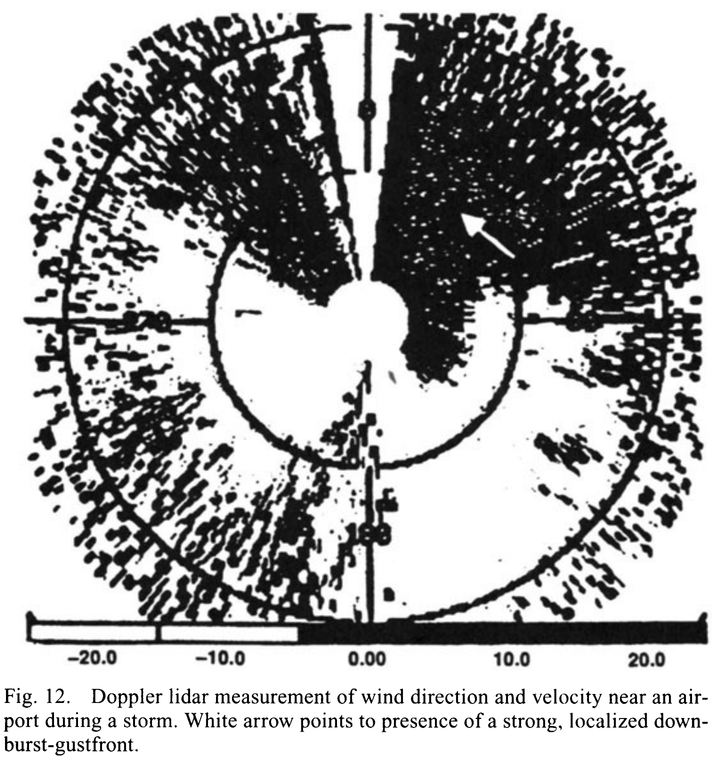

Fig. 12. Doppler lidar measurement of wind direction and velocity near an airport during a storm. White arrow points to presence of a strong, localized downburst-gustfront.

TABLE 2. SUMMARY OF LIDAR SYSTEM CHARACTERISTICS

Type of Lidar	Type of Laser Used	Nominal Accuracy	Range (km)	Atmospheric Targets
Atmospheric backscatter	Ruby, Nd:YAG	1–10%	10–50	Dust, clouds, volcanic ash, smoke plumes
DIAL, Raman	Dye, CO_2, optical parametric amplifier, CO:MgF_2	1 ppb–100 ppm	1–5	H_2O, O_3, SO_2, NO, NO_2, N_2O_3, C_2H_4, CH_4, HCl, CO_2, CO, Hg, SF_6, NH_3
Fluorescence	Dye	10^2–10^7 atoms/cm	1–90	OH, Na, K, Li, Ca, Ca^+
DIAL, Raman	Dye, Nd:YAG	1 K, 5 mbar	1–30	Temperature, pressure
Doppler	CO_2	0.5 m/sec	15	Wind speed

After Leone (1987).

tofore had to be inferred from samples found on the surface. Spontaneous Raman scattering of crystalline and amorphous solids at high pressure demonstrates that dramatic changes in structure and bonding occur on compression. High-pressure Brillouin scattering is sensitive to the presure variations of single-crystal elastic moduli and acoustic velocities. Laser heating techniques with the diamond anvil cell can be used to study phase transitions, including melting, under deep-earth conditions. Laser-induced ruby fluorescence has been essential for the development of techniques for generating maximum pressures now possible with the diamond anvil cell, and currently provides a calibrated in situ measure of pressure well above 100 gigapascals. Hemley, Bell, and Mao (Carnegie Institution of Washington) point out that applications of new spectroscopic techniques, such as double resonance, ultrafast kinetics, Fourier-transform, Raman, and nonlinear optical methods, are likely prospects in future work on geophysical problems with the diamond anvil cell. Recent high-pressure studies involving the use of picosecond spectroscopy and hyper-Raman scattering of perovskites may be representative of this trend. Time-resolved studies may permit the detailed investigation of the kinetics of high-pressure phase transitions and the rheology of minerals under in situ deep-earth conditions. The combination of spectroscopic and x-ray diffraction probes with laser-heating techniques may yield detailed structural information on earth materials at high temperatures and pressures, thus advancing an understanding of the connection between atomic-scale properties and global deep-earth processes.

Laser Metrology

Lasers are widely used for making precision measurements of geometric variables. In the early 1960s, laser pioneers demonstrated precise measurements with the device. Lasers introduced the concept of frequency metrology as contrasted with wavelength metrology. In an early experiment with a super-stabilized laser, scientists at the Massachusetts Institute of Technology during the early 1960s worked out a laser version of the famous Michelson-Morley experiment at Case Institute of Technology in Cleveland. The MIT scientists concluded that an advance in measurement sensitivity by a factor of 1000 over the Michelson-Morley data was potentially available through the use of frequency rather than length metrology. In 1962, the first laser measurement of the speed of light (*c*) was made, yielding a value of 299,792,462 ± 18 meters per second. The generally published value of *c* is 299,792,500 meters per second.

Since that time, several more sophisticated determinations have been made by leading metrology laboratories, including the National Institute of Standards and Technology (Boulder, Colorado), the National Physical Laboratory (United Kingdom), the Laboratoire de Physique des Lasers (Villetaneuse, France), and the Laboratory for Spectroscopy (Russia), among others.

At the practical manufacturing level of metrology, laser guidance systems can be used. Mergler (Case Western Reserve University) introduced a machine in 1978 along these lines. In a conventional machining operation, a part is cut, then measured, then remachined until the required dimensions are obtained. Manual measuring methods are tedious, time consuming, and somewhat limited in accuracy. In Mergler's system, a small modulated gas laser beam follows the surface of the part being machined and, within a precision of 1/5000 inch (0.005 millimeter) measures the piece as it is being cut. In later systems, the gas laser was replaced by a solid-state laser which occupies less space. See also laser interferometer described in entry on **Interferometer**; and **Electron Beam Lithography**.

Laser Doppler Flowmeter

As shown in Fig. 13 fluid flow can be determined by measuring the doppler shift in laser radiation scattered from particles in the moving fluid stream. No sensor is required in the moving stream. The laser radiation focal point can be moved across the flow tube to measure velocity profiles. Fluid linear flows from 0.01 to 5000 inches (0.03 centimeter to 127 meters) per second have been measured. Contaminants, such as smoke, may have to be added to gases to provide scattering centers for the laser beam.

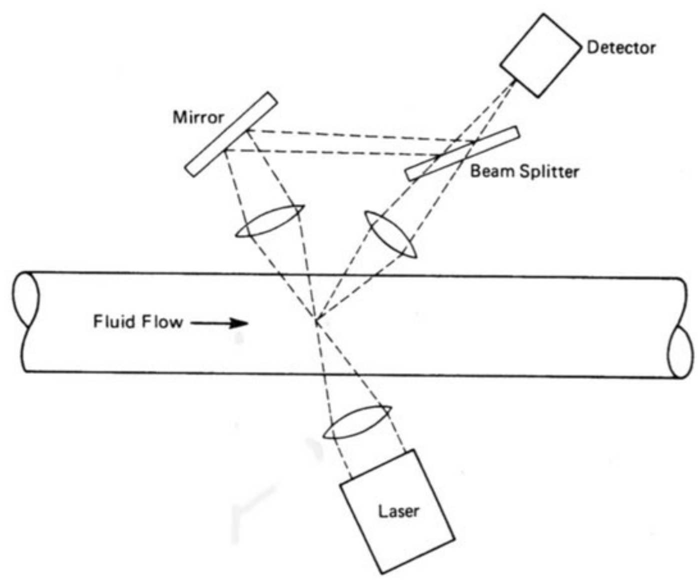

Fig. 13. Operating principle of laser doppler flowmeter.

Laser Gyroscope

As early as the beginning of this century, some investigators suggested that light will exhibit gyroscopic behavior, that is, the time required by light to traverse a circular pathway depends on whether the pathway is stationary or rotating. Thus the time difference can be used as a measure of the amount of rotation. The practical application of this observation, however, had to await vast improvements in optical systems, including the discovery of the laser, advances in fiber-optics, and better reflective mirrors. Within recent years, this principle has been applied in two configurations—fiber gyroscopes and ring-laser gyroscopes. The latter is described briefly here. As of the late 1980s, several aircraft depend upon ring-laser gyroscopes instead of their mechanical counterparts. The ring-laser gyroscope is more sensitive, has virtually no moving parts, and is as accurate as the best mechanical instruments. The rotation-induced difference in length of light path traversed is called the Sagnac effect, after the researcher who first demonstrated the phenomenon in 1913.

As previously mentioned in this article, a laser is a resonant cavity. C.V. Heer (Ohio State University) in 1958 proposed that a resonant cavity could be used to measure rotation rates. In such an instrument, light circulates many times around a given path, not just back and forth between two mirrors. The first gyroscopes of this kind were constructed on a large scale, consisting of four glass tubes, each a meter long and arranged in a square. Light traveled around the device by placing a mirror in each corner. Over the last several years, the device has been markedly reduced in size (fits in the palm of the hand). Contemporary gyroscopes of this type are made from a single block of glass, into which a square channel is drilled. The channel is filled with a mixture of helium and neon. The laser is completed by attaching a small number of electrodes and four mirrors. As explained by Anderson (see reference), some ring-laser gyroscopes have a triangular channel and three mirrors; other have a hexagonal channel and six mirrors.

Beyond the scope of this article, Anderson explains the operation of the gyroscope in intimate detail and describes two problems that have proved most vexing to manufacturers of the ring-laser gyroscope, namely, frequency locking at low rotation rates and the bias effect. Improvements in this instrument are expected in the relatively near future because of what scientists have recently learned pertaining to the phenomenon of optical phase conjugation. See **Light.**

Lasers in Manufacturing Operations. A principal advantage of the laser in manufacturing operations is its ability to apply an extremely high flux of energy to the surface of a workpiece, as compared with traditional heat sources, such as flames, torches, electric arcs, and plasma jets. For manufacturing operations, lasers are usually placed in two categories. (1) *Light-duty lasers* range from a few tens of watts to a few hundred watts. Typical applications include cutting and drilling ceramic substrates in the electronics industry, drilling gems (for example, rubies in watchmaking), and cutting light-gauge metals as well as cloth, plastics, wood, and a variety of materials. Light-duty lasers that have been used include ruby lasers, neodymium-doped glass lasers, and neodymium-doped yttrium aluminum garnet lasers, among others. Depending upon the particular laser selected, the laser may operate in a pulsed or continuous mode. Argon and CO_2 lasers usually are operated in the continuous-wave mode. (2) *Heavy-duty lasers* range from a few kilowatts to a few tens of kilowatts. Typical applications include pipeline welding, automobile part welding, surface heat-treating of engine and other parts, with the applications expanding as experience is gained.

The high flux of electro-magnetic energy applied to the surface of a workpiece by a laser is absorbed in an outer layer only about 10 nanometers thick. Thus the heat source is confined essentially to a thin film. Through careful design of equipment, the heat energy required is maintained in a comparatively small region, thus preventing or reducing thermal damage to the rest of a given part and achieving a very high energy efficiency, estimated to range from 10 to 1000 times greater than can be achieved with conventional energy sources.

The electronics industry utilizes laser welders for joining dissimilar materials, fixing electrodes to batteries and connectors to a host of devices. A whole new area of laser technology, sometimes called laser microchemistry, has been exploited in the microstructure engineering of semiconductors. Lasers are used to initiate chemical reactions which result in deposition of material at a surface, for removing materials, and for alloying or diffusively mixing two or more solids on microscopic spatial scales. Lasers thus have played a major role in establishing new dimensions in microfabrication technology. It is possible to use a single laser to produce both gas-phase photolysis and surface heating. As described by Christensen (see reference), solar cells have been fabricated by using a UV laser to photodissociate trimethylboron, $B(CH_3)_3$, over a silicon surface in the manufacture of solar cells. The laser also heats the surface so that the boron atoms absorbed on the surface after the photolytic step rapidly diffuse into the bulk of the material. After irradiation, the silicon is heavily doped with boron near the surface, and the p-n junction thus formed functions as a photovoltaic cell. In some other applications, it has proved advantageous to use two lasers of different wavelengths to separately achieve photolysis and heating.

Perspective. The industrial applications for lasers developed comparatively slowly. As previously mentioned, lasers depend upon raising active molecules of the lasing medium to what might be called an *upper laser level* of energy, after which they relax to a *lower laser level*. Energy is given up during this process. Part of this energy is represented by photons of which the laser beam is composed. The other part is waste heat, which raises the temperature of the lasing medium. Thus, an excess of waste heat be removed so that the upper-level population can be maintained, a significant problem in the case of a continuously emitting high-power laser. Higher packets of energy in pulses can be attained, but waste heat must be removed by conduction between pulses. The end result is a pulsed high-power laser, but one that has a comparatively low average energy level simply because of the pauses in between.

In early laser designs, the quantity of waste heat generated was a limiting factor and consequently the average power output was low. Such lasers were excited by diffuse longitudinal electric discharges in long tubes with relatively large diameters. Heat generated at the center of the tube diffused to the side walls essentially by conduction, and the rate of heat transfer varied inversely with the tube radius and essentially directly with the length of the tube. Thus, the length of laser tubes increases as greater output power was sought.

Various component cooling schemes were proposed and used, but a major improvement was made when the concept of cooling a flowing laser medium was proposed, thereby taking advantage of the far more effective cooling by convection than by conduction. Gas lasers were considered the most apt for application of this concept and this led to the gas dynamic laser.[1]

With this concept, gas dynamic lasers increased in power outputs from less than 10 kilowatts by a factor of 13 to 14 within less than a decade (by the late 1960s). Success with the early gas lasers in this respect catalyzed a number of other refinements and improvements. However, the problem of maintaining a high-pressure glow discharge remained. Population inversion can be produced when electrons in an ionized gas are at a temperature relatively high as compared with the kinetic temperature of ions or molecules. This is a condition referred to as *glow discharge*. But an are may form when the discharge is destabilized as the result of greatly increased gas pressure. Overheating of the molecules and ions destroys the population inversion. This problem was overcome by the concept of the ionizer/sustainer.

With the availability of high-power continuous electric discharge lasers capable of operating at up to 20 kilowatts output, a number of the previously predicted applications for lasers became practical. One of the first uses of a laser beam strictly for its power in cutting (exploding) a material was the fabric cutting system developed by Hughes Aircraft Company (circa 1966-1967) for which U.S. Patent No. 3,761,675 was awarded to W.J. Mason, D. W. Wilson, D. M. Considine, F. J. Viosca, and J. P. Wade on September 25, 1973. See Fig. 14. In this system cloth is carried in a single layer into a cutting area where a laser beam focused on the cloth is directed by computer commands to travel within the cutting area so as to cut many patterns in the cloth rapidly and accurately. The cut produced by the focused laser beam is sharp and narrow, leaving the fabric unfrayed. With synthetic materials, such as nylons and Dacrons®, the laser beam also serves to seal the cut edges by melting them during the cutting process. Unlike a mechanical blade, the

[1]Invented by Kantrowitz in the late 1960s. In essence, the device had two compartments separated by a nozzle. In the first compartment, gas was held at a temperature of about 1400 K and pressure of 17 atmospheres. This high-pressure compartment held about 10% of the active CO_2 molecules in the total system. Expansion of this gas through an orifice caused cooling. Because of the cooling, the lower-level population essentially vanished a few centimeters downstream from the nozzle. This occurred before the upper-level population had an opportunity to decline significantly. The population "inversion" resulting was adequate for effecting a laser beam of considerable power.

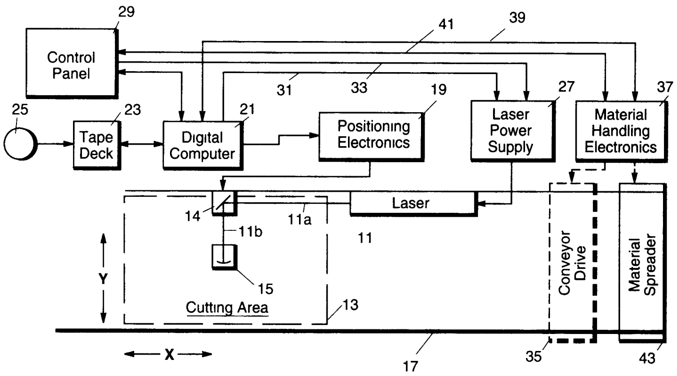

Fig. 14. A cloth cutting system wherein cloth is carried in a single layer into a cutting area, where a laser beam is focused on the cloth and is directed by computer commands to travel within the cutting area so as to cut a plurality of patterns through the cloth rapidly and accurately. Invented by W. J. Mason, D. W. Wilson, D. M. Considine, and J. P. Wade (*Hughes Aircraft Company*) in 1973, this was one of the very early and successful industrial applications of the laser. Diagram is part of U.S. Patent 3,761,675.

laser beam does not dull; its cut remains uniform and is effective in cutting a wide range of materials, even those having metallized threads.

In the early 1980s, a helium-neon laser was used in a scanner system for inspecting textiles. The system uses laser output split into three beams, each of which scans the fabric independently in a pattern covering the entire surface. The system, moving at a rate of four meters per second, detects flaws through changes in reflected light and flags these areas for elimination or repair. In terms of economics, one laser system working one shift performs the same function as human inspectors at two plants working two shifts.

High-power lasers can perform many metalworking operations, including welding, cutting, surface hardening, and surface alloying. For small devices, the laser can perform much as a conventional electron beam, but without requiring the need for operation under a vacuum. High power densities can be achieved—up to 10^6 watts per square centimeter. It has been shown that a 16-kilowatt laser can make a 0.75-inch (1.9-centimeter) penetration weld in stainless steel at a rate of about 30 inches (76 centimeters) per minute. The laser beam can be directed by mirrors, thus making it effective for welding pipe from the inside. It also has been shown that a continuous-wave carbon dioxide device (15 kilowatts) can be used for welding half-inch (1.2-centimeter) thick steel plates at the rate of about 50 inches (127 centimeters) per minute. If the laser is focused to a spot size of about 0.03 inches (0.08 centimeter) in diameter, power densities of some 2200 kilowatts per square centimeter are produced.

Laser Recording. For many years, it has been known that lasers can be used to encode information on materials that respond in an irreversible manner to exposure to high-intensity light. However, it is only comparatively recently that the concept has been reduced to commercial practice—with the almost sudden appearance of optical disk recording (compact disk) in the entertainment field. It is because of the coherence and relatively short wavelength of laser radiation that such large volumes of information can be written onto a very small space of the recording medium. The potential for microlasers in this field is discussed earlier in this article.

Lasers in Medicine

Medical applications for lasers have developed at a very rapid rate and one that is continuing to expand today in an exponential fashion. Much of the laser equipment for the manipulations required in surgical procedures is available. Training professionals in the effective use of the equipment and equipment costs are the only factors retarding even faster growth of the field.

Some of the applications of lasers in various medical situations are described in other articles in this encyclopedia. Some of the more widely accepted laser surgery procedures include:

Abdominal cavity—Repairing hernias.
Brain—Shrinking or removing benign or malignant tumors.
Ears—Repairing damaged portions of inner ear.
Eyes—The earliest medical use of lasers was in eye surgery. Reattaching torn retinas; opening blocked tear ducts that cause dry-eye syndrome; coagulating bleeding blood vessels that cause diabetic retinopathy; reducing fluid buildup in glaucoma; clearing the cloudiness that sometimes remains after conventional cataract surgery.
Feet—Removing corns, plantar warts, and ingrown toenails.
Gastrointestinal tract—Removing hermorrhoids and intestinal polyps.
Genital-reproductive tract—Vaporizing venereal warts, fibroids, external genital growths, and tumors in the cervix, vagina, and perenium; vaporizing excess uterine lining that causes bleeding.
Mouth—Removing or vaporizing superficial tumors.
Nose—Vaporizing polyps; removing adenoids and the excess tissue that causes sinusitis.
Skin—Erasing port-wine stains, birthmarks, tattoos, age spots, and freckles; removing warts, broken facial capillaries, and precancerous patches.
Urinary tract—Vaporizing tumors; opening a narrow urethra.

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LASER ALTIMETER. See **Moon (The).**

LASER DOPPLER FLOWMETER. See **Flow Measurement.**

LASER GLASS. See **Glass.**

LASER IONIZATION. See **Mass Spectrometry.**

LASER (Telephony). See **Telephony.**

LASSA FEVER (African Hemorrhagic Fever). Lassa (arenavirus) disease is named after the village in Nigeria where the causative agent was first isolated and the disease was first described. The natural reservoir is the multimmamate rat (*Mastomys natalensis*) through which the virus is passed and continued via the uterus or maternal milk. In humans, the disease is acquired through contamination of food or water by rat urine or feces, aerosol transmission, or accidental inoculation or blood contact.

The incubation period ranges from 7 to 18 days. Lassa fever virus may cause as many as 300,000 human infections and 5000 deaths per year in areas where the disease is endemic in central and west Africa, notably in Liberia and Sierra Leone.

The disease rarely occurs in North America, Europe, and other regions outside the aforementioned regions of Africa. When the disease does occur elsewhere, it has been imported by a person who has had recently returned from a visit to Africa. Such a case was reported in Illinois in 1989. That case was subjected to considerable analysis. Similar imported cases have been reported in Canada, Europe, Israel, and Japan.

As the result of the study, involving patients and persons with whom the patient had been in contact, the Centers for Disease Control (U.S.) now suggest that complete isolation, as was the former practice, can be mitigated to meticulous barrier nursing procedures and universal precautions to prevent contact with contaminated blood or other body fluids. The revised guidelines allow greater medical access to patients with Lassa fever and thus better overall care for the patients. Prophylactic use of ribavirin also may prevent spread of the disease to exposed persons. Detailed procedures are given in the Holmes reference listed.

At autopsy of persons who succumb to the disease, one finds hemorrhages in the gastrointestinal tract, focal lesions, congestion and pneumonitis in the liver, and edema and petechial hemorrhages in the spleen and myocardium. Incipient mortality is signalled by high fever and conjunctival hemorrhage.

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R. C. V.

LASTUS RECTUM. See **Parabola.**

LATENT HEAT. Heat which is gained by a substance or system without an accompanying rise in temperature during a change of state. As examples, the latent heat of fusion is the amount of heat necessary to convert a unit mass of a substance from the solid state to the liquid state at the same temperature, the pressure being that to allow coexistence of the two phases. A considerable part of the latent heat arises from the entropy increase consequent on the greater disorder of the liquid state. The latent heat of sublimation is the amount of heat necessary to convert a unit mass of a substance from the solid state to the gaseous state at the same temperature, the pressure being that to allow coexistence of the two phases.

LATERAL. A force which acts on a structure or a structural member in a transverse direction is sometimes called a lateral load. The wind blowing upon the exposed surface of a bridge or building at right angles to its length or upon the stationary or moving traffic using the bridge constitutes one type of lateral load. The sway of a moving train on a bridge or the centrifugal force transmitted if the bridge is on a curve is a type of lateral loading. A moving crane supported on girders exerts a side thrust on the girders which may also be included in this classification.

Trusses and girders, which constitute the main load-carrying members of bridges, are not ordinarily designed to carry side loads of this nature, and consequently have very little strength in that direction. For this reason, the trusses of girders of a bridge are joined together in a horizontal plane by a system of lateral bracing composed of struts and diagonals. These members are often referred to as the laterals or the lateral system. This lateral bracing stiffens the whole bridge and opposes any sidewise deflection or vibration.

The term lateral is also used in connection with sewerage systems. Any sewer which serves the abutting property owners and in which each owner has an equal right is a common sewer. A lateral sewer is one which has no other common sewer flowing into it.

LATERAL LINE. See **Fishes.**

LATERITE. The sub-aerial decay of rocks in tropical regions, having a distinctly moist or rainy climate, results in the development of a residual, reddish, and usually sticky soil frequently containing concretions. The principal products of laterization are the hydrated oxides of aluminum and iron either in the crystalline or amorphous form. If the concentration of iron oxide is sufficiently high the laterite may be valuable as an iron ore. If, on the other hand, the concentration of alumina is high the laterite may be valuable as an ore of that metal.

LATEX. Latex is a milky substance found in many plants. It is a complex emulsion in which such substances as proteins, alkaloids, starches, sugars, oils, tannins, resins, and gums are found. In most plants the latex is white; but in some it is yellow; in others, orange or scarlet.

The cells or vessels in which latex is found make up the laticiferous system. There are two very different ways in which this system may be formed. In many plants the laticiferous system is formed from cells laid down in the meristematic region of the stem or root. Rows of these cells are formed. The cell walls separating them are dissolved, so that continuous tubes, called latex vessels, are formed. This method of formation is found in the poppy family; in the rubber plant, *Hevea brasiliensis*; and in the *Cichorieae*, a section of the composite family distinguished by the presence of latex in its members. Dandelion, lettuce, hawkweed, and salsify are members of the *Cichorieae*. See also **Rubber (Natural).**

LATIN SQUARE. An experimental design based on a $p \times p$ array of p letters such that each letter occurs once and only once in each row and column; e.g., for $p = 4$ and the letters, A, B, C, D:

$$\begin{matrix} A & B & C & D \\ B & D & A & C \\ C & A & D & B \\ D & C & B & A. \end{matrix}$$

A layout of this type, for example, may correspond to 16 plots and four treatments represented by the letters.

The design may be generalized to allow for a further treatment represented by Greek letters and it is then known as a Graeco-Latin square, e.g.,

$$\begin{matrix} A\alpha & B\beta & C\gamma & D\delta \\ B\gamma & A\delta & D\alpha & C\beta \\ C\delta & D\gamma & A\beta & B\alpha \\ D\beta & C\alpha & B\delta & A\gamma \end{matrix}$$

In this case, no combination of Roman and Greek letters occurs more than once. More general designs are sometimes known as Hyper-Graeco-Latin squares. The purpose of all these designs is to provide independent comparisons of row, column, and treatment effects. See also **Orthogonal Squares.**

LATITUDE. The celestial latitude of a point on the celestial sphere is the spherical coordinate measured from the plane of the ecliptic along a great circle passing through the object and the poles of the ecliptic.

Because of the fact that the earth is not a perfect sphere, there are several different sorts of terrestrial latitude in use. In Fig. 1, we have an ellipse $PEP'E'$ representing a section of the earth in the plane of a meridian. C is the geometric center of the earth, and the line COZ' is the line to the geocentric zenith of the point. O. The angle ECO (ϕ') is the geocentric latitude of the point O.

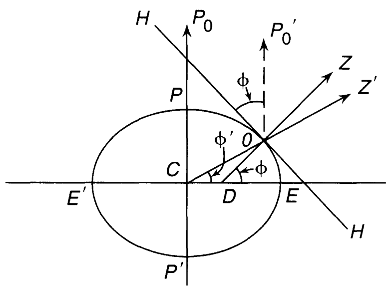

Fig. 1. Ellipse representing a section of the earth in the plane of a meridian.

The line DOZ represents the direction of gravity at the point O and extends to the astronomic zenith of O. The angle EDZ (ϕ) is the astronomic latitude of the point O. The difference between the astronomic and geocentric latitude of a point, the angle $COD = \phi - \phi'$, is defined as the reduction of latitude for the point O.

Because of local influences, such as massive mountains in the vicinity, the direction of the plumb line may not be strictly perpendicular to the surface of the earth. The geographical latitude of a point is the angle, measured in the plane of the local meridian, between the equator and a line drawn perpendicular to the theoretical geoid (surface of the earth) through the point in question. The difference between astronomic and geographic latitude is always relatively small, but by no means an inappreciable angle, and is known as station error. Station error is commonly between 4 and 6 seconds of arc, but occasionally amounts to 30 or 40 seconds.

In Fig. 1, CP represents the axis of rotation of the earth, which, if extended, will pierce the celestial sphere in its pole of rotation. The parallel line OP'_0 is the line from the observer at O to the pole of rotation of the celestial sphere, and the line HOH represents the plane of the astronomic horizon at O. HOP'_0 is the altitude of the pole of rotation at O, and inspection of the figure will indicate that this is equivalent to the angle EDZ. This gives rise to the common definition of the astronomic latitude of a point as the altitude of the pole of rotation of the celestial sphere at the point.

Astronomic latitude may be determined in a variety of ways by observation of the celestial objects. The most direct method is to observe the altitude of some object on the meridian whose declination is known. In Fig. 2, we have a representation of the celestial sphere drawn in the plane of the local meridian of the point O. In the figure, HOH' represents the plane of the horizon; $HPZQH'$ represents the local meridian; OP the direction of the pole of rotation; OQ the direction of the equator. $HOP = \phi$ (the astronomic latitude of O), and $H'OQ = 90 - \phi$. Since $H'S$ represents the altitude of a celestial object, S, which is on the meridian, and QS represents the declination, δ, of the object, we have at once the relation: $\phi = \delta + 90 -$ altitude. This is the method of determination of latitude most commonly used at sea, and it presents two fundamental difficulties to the navigator. The instant that the object is on the meridian must be accurately known and, also, the declination of the object must be observed. If both the Greenwich time and the longitude are known, the instant that the object should reach the meridian may be calculated in advance from the right ascension of the object; and the observation of altitude is taken at the predetermined instant. (Before chronometers came into use, it was necessary to watch the object very carefully and to record the maximum altitude attained by the object. If the object was the sun, the time that the maximum altitude was obtained became the local apparent noon, and was used by the navigating officer for setting the watch time for the ship.) If the observed object is a star, the declination may be immediately obtained from star catalogues; but if the sun, whose declination is changing rapidly, is the observed object, the Greenwich time of observation must be used to obtain the declination from the ephemeris.

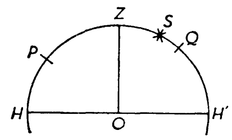

Fig. 2. Representation of the celestial sphere drawn in the plane of the local meridian of the point O.

Should the meridian observation be missed, because of cloud cover or for any other reason, the astronomical triangle may be solved to obtain the latitude if the local time of observation and the declination of the object are both known. If the object is observed very close to the meridian, and if the approximate latitude as well as the local time is known, the observation may be "reduced to the meridian" by tables. See also **Celestial Sphere and Astronomical Triangle.**

Modern navigational methods for determining latitude are discussed elsewhere. See **Navigation.**

A meridian altitude of an object is always effected by the correction for astronomical refraction, which is always subject to error unless the object observed is close to the zenith. For accurate determination of latitude for purposes of geodetic surveying, the zenith telescope is used. See also **Earth.**

LATITUDE (Geomagnetic). See **Geomagnetic Latitude.**

LATTICE COMPOUNDS. Chemical compounds formed between definite stoichiometric amounts of two molecular species which owe their stability to packing in the crystal lattice, and not to ordinary valence forces.

LATTICE CONSTANT. A length representing the size of the unit cell in a crystal lattice. In a cubic crystal, this is just the length of the side of the unit cell, but such a simple definition is not in general possible, and the lattice constant must be chosen according to the geometry of the structure in each case.

LATTICE DESIGNS. Lattice designs form a class of experimental designs enabling a large number of unrelated treatments to be compared in randomized blocks of a reasonable size. If there are n treatments where $n = p \times q$, the treatments are thought of as generated by the combinations of two pseudo-factors A and B, one at p, the other at q levels (see **Factorial Experiment**). Two types of replicates are then laid down, confounding the main effect of A in one type and the main effect of B in the other. The most useful case is that in which $p = q$, which gives rise to blocks of equal size.

LATTICE DIMENSIONS. According to the Bragg formula the spacing of the atomic planes can be deduced from the x-ray diffraction pattern and a knowledge of the x-ray wavelength, which can itself be measured by diffraction from a ruled grating.

LATTICE ENERGY OF CRYSTAL. The decrease in energy accompanying the process of bringing the ions, when separated from each other by an infinite distance, to the positions they occupy in the stable lattice. It is made up of contributions from the electrostatic forces between the ions, from the repulsive forces associated with the overlap of electron shells, from the van der Waals forces, and from the zero-point energy.

LATTICE (Mathematics). A set S of elements $a, b, \ldots,$ is *partially ordered* if a binary relation often denoted by the symbol $\leq$, which is reflexive, antisymmetric and transitive, is defined for certain of its elements. For example, let $a, b, \ldots$ denote the subsets of S, and let a stand in the given relation to b if the subset a is included in the subset b. A partially ordered set is a lattice if for any two elements a, b there exists an element c which is a least upper bound for a, b; that is, such that $a \leq c$, $b \leq c$ and if $a \leq e$, $b \leq e$, then $c \leq e$, and also an element d which is a greatest lower bound for a, b; that is, such that $d \leq a$, $d \leq b$ and if $f \leq a$, $f \leq b$ then $f \leq d$. These elements c and d are called the *join* (or *union*) and the *meet* (or *intersection*), respectively, of a and b, and are denoted by $c = a \cup b$ and $d = a \cap b$. The terms cup and cap are also used, and it is common to write $a + b$ for $a \cup b$ and ab or $a \times b$ for $a \cap b$. See **Boolean Algebra.**

LATTICE WATER. See **Hydrate.**

LAUAN TREE. See **Mahogany Trees.**

LAUNCH WINDOW. The postulated opening in the continuum of time or space, through which a spacecraft or missile must be launched in order to achieve a desired encounter, rendezvous, impact, or the like.

LAUREL FAMILY (*Lauraceae*). Approximately one thousand species make up this family of trees and shrubs. They are characterized by alternate, simple, often evergreen leaves, and by panicles or umbels of flowers with one-seeded drupes or berries. Some of the more familiar and economically important members of *Lauraceae* are described here. See accompanying table.

Avocado. This tree, sometimes called the alligator pear, is a native of the lowlands of tropical America, but has been extensively cultivated in tropical and subtropical regions. The tree was introduced into California and Florida many years ago. The avocado is now of considerable economic importance in California. The avocado tree is attractive, with large oval to elliptical leaves and small yellowish flowers. The large green-to-brown fruit varies in shape from nearly spherical to that of a pear. The fruit is very nutritious and is rich in oil. The flavor is quite subtle and often is garnished with salt, vinegar, or salad oil. The Guatemalan avocado has an oil content up to 25%. The tree can withstand temperatures as low as 25°F (−4°C.) without damage. The Mexican species is the hardiest and of excellent quality. It can withstand temperatures as low as 20°F (−6.7°C.) if not prolonged. California growers bud this variety, using patch-a-bud technique. In Florida, the side graft is commonly used. The pulp is about 69% water, 20% oil, and contains close to 2.5% protein. The Western Avocado of the West Indies has the tenderest fruit and is of a low oil content, ranging from 4 to 7%. The peel is smooth and purple. The tree cannot withstand temperatures below 28°F (−2.2°C.).

J. M. Haller (*American Forests* magazine, p. 29, May 1982) observes some of the unusual characteristics of the avocado. "Flowers, which in other species are certifiably and consistently male, female, or hermaphroditic, may on the avocado be male in the morning and female in the afternoon! Other species are rigorously grafted to prevent reversion to a primitive type bearing inferior fruit, but the avocado, though it may be and often is similarly grafted, will produce an astonishing variety of viable types from seed, most of them equal to any given grafted line and many of them superior. Other trees are either deciduous (leaf-shedding) or evergreen. The avocado manages both at the same time, shedding its leaves regularly each spring but not until the new season's crop is ready as a replacement (hence always green)."

Bay. This tree is native to the West Indies, but is found in France, Germany, and the coastal areas of the Americas. It is a small-to-medium-size tree, attaining a height of from 35 to 40 feet (10.5 to 12 meters), with a trunk of about 5 to 12 inches (12.7 to 30.5 centimeters) in diameter. The tree is related to the allspice and sassafras trees. The fruit is a berry. The oil from the fruit is yellow and aromatic and is the basis of bay rum. The bay tree sometimes is called the wax myrtle tree.

Cinnamon. The spice is obtained from a small tree native to Sri Lanka and India, where it is cultivated extensively. The tree grows to a height of from 25 to 40 feet (7.5 to 12 meters) and has shiny dark green, leathery leaves, small whitish flowers which have a rather disagreeable odor, and dark purple fruits. The bark of young twigs is smooth and somewhat mottled; in the older branches and the main stem, the bark becomes thick, rough and of little value. To insure the desideratum of many young branches, the limbs are severed so that many slender branches will form, a practice known as coppicing. From these slender stems, the bark is removed by lengthwise splitting and partial loosening from the stem. As it dries, it rolls back. It is then removed from the stem, the dry useless periderm scraped off, and the inner bark remaining allowed to dry completely. During drying, its color changes from pale yellow to deep brown. The tight rolls of dried bark are packed together in bundles, called pipes, and are ready for marketing as cinnamon. The bark contains considerable amounts of a powerful drug which in large doses is a dangerous poison. The principal use of cinnamon is as a spice for pastries. By distillation of cinnamon stems and leaves, oil of cinnamon is obtained, used in flavoring candy and in scenting soaps.

Laurel. The laurel tree grows along the coastal mountains and in the Sierra Nevada mountains of California—at an altitude of about 4,000 feet (1220 meters). The California laurel and Oregon myrtle are essentially the same tree. The tree attains a height of from 50 to 80 feet

RECORD LAUREL TREES IN THE UNITED STATES[1]

	Circumference[2]		Height		Spread		
Specimen	Inches	Centimeters	Feet	Meters	Feet	Meters	Location
American avocado (1991) (*Persea americana*)	168	427	60	18.3	75	22.9	California
California laurel (1978) (*Umbellularia californica*)	501	1273	88	26.8	70	21.3	Oregon
Loblolly bay (1983) (*Gordonia lasianthus*)	161	409	94	28.7	52	15.8	Florida
Mountain Laurel (1991) (*Kalmia latifolia*)	58	127	25	7.6	28	8.5	North Carolina
Redbay (1972) (*Persea borbonia*)	164	417	58	17.7	68	20.7	Georgia
Sassafras (1982) (*Sassafras albidum*)	253	643	76	23.3	69	21.0	Kentucky
Swampbay (1991) (*Persea borbonia var. pubescens*)	161	409	83	25.3	29	8.8	North Carolina
PAWPAWS							
Asimina triloba (1986)	92	214	60	18.8	30	9.1	Mississippi

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

(15 to 24 meters), with a trunk of 2 to 3 feet (0.6 to 0.9 meter) in diameter in mature trees. The branches are erect, long, and thick. The bark is thin, scaly, and dark brown. The leaves are about 2 to 5 inches (5 to 12.7 centimeters) in length and one-half to one or more inches (1.3 to 2.5 centimeters) wide. The underside is light green; the top side is leathery, glossy, and thick. The flowers are in clusters, pale yellow and small. The fruit of approximately 1 inch in length hangs in clusters of two and three. It is about the size of an olive and is a yellow-green color, containing one seed. Laurel wood, used for cabinet work, veneers, and garden tool handles, is fine-grained and hard and weighs approximately 40.5 pounds per cubic foot when dry (649 kilograms per cubic meter).

Pawpaw. This tree is found in the southern and midwestern states of the United States. It is related to the banana plant. The tree is small, usually grows wild, and is found in woodland areas. The fruit is exotic in appearance with a rich golden color. The fruit ranges from 3 to 5 inches (7.6 to 12.7 centimeters) in length and hangs from the tree in clusters. Pawpaw wood is spongy and weak and of no commercial value. The flower is purple and fragrant.

P. Stevenson presents an interesting portrait of the pawpaw in the March/April 1990 issue of *Amer. Forests,* page 46.

Sassafras. Frequently more of a shrub than a tree, the sassafras plant is found in the New England states, west to Wisconsin, and south to the Gulf coast. In many areas, the sassafras grows into a large tree, ranging up to 30 or 40 feet (9 to 12 meters) in height, with a trunk measuring from 8 inches to 2 feet (20 centimeters to 0.6 meter) in diameter. The record sassafras tree in the United States, as reported by The American Forestry Association, is located in Owensboro, Kentucky. See table.

All parts of the sassafras tree have a characteristic fragrance. The branch grows horizontal. The leaf is bright green and glossy and well known for its "mitten" shape. In the autumn, the tree turns a golden red and is quite showy.

The flower is small, yellow, and occurs in clusters. The flower appears before the leaf and is staminate with a six-lobed calyx, orange stalked glands, and nine stamens. The fruit is a dark blue, thin and fleshy berry of oblong shape. Although eaten readily by birds, the fruit is not enjoyed by humans. The bark is thick, scaly, and gray with longitudinal ridges. The wood is brittle and coarse grained and, although it resists moisture decay well, it is seldom considered of commercial value. Sassafras oil is sometimes used in soaps and toiletries.

H. Clepper elucidates further details of the sassafras tree in the March/April 1989 issue of *Amer. Forests,* page 33.

LAURIC ACID. Also called dodecanoic acid, formula $CH_3(CH_2)_{10}COOH$. A fatty acid that occurs in many vegetable oils and fats as the glyceride, especially in coconut oil and laurel oil. See also **Vegetable Oils (Edible).** Combustible. It takes the form of colorless needles at room temperature. Specific gravity 0.833; mp 44°C; bp 225°C (100 millimeters pressure). Insoluble in water; soluble in alcohol and ether. It is derived by the fractional distillation of coconut oil. Lauric acid is used in alkyd resins; wetting agents; soaps; detergents; cosmetics; insecticides; food additives.

LAVA. Molten material which has poured out on the surface of the earth and, due to relief of pressure, may have lost much of its original gas and water content during its relatively rapid consolidation. The term lava is used for both the liquid and the consolidated state of the igneous material. Lava may be erupted either by volcanoes or from fissures. The most extensive lava flows are fissure eruptions, such as the Columbia Plateau basalts in Oregon or the plateau basalts of the Deccan, India, which are derived from basic magma. Had this magma, either basic or acid, cooled slowly beneath the surface of the earth under great pressure and with all its original gases, the resulting rock would have had a coarser texture and somewhat different mineral content.

See also **Earth, Ocean,** and **Volcano.**

LAVAGE. See **Empyema.**

LAW OF AREAS. See **Kepler's Laws of Planetary Motion.**

LAW OF COSINES. See **Direction Cosine; Pythagorean Theorem.**

LAW OF LARGE NUMBERS. There are various laws of large numbers but the essential idea is exactly the same in each case. If the size of a sample is increased indefinitely or becomes very large, good sample estimates of population parameters will tend to concentrate more and more closely about the true value. Bernoulli's theorem is perhaps the simplest illustration of a law of large numbers.

Put another way, such laws state conditions under which random variables converge in probability to constants as some parameter n (usually a sample number) tends to infinity. *Strong* laws are concerned with showing that, for example, a variable x converges to a value μ with probability unity. Weak laws consider conditions under which the probability that $|x - \mu|$ is greater than some given ϵ, tends to zero.

LAWRENCIUM. Chemical element symbol Lr, at. no. 103, at. wt. 257 (mass number of known isotope), radioactive metal of the Actinide series, also one of the Transuranium elements. ^{103}Lw was identified in 1961 by A. Ghiorso, T. Sikkeland, A. Larsh, and R. Latimer at the University of California at Berkeley.

This method used to produce and identify lawrencium was similar to that used in the later, direct-counting experiments performed in connection with the production of nobelium at Berkeley. About 3 micrograms of a mixture of californium isotopes were bombarded with boron ions accelerated in the heavy-ion linear accelerator. The atoms of lawrencium recoiled from the target into an atmosphere of helium, where they were electrostatically collected on a copper conveyor tape. This tape was then periodically pulled into place before radiation detectors to measure the emission rate and the energy of the alpha particles being emitted. By this means, it was possible to identify the lawrencium isotope ^{257}Lr, with a half-life of 8 seconds. At present, because of the short half-life and the lack of a suitable daughter isotope, available in the case of nobelium, it has not been possible to perform a chemical identification.

Another isotope, ^{256}Lr, half-life about 45 seconds, was reported by the Soviet Union in 1965. It was produced by impact of oxygen atoms (^{18}O) on americium (^{243}Am). It decayed by alphaparticle emission and electron capture to form ^{252}Fm. See also **Chemical Elements.**

Lawrencium has been found to behave quite differently from dipositive nobelium and, in fact, it is comparable to the tripositive elements that appear earlier in the Actinide series.

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LAWSON CRITERION. See **Nuclear Reactor.**

LAWSONITE. This calcium aluminum silicate mineral, $CaAl_2(Si_2O_7)(OH)_2 \cdot H_2O$, is found as grains and veins within the metamorphic rocks, gneisses, and schists. It was found originally on the Tiburon Peninsula, San Francisco Bay, California, but also occurs in schistose rocks in France and New Caledonia. The mineral has a hardness of 7; specific gravity of 3.09. It is colorless, pale blue to bluish gray, translucent, with vitreous to greasy luster. The mineral crystallizes in the orthorhombic system.

LAXATIVE. See **Constipation.**

LAZULITE. This mineral crystallizes within the monoclinic system, a basic phosphate of magnesium and aluminum, $MgAl_2(OH)_2(PO_4)_2$. Ferrous iron can substitute for the magnesium and the isomorphous mineral scorzalite is the product. Usually occurs massive but acute pyramidal crystals are not uncommon. Color is azure-blue to bluish-green, usually translucent (rarely transparent), with vitreous luster. It has a hardness of 5.5–6, with specific gravity of 3–3.1.

Lazulite is a rare mineral found principally within high-grade metamorphic rocks. Notable world crystal occurrences are Salzburg, Austria; Syria; Hörnsjöberg, Sweden; Madagascar; Brazil; and Graves Mountain, Georgia. When transparent, the mineral can be cut into gem stones.

LAZURITE. The mineral lazurite or lapis lazuli has been used since ancient times for jewelry and other ornamental purposes. Ground to powder it forms the pigment ultramarine, now, however, largely superseded by artificial preparations. Lapis lazuli is a mixture of minerals, lazurite being the chief component. This mineral is isometric, and chemically a sodium, calcium, aluminum sulfo-chlorosilicate. A general formula is $(Na, Ca)_8(Al, Si)_{12}O_{24}(S, SO_4)$. Lapis lazuli has a hardness of 5–5.5; specific gravity, 2.4; color, various shades of blue; luster, vitreous to greasy; translucent to opaque. Localities are Afghanistan, Siberia, Chile, and California.

LAZY EYE. See **Vision and the Eye.**

L-DOPA. See **Parkinson's Disease.**

LEAD. Chemical element symbol Pb. at. no. 82, at. wt. 207.2, periodic table group 14, mp 327.5°C. bp 1740°C, density 11.35 g/cm^3 (20°C). Elemental lead has a face-centered cubic structure with an edge length of 4.950Å.

Lead is a white to bluish-gray metal, soft, malleable, and slightly ductile; tarnishes in air, forming a film of oxide, forms oxide scum upon heating the molten metal in air; soluble in dilute HNO_3; HCP or H_2SO_4 attack lead only slightly, the extent depending markedly upon the concentration and the temperature; slowly dissolves in H_2O and consequently the use of lead constitutes a health hazard due to its toxic effect; attacked by solutions of organic acids or sodium hydroxide. Lead is one of the four most largely produced and utilized metals, and considerable scrap metal is recovered. Used (1) in construction and apparatus where workability is demanded, and definite resistance to corrosion is supplied by the metal, (2) as a constituent of various alloys, especially solder, type metal, pewter, and fusible alloys, (3) for storage battery plates, (4) for shot and bullets, (5) as a protective coating for iron and steel.

Lead has four naturally occurring isotopes. In order of abundance, these are ^{208}Pb, ^{206}Pb, ^{207}Pb, and ^{204}Pb. There are ten unstable isotopes, 200–203, 205, and 209–214. See also **Radioactivity.** In terms of abundance, lead is scarcely represented in the earth's crust, the average composition of igneous rocks containing only 0.002% Pb by weight. In terms of cosmic abundance, an estimate made by Harold C. Urey in 1952, using silicon as a basis with the figure of 10,000, lead had an abundance figure of less than 0.02. In terms of presence in seawater, lead is 27th among the elements, with an estimated 14 tons per cubic mile (3 metric tons per cubic kilometer) of seawater. In this regard, it is comparable to tin, copper, arsenic, protactinium, and selenium.

The atomic weight varies because of natural variations in the isotopic composition of the element, caused by the various isotopes having different origins: ^{208}Pb is the end product of the thorium decay series, while ^{207}Pb and ^{206}Pb arise from uranium as end products of the actinium and radium series respectively. Lead-204 has no existing natural radioactive precursors. Electronic configuration $1s^22s;\&22p^63s^23p^6 3d^{10}4s^24p^64d^{10}4f^{14}5s^25p^65d^{10}6s^26p^2$. Ionic radius Ph^{2+} 1.18Å. Pb^{44} 0.70Å. Metallic Metallic radius 1.7502Å. Covalent radius (sp^3) 1.44Å. First ionization potential 7.415 eV; second, 14.97 eV. Oxidation potentials $Pb \rightarrow Pb^{2+} + 2e^-$, 0.126 V; $Pb^{2+} + 2H_2O \rightarrow PbO_2 + 4H^+ + 2e^-$, -1.456 V; $Pb + 2OH^- \rightarrow PbO + H_2O + 2e^-$, 0.576 V; $Pb + 3OH^- \rightarrow HPbO_2^- + H_2O + 2e^-$, 0.54 V. Other physical properties are given under **Chemical Elements.**

Lead is of interest as being the terminal product of radioactive decay. Thus while ordinary lead has the atomic weight 207.19 (being composed of 1.37% ^{204}Pb, 26.26% ^{206}Pb, 20.8% ^{207}Pb and 51.55% ^{208}Pb), the isotopic composition, and hence the atomic weight, varies somewhat in lead from meteorites, from deep-seated rocks and from uranium ores (the last being somewhat less dense, as would be expected from the fact that ^{206}Pb is the end-product of the uranium series). These variations in isotopic composition of lead permit of calculations of the age of the earth (and the meteorites).

Lead Melting Point as a Standard. Melting, defined as the equilibrium transition between crystalline and liquid states, is of large concern in the development of the physical and materials sciences. To date, some of the purest crystals of silicon, diamond, and other technologically important materials have been produced from melts. Studies of melts also are of significance in understanding the interiors of terrestrial plants and, in fact, of Earth. In research at the University of California (Berkeley), studies of the effects of high pressure on the fusion temperature of lead have been underway. The advantages of studying lead are outlined by the investigators as: (1) the melting temperature of lead at ambient pressures is low and well determined, (2) lead is highly compressible and therefore should show the effects of pressure, (3) the behavior of lead under pressure is relatively simple, involving only one known polymorphic transition (from face-centered cube to hexagonal close-packed crystal structure), and (4) shock-wave experiments have been carried out previously to document the compression of both crystalline and molten lead at simultaneously high pressures and temperatures.

Occurrence and Processing. Galena, PbS, is the source of over 95% of the lead currently produced. Bodies containing galena range from 3% to 30% lead. One of the most widely distributed sulfide minerals, galena frequently occurs along with sphalerite, ZnS. The lead-zinc ores processed usually contain recoverable quantities of copper, silver, antimony, and bismuth. Principal sources being worked are in Australia's Broken Hill area in New South Wales, the western United States, Canada, Mexico, Peru, former Yugoslav Republics, and the former Soviet Union. When groundwater reacts with galena, cerussite, $PbCO_3$, is formed; when galena is in contact with sulfate solutions generated by the oxidation of sulfide minerals, anglesite, $PbSO_4$, may be formed. See also **Anglesite; Cerussite;** and **Galena.**

In processing, the ore first is crushed, wet-ground, and classified to a point where it is at least 90% less than 200 mesh. Separation of the sulfide ore from the gangue is aided by flotation agents. The resulting concentrates contain from 45% to 60% lead, from zero to 15% zinc, and often a few ounces (~50 grams) of gold and up to 50 ounces (1.4 kilograms) of silver per ton. Copper content may be as much as 3%, arsenic, 0.4%, and antimony, 2%. The sulfur content (10 to 30%) is reduced by roasting in a Dwight-Lloyd sintering machine. This sulfur reduction is necessary because PbS is not reduced by carbon or carbon monoxide at blast-furnace temperatures. Once formed, the sinter, together with limestone and coke, is fed into a blast furnace. Further oxidation and electrolytic methods may be used to refine the lead. Lead is commercially produced to standards of very high purity. The minimum lead content permitted by specifications for Pig Lead (7 classifications) is 99.73%. Fully refined lead averaging 99.99% lead is obtainable. Large quantities are used for production of chemicals. At one time, primary uses for lead chemicals were in the production of paint pigments and lead tetraethyl gasoline additive.

Lead Metals and Alloys. Lead is soft and ductile and is readily worked by common methods, predominantly by rolling and extruding. Lead is easily formed and readily joined by welding (burning), or by soldering and can be bonded to steel, or used as a liner for steel, wood, concrete, and other materials. Lead is widely used in this manner because of its excellent resistant to atmospheric and soil corrosion, and attack by sulfuric and phosphoric acids. Lead generally does not resist the action of the organic acids, nor the oxidizing mineral acids, such as HNO_3. Lead is attacked by alkalies.

Due to its low melting point, pure lead will very gradually flow or creep at room temperature. Thus, lead sheeting used as a roofing material on old buildings will usually be thicker at the lower edge than at the upper edge. Other examples of creep occur under low sustained stresses due to the oil pressure in lead-covered power conducting cable, for example, or due to the weight in the case of a deep tank lined with sheet lead. To counter the effects of creep, lead containing 0.06% copper (*chemical lead* or *acid lead*) is preferred.

The addition of antimony in amounts up to 12% greatly improves the casting properties and increases the hardness very materially. These properties make possible the casting of intricately shaped antimonial lead storage-battery grids which, including the weight of the lead oxide paste applied to them, constitue the largest single use for the metal.

Tin and lead in various proportions form a highly useful series of alloys generally known as the soft solders which are used for joining copper, iron, nickel, lead, zinc and even glass. The solders can be applied by means of a soldering tool, by wiping, by hot-dipping, or by special machines as in the tin-can industry. Numerous compositions are used, the most popular of which are listed in the accompanying table.

Further additions of bismuth, cadmium, and antimony to the tin-lead alloys result in the low melting or "fusible" alloys widely used as safety devices, the melting points of which can be varied to suit a wide range of requirements. The type metals of the printing industry are lead-tin-antimony alloys having the requisite hardness and good casting properties needed for high-fidelity reproduction.

Babbitt metals (white-metal bearing alloys) are generally classified as either tin-base or lead-base. The true tin-base Babbitts contain only tin, antimony and copper, and have been used for many years. The practice of adding up to 25% lead to the tin Babbitts to reduce their cost is to be avoided since the net result is an expensive series of alloys with inferior properties to the inexpensive lead-base Babbitts. The lead-base bearing alloys of the older type usually contain lead, antimony and tin, and while not considered the equal of the tin-base alloys for severe service have been widely employed due to their low cost. The lead-base alloy containing arsenic has found extensive use and has come to the fore of this group since it has successfully met many automotive and other severe service requirements. All of these alloys render their most efficient service when used in the form of a thin lining bonded to a bronze or steel shell. See accompanying table.

Lead Eliminated from Free-cutting Alloys. Among the numerous efforts being made to eliminate lead from the environment, including the potable water plumbing systems, free-cutting copper alloys that contain no lead have been developed. As reported in late 1991, bismuth, as a replacement, has significant potential as a nontoxic alternative to lead to enhance the machinability of copper. When bismuth is used alone, however, the element embrittles copper because of its tendency to "set" grain boundaries. J. T. Plewes (see reference) ascribes this characteristic to the large difference in surface tension between copper and bismuth. It has been found that adding a third element in modest amounts removes this limitation of bismuth. Such elements include phosphorus, indium, and tin.

Chemistry of Lead. A number of oxides of lead are known, but not all are daltonide compounds. Thus, lead(I) oxide, Pb_2O, made by heating lead(II) oxalate, has been shown by x-ray analysis to be a mixture of the metal and lead(II) oxide, PbO. The latter is obtained by heating lead in air, which yields a yellow, rhombic material, which has a peculiar layer structure having each lead atom attached to four oxygen atoms all lying on the same side of it, forming a square pyramid with the lead at the apex. Each oxygen is surrounded tetrahedrally by four lead atoms. Another form of PbO, somewhat more stable and soluble in water, red in color, and tetrangonal in structure, may be obtained along with the yellow form by alkaline dehydration of $Pb(OH)_2$. PbO is amphiprotic, but only weakly acidic. Lead(IV) oxide, PbO_2, is obtained by action of chlorine on alkaline solutions of lead(II) oxide or acetate. The reaction is $Pb(OH)_3^- + ClO^- \rightarrow PbO_2 + Cl^- + OH^- + H_2O$. PbO_2 can also be produced on a lead or platinum anode by electrolysis in acidic solution. Like the lower elements of main group 4, lead(IV) forms tetrahedral bonds exhibiting sp^3 hybridization. In its relatively more stable salts, however, the $6s^2$ electrons are unused, and Pb^{2+} ions are formed by loss of the $6p^2$ electrons. These facts explain the marked difference between the essentially covalent character of many of the tetravalent compounds and the essentially electrovalent character of the divalent compounds, as well as the peculiar structure of PbO and many other Pb(II) compounds.

REPRESENTATIVE LEAD AND TIN ALLOYS

Name	Pb	Sn	Sb	Cu	Bi	Ag	Cd	Typical Application
Lead Alloys								
Chemical or acid lead	99.9			.06				Tank linings, coils, etc., power cable sheath.
Cable sheath	98.9		1.0					Telephone cable sheath.
Hard lead	96–92		4–8					Cast shapes, wrought sheet and pipe.
Battery grid metal	92–88	.25	8–12					Cast battery grids.
Solders								
Soft solder	50	50						General purposes, most popular solder.
Wiping solder	60	40						For wiping joints in cables, lead
	60	37.5	2.5					pipes, etc.
"Fine solder"	40	60						For making joints at low temperature.
Solder	95–97.5					5–2.5		High temperature solder.
Fusible Alloys								
Wood's metal	25	12.5			50		12.5	Melts in hot water at 154°F. Wets glass. Wide range of melting points possible with changes in composition for automatic sprinkler systems and other safety devices.
Matrix metal	28.5	14.5	9		48			For anchoring punches, etc., in jigs and fixtures. Expands on freezing.
Bending alloy	26.5	13.5			50		10	Filler for tubes, etc., during bending. Melts out in hot water.
Type Metals								
Electrotype	93	3	4					
Linotype	84	4	12					
Stereotype	80.5	5.75	13.75					
Monotype	76	8	16					Single type.
Tin Base Babbitts								
		89	7.5	3.5				General usage.
		83.3	8.3	8.3				Hard Babbitt.
Lead Base Babbitts								
	82.5	1.0	15	.5	1.0 As			General usage.
	80	5	15					General usage.
	75	10	15					General usage.

NOTES: Figures given in percent. Wood's metal melts at ~68°C in water.

The dioxide, PbO_2, has rutile structure, and the compound is a strong oxidizing agent. It is also amphiprotic, giving unstable lead(IV) salts with acids, and orthoplumbates, $M^{I}_4PbO_4$, or metaplumbates, $M^{I}_2PbO_3$, upon fusion with alkalies. Lead dioxide dissolves in aqueous alkali with formation of the ion $Pb(OH)_6^{2-}$, the alkali salts of which are isomorphous with the corresponding stannates and platinates. Lead sesuqioxide, Pb_2O_4, has been shown not to exist as a stable phase.

Lead orthoplumbate, Pb_2PbO_4, red lead, is similarly described as a salt, in this case an orthoplumbate of divalent lead, Pb_2PbO_4, because on treatment with nitric acid, two-thirds of the lead dissolves and one-third remains as PbO_2. It is prepared in the red form by atmospheric heating of PbO, and in a black form by reaction of PbO with pure oxygen. Red lead is formed of PbO_6 octahedra (with one common edge) linked by lead atoms covalently bonded to three oxygen atoms.

The lead dihalides are known for all four of the common halogens. They are not strictly ionic in the anhydrous state, but they dissolve in (hot) water to give Pb^{2+} ions, more or less hydrated. They are much less soluble in cold water. They also form complex compounds such as M_2PbCl_4, MPb_2Cl_5, M_4PbF_6, and $MPbF_3$, where M is an alkali metal. The compound formed, especially of the fluoroplumbates(II) depends somewhat on the alkali metal, some of which form/nondaltonide (berthollide) compounds. Of the lead tetrahalides, only PbF_4 and $PbCl_4$ are known, the fluoride being prepared by fluorination of PbF_2. The chloride, which easily loses chlorine, is made by careful acidification of a hexachloroplumbate(IV). $PbCl_4$ forms the complex compound ammonium hexachloroplumbate, $(NH_4)_2PbCl_6$, upon addition to its solution of solid ammonium chloride.

Lead(II) inorganic compounds and salts of organic acids are far more numerous than those of lead(IV), as is to be expected from the essentially covalent character of the latter. In addition to the oxides and halides already discussed, there are lead(II) compounds of essentially all of the common anions, including many basic compounds. Thus lead(II) chloride forms such basic compounds as $PbCl_2 \cdot Pb(OH)_2$, $PbCl_2 \cdot PbCl_2 \cdot 2PbO$, $PbCl_2 \cdot 3PbO$, and $PbCl_2 \cdot 7PbO$. In fact, a whole series of lead salts are derived from the hydroxide, some of which are double compounds, such as $PbX_2 \cdot 2Pb(OH)_2$ and some of which, of composition Pb(OH)X, have been shown to be dimeric of the general formula

$$\left[\mathrm{Pb}\left(\begin{matrix}\mathrm{HO} \\ \\ \mathrm{HO}\end{matrix}\quad\mathrm{Pb}\right)\right]X_2$$

Other lead compounds include the following:

Acetates. Lead acetate, "sugar of lead" $Pb(C_2H_3O_2)_2 \cdot 3H_2O$, white crystals, soluble, formed by reaction of lead oxide and acetic acid, and then crystallization. Used (1) to furnish a soluble lead salt, (2) as a mordant in dyeing and printing textiles, (3) as a paint and varnish drier, basic lead acetate, white crystals, soluble, formed by reaction of lead acetate solution and lead oxide, and then crystallization. Used as a coagulating, clarifying, and deacidifying agent for many organic solutions.

Arsenate. Lead arsenate, arsenate of lead $Pb_3(AsO_4)_2$, white precipitate, formed by reaction of soluble lead salt solution and sodium arsenate solution. Used as an insecticide. Banned or tightly controlled in some countries.

Azide. Lead azide PbN_6, white precipitate, formed by reaction of soluble lead salt solution and sodium azide solution (white solid, formed by reaction of sodamide $NaNH_2$ upon heating in nitrous oxide N_2O gas). Used as a detonator.

Borate. Lead borate $Pb(BO_2)_2$, white crystals, insoluble, by reaction of lead oxide and boric acid solution. Used in preparing special types of glass.

Carbonates. Lead carbonate $PbCO_3$, white precipitate, formed by reaction of soluble lead salt solution and sodium carbonate solution in the cold; basic lead carbonate, formed by reaction of (1) soluble lead salt solution and hot sodium carbonate solution, (2) lead sheets, carbon dioxide and acetic acid, and pigment, the quality depending largely upon the conditions of the reaction.

Chromates. Lead chromate, "chrome yellow" $PbCrO_4$, yellow precipitate, by reaction of soluble lead salt solution and sodium dichromate or chromate solution, melting point of lead chromate 844°C. Used as a pigment; basic lead chromate, red solid, insoluble, formed by heating lead chromate and sodium hydroxide solution.

Nitrates. Lead nitrate $Pb(NO_3)_2$, white crystals, soluble, formed by reaction of lead oxide and nitric acid, and then crystallization, decomposes on heating leaving lead oxide residue. Used to furnish a soluble lead salt; basic lead nitrate, formed by reaction of lead nitrate solution and lead oxide.

Oxalate. Lead oxalate PbC_2O_4, white precipitate, formed by reaction of soluble lead salt solution and ammonium oxalate solution, yields lead suboxide on heating at 300°C out of contact with air.

Phosphate. Lead phosphate $Pb_3(PO_4)_2$, white precipitate, by reaction of soluble lead salt solution and sodium phosphate solution.

Sulfates. Lead sulfate $PbSO_4$, white precipitate, formed by reaction of soluble lead salt solution and sulfuric acid or sodium sulfate solution; basic lead sulfate, "sublimed white lead," white solid, formed (1) by reaction of lead sulfate and lead hydroxide in water (slow reaction), (2) by roasting galenite in a current of air.

Sulfide. Lead sulfide PbS, brownish-black precipitate, formed by reaction of soluble lead salt solution and hydrogen sulfide or sodium or ammonium sulfide, soluble in dilute nitric acid.

In the great majority of organometallic compounds of lead, the metal is tetravalent and covalently bonded, although the organolead group includes many compounds with both organic radicals and halogen atoms attached to Pb which are not to be described merely as covalent compounds. More than five hundred organometallic compounds of lead have been reported, many of which are named as substituted plumbanes, although PbH_4 is not a starting point in their production. Tetraethyl lead, $Pb(C_2H_5)_4$, is made from a sodium-lead alloy and ethyl chloride.

Like carbon and silicon, and to a lesser extent, germanium and tin, lead forms binary compounds with metals, such as Na_4Pb_7 and Na_4Pb_9. These materials are essentially salt-like, and contain polyplumbide anions. They are of theoretical interest, because they are intermediate in character between stoichiometric compounds (daltonide compounds) and intermediate phases. The two compounds cited dissolve in liquid ammonia, electrolyze in such solutions to give the metals, and apparently form ions such as $[Pb_7]^{4-}$ and $[Pb_9]^{4-}$ which readily form amine complexes.

Lead in Biological Systems—Toxicity

Lead has been identified as a biological system deterrent for decades. It is only recently, however, that studies pertaining to low-dosage exposures of lead have been published despite the fact that probably millions of words have appeared in various publications on the overall topic of lead poisoning.

From a qualitative standpoint, exposure to lead results in a clinical picture of hypertensive encephalopathy, neuropathy, and hemolytic anemia characterized by coarse basophilic stippling in red blood cells. The mechanism of lead's action on human tissue is complex. For one thing, lead blocks heme synthesis. This leads to a build-up of red blood cell protoporphyrin. Lead interferes with cell metabolism by causing a deficiency of pyrimidine 5'-nucleotidase. Lead attacks erythrocyte membrane phospholipids with resultant loss of potassium and interference with the sodium-potassium balance. Diagnosing lead poisoning may involve a determination of the free erythrocyte proptoporphyrin level as well as determination of blood and urine levels. Once confirmed, further exposure to lead must be stopped immediately. Chelating compounds, such as $CaNa_2EDTA$, may be administered intravenously over an 8-hour period for several days. This may be followed by treatment with oral penicillamine for several days.

Lead poisoning can lead to chronic renal failure. In its effect on kidney function, lead acts much like cadmium. Chronic exposure to or ingestion of practically any heavy metal, such as lead, is the most common path to polyneuropathy. Where effects of heavy metals on the peripheral nervous system are suspected, many physicians will require testing for metal in hair, fingernails, serum, and urine of the patient. Habitual sniffing of leaded gasolines can lead to lead poisoning. Robinson, Scientists have compared the lead concentration in the diets of present Americans (0.2 part per million) with the diets of prehistoric peoples (estimated to be less than 0.002 part per million). Some investigators believe that the presence of "natural" lead contamination has been grossly overestimated and that what has appeared to be natural has been the result mainly of a gradual build-up of lead pollution in the air derived from anthropogenic sources. The principal sources of atmospheric lead contamination include (1) natural sources, such as wind-blow volcanic dust, sea spray, forest foliage, and volcanic sulfur compounds; and (2) antropogenic sources, such as lead alkyls (present in fuels), iron smelting, lead smelting, zinc and copper smelting, and the burning of coal. Much remains by way of research into the sources of lead contamination, including the contributions of atmospheric pollution, of food containers, and of food processing equipment.

In 1990, H. I. Needleman and co-researchers (University of Pittsburgh, Boston University, and Harvard University) reported their findings on the long-term effects of exposure to low doses of lead in children. An abstract of the report is as follows:

> To determine whether the effects of low-level lead exposure persist, we reexamined 132 of 270 young adults who had initially been studied as primary school-children in 1975 through 1978. In the earlier study, neurobehavioral functioning was found to be inversely related to dentin lead levels. As compared with those we restudied, the other 138 subjects had had somewhat higher lead levels on earlier analysis, as well as significantly lower IQ scores and poorer teachers' ratings of classroom behavior.
>
> When the 132 subjects were reexamined in 1988, impairment in neurobehavioral function was still found to be related to the lead content of teeth shed at the ages of six and seven. The young people with dentin lead levels >20 ppm had a markedly higher risk of dropping out of high school (adjusted odds ratio, 7.4; 95 percent confidence interval, 1.4 to 40.7) and of having a reading disability (odds ratio, 5.8; 95 percent confidence interval, 1.7 to 19.7) as compared with those with dentin lead levels <10 ppm. Higher lead levels in childhood were also significantly associated with lower class standing in high school, increased absenteeism, lower vocabulary and grammatical-reasoning scores, poorer hand—eye coordination, longer reaction times, and slower finger tapping. No significant associations were found with the results of 10 other tests of neurobehavioral functioning. Lead levels were inversely related to self-reports of minor delinquent activity.
>
> We conclude that exposure to lead in childhood is associated with deficits in central nervous system functioning that persist into young adulthood.

An interesting professional critique of the Needleman report is summarized by J. Palca (reference listed).

The lead elimination and clean-up problem has numerous parallels with the asbestos pollutive problem. The main problem is not one of finding substitues for these substances, because lead-free paints, for example, have been available for several years, just as substitute insulating materials for asbestos have been found. As pointed out in an excellent article by Pollack (reference listed), the problem (or dilemma) lies with the cleanup of old structures that have such materials installed and to what limits must one go, within the limitations of financial resources, to remove such materials and, once removed, how to dispose of them safely. While removing all lead-painted surfaces from schools, for example, ultimately can provide assurance that children will not be exposed to the long-term effects of lead, a great deal of new exposure to workmen and the immediate neighborhoods of such removal projects can occur. Modern technology has designed equipment to protect the safety of restoration or demolition crews, but there remains the enforcement of their using such equipment. One crux of the problem is that posed by the apparent effects of lead in very low dosages. Such pollutants of low concentration, as may be typified by the creation of dust and windborne aerosols, definitely exacerbates the problem. Obviously, the problem becomes one more hampered by socioeconomic measures than by technology.

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LEAD-ACID BATTERY. See **Battery; Electric Car.**

LEAD GLASSES. See **Glass.**

LEAD SCREW. The screw which controls the longitudinal motion of a tool on a lathe or other machine tool. Also, the screw which drives the cutting head across a recording disk in the initial recording process.

LEAD SULFIDE. See **Galena.**

LEAF. The food-manufacturing organ of a plant. Typically, leaves consist of a broad thin blade borne on a slender stalk. The important chemistry taking place within a leaf is described under **Photosynthesis.**

The leaf orginates as a small protuberance from the surface of the growing tip of the stem. Numerous divisions of the cells of this protuberance produce a structure from five to eight cells thick. Many of these leaf primordia are borne together on the stem tip, and, together with any protecting scales which may cover them, form the buds of the stem. At first, all the cells of these leaf primordia are alike. Very early in their existence, however, certain cells become distinct by their somewhat elongated shape. These cells are the beginnings of the vascular elements. Cells divisions continue in these small bodies until there are present in the bud recognizable but very small leaves which are folded in various ways. In woody plants, this development takes place in the year previous to that in which the leaf will unfold. With the advent of the new growing season, growth of the many minute cells of these tiny leaves is very rapid, so that within a few days' time the leaf has unfolded and grown to its mature size. During this enlargement, many changes have taken place in the cells of the leaf.

The mature leaf is commonly composed of two distinct parts, the broadly expanded, thin green blade, and the petiole or stalk which supports it and connects it with the stem. In many plants there is formed at the base of the petioles a pair of outgrowths called stipules, which in some plants may take the form of a complete sheath. See Fig. 1. This sheath is well developed in members of the Umbelliferae. Sometimes the petiole is completely lacking, the blade being attached directly to the stem; leaves of this kind are called sessile leaves. Less frequently the blade of the leaf is lacking, the petiole being expanded into a flattened object looking much like a blade. Certain Australian trees, species of *Acacia* and *Eucalyptus*, exhibit this peculiarity. Leaves of such plants often show progressive changes from those having well-developed blades to those in which the blade is completely lacking, showing clearly that the flattened portion present is a modified petiole. Such

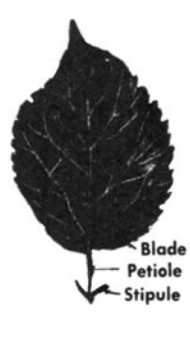


Fig. 1. Leaf of apple, illustrating all parts—blade, petiole, and stipules.

flattened petioles are not uncommon, but usually the blade is present, as is the case in the lemon tree. Leaves may be deciduous, falling off at the end of a single growing season, or evergreen and persistent through several seasons. In nearly all cases the leaf fall is brought about by the development of a definite abscission layer. In many plants, such a layer is formed not only at the base of the petiole, but also at the point where the petiole joins the blade.

The shape of the blade is extremely varied, ranging from very slender linear leaves to those which are broader than they are long. The margin of the leaf may be entire, that is, without indentations of any sort, or toothed or lobed in various ways, until some are incised nearly to the midrib. If the leaf is completely divided into separate segments, it is said to be a compound leaf, in contrast with the undivided leaves, which are simple leaves, no matter how deeply they may be lobed. If the sections of a compound leaf all come from a common point, the leaf is said to be palmately compound; if they are borne along a central axis, the leaf is pinnately compound. While such infinite variations do exist, the leaves of any single species of plant are recognizably constant in shape.

The blade of the leaf is supported by a framework of veins which are also very characteristically arranged. In many leaves, especially in dicotyledons, one vein, usually extending through the center of the blade, is more prominent than the others. This is called the main vein or midrib. The others are lateral veins. In most dicotyledons, the veins branch abundantly to form an intricately anastomosing network, which reaches all parts of the leaf. In most monocotyledons, the midrib and lateral veins extend in parallel lines from base to apex of the leaf. Between these, many minute veinlets exist, too small to be readily seen, reaching all parts of the leaf.

The cellular structure in leaves is very constant. See Fig. 2. Covering the entire surface of the leaf is the epidermis, a layer of tabular cells. On the upper surface of the leaf, the epidermal cells are frequently covered with a layer of cutin, a waxy substance which is impervious to water and so greatly reduces the loss of water by evaporation from the leaf surface. Epidermal cells contain a scant peripheral cytoplasm, and a large central vacuole full of cellsap. Usually, there are no chloroplasts present in the epidermal cells. The cells of the epidermis of the lower surface are similar to those of the upper, but with a less evident cuticle. In the epidermis of the leaf, particularly that of the lower surface, there are many minute openings, called stomata, which permit a ready exchange of gases between the interior of the leaf and the external air. Each stoma is surrounded by a pair of guard cells containing chloroplasts. These cells close the stoma by collapsing and open it by expanding. All cells occurring between the upper and lower epidermal layers are called mesophyll cells. See Fig. 3. Beneath the upper epidermis the mesophyll cells form a very distinct layer, called the palisade mesophyll. These are elongated cells with their long axis perpendicular to the surface of the leaf. They contain large numbers of chloroplasts. In them, furthermore, active photosynthesis takes place. Occupying all the rest of the leaf is a loose tissue composed of irregularly arranged rounded cells known as the spongy mesophyll. Numerous intercellular spaces separate these cells from one another. Ramifying through the leaf just below the palisade cells are the veins. Each vein is composed of three types of cells. Some of them are thick-walled xylem cells which carry water and dissolved mineral matter to all parts of the leaf. Others are phloem cells which carry food away from the green cells of the leaf where they are elaborated. The xylem cells are towards the top of the leaf, the phloem cells towards the bottom. Outside these and often forming a conspicuous tissue are the masses of fibers, or collenchyma, thick-walled cells which give support to the leaf.

Leaves are often greatly modified. See Fig. 4. In many plants, they become greatly enlarged and fleshy, and serve as organs of storage of water and food. Many rock garden plants, such as species of *Sedum*,

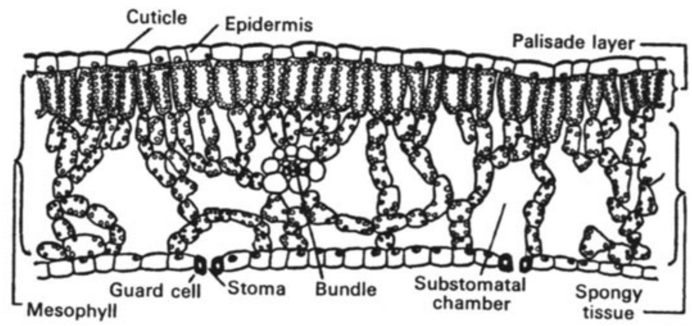


Fig. 3. Cross section of an apple leaf.

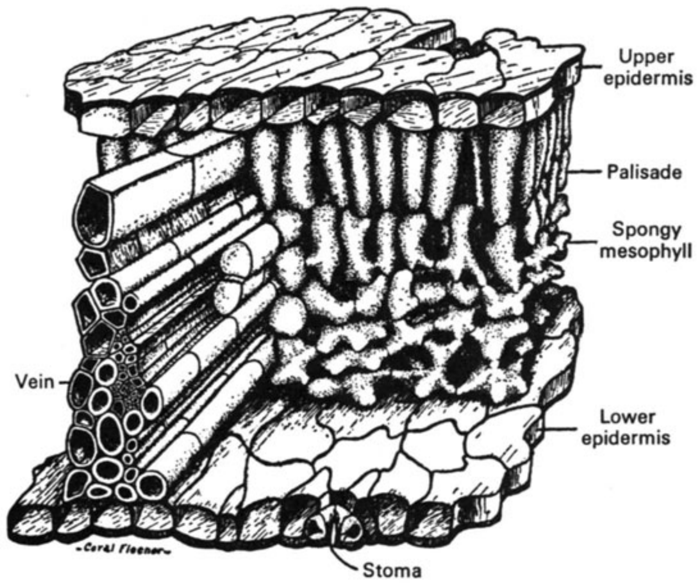


Fig. 2. A portion of the blade of a leaf cut so as to show the internal structure. The cell contents are not shown.

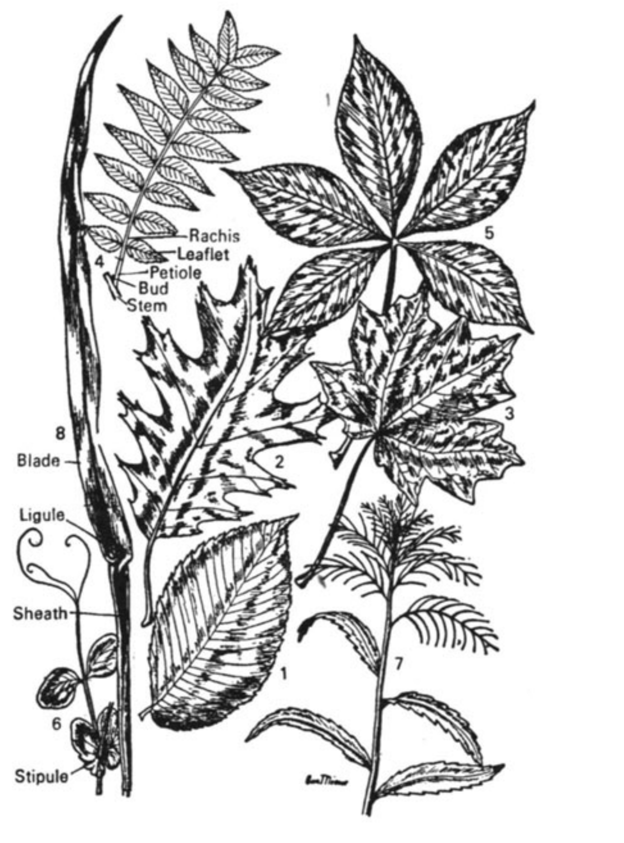


Fig. 4. Types of leaves: (1, 2) elm leaf and oak leaf, both pinnately netted veined; (3) maple leaf, palmately netted veined; (4) black walnut leaf, pinnately compound; (5) buckeye leaf, palmately compound; (6) a pea leaf, with stipules, tendrils, and two unmodified leaflets; (7) portion of a plant of the water mermaid, *Proserpinaca*, with upper leaves modified by immersion in water; (8) grass leaf.

have leaves of this type. Of similar nature are the scale-like leaves which form the greater part of many bulbs, such as those of many lilies. The common onion is composed of the closely enwrapped bases of leaves, swollen with food material. In other plants, modification of the leaves becomes extreme, as, for example, in the pitcher plants and bladderworts. (See **Insectivorous Plants.**)

In other plants, such as the common barberry, the leaf is reduced to sharp-pointed branched spines; in many cases all gradations between these spines and typical leaves may be found on a single branch. In some plants, as the Locust, *Robinia pseudoacacia*, only the stipules are modified to short sharp spines. Many plants have leaves modified into tendrils, slender thread-like objects which twine tightly around any suitable object with which they may come in contact. Sometimes only the tip of the blade functions in this way, and sometimes only the stipules are thus modified, as in the Carrion flower, *Smilax herbacea*. Many plants of the legume family have pinnately compound leaves, some of the segments of which are changed into tendrils. Weirdest of all are the leaves of species of *Nepenthes*, one of the pitcher plants. (See article on **Insectivorous Plants,** where this leaf is described.)

In a few plants, the leaf becomes a vegetative reproductive body, having in the notches of its margin, at its tip, or less commonly on its surface, groups of meristematic cells which, when the leaf is mature, give rise to tiny plants which remain attached to the parent leaf for some time. Among the plants in which reproduction of this type occurs are species of *Bryophyllum* and *Kalanchoë*.

The principal function of the leaf is to carry on photosynthesis. To do this, the leaf must receive adequate light. Leaves are not distributed haphazardly on the stem, but in a very definite way which assures them the maximum of light. In many plants, the leaves are in pairs on opposite sides of the stem. Each successive pair usually grows out at right angles to the pair beneath it, thus preventing overshadowing. Leaves may occur in whorls, in which case there will be three or more leaves growing from each node of the stem. In many plants the leaves are alternate, each node bearing a single leaf. In every case, alternate leaves arise from the stem in such a way that a line passing around the stem and through the junction of the petiole with the stem forms a regular spiral. Examination of this spiral shows that the leaves are distributed on it in a very exact mathematical arrangement. In the simplest case, the leaves are in two longitudinal rows along the stem, every third leaf being directly above the first; in the next arrangement there are three longitudinal rows, the fourth leaf of the spiral being above the first. Other more complicated arrangements are found. The arrangement of leaves on a stem is called phyllotaxy.

Sometimes the exact arrangement is more or less obscured by twisting of the stem during growth. The leaves themselves turn considerably during their growth, petioles twisting to one side or the other, or elongating unequally, in such a way as to bring the blade into a position to receive the most favorable light.

LEAF HOPPER (*Insecta, Homoptera*). Any insect of the large family *Cicadellidae* or *Jassidae*. They are small to moderate jumping insects which often come freely to light.

Many species are of economic importance and, since they have sucking mouths, they must be attacked with contact poisons such as nicotine sulfate or kerosene emulsion. These sprays are effective against the tender immature insects.

The body of the insect is slender and long, with a round head. The four legs are stout and strong. A hairlike antenna is below each eye. Sometimes they are called "dodgers" because of their habit of dodging around various objects to escape attention.

The leaf hopper exudes a sweet liquid from its abdomen in a fashion similar to the aphid.

In Australia, the *Eurymela* group lives on eucalyptus leaves. These insects are attended by ants, much as the aphids. The potato and apple leaf hopper is the *Empoasca fabae*. The species *Eutettix tenellus* is a pest to gardens. It inserts a virus with its tiny mouth parts, causing leaves to curl. The species *Nephotettix* is a pest to rice fields and has caused much damage in India. The insect stunts the growth of the plant by robbing juices and causing wilting.

Other leaf hopper species include the *apple leaf hopper* (*Empoasca mali*). The adults are of several colors, ranging from green to brown and yellow; or they may be striped. The insects are wedge-shaped and attain a length of about $\frac{1}{8}$ inch (3 millimeters). The nymphs are similar to the adults, but smaller, and crawl sideways like crabs. Action of the insects causes leaves to curl and turn yellow or reddish brown. The apple leaf hopper is found throughout the United States.

Young trees are the most seriously infested by leaf hoppers. The insects usually attack the underside of the leaves. Control should be directed toward the young nymphs; adult leaf hoppers often escape by flying away when disturbed. To control young trees infested by leaf hoppers, the tips of affected branches can be dipped into a container of soap solution, using about 1 pound of soap per 8 gallons of water (about 250 grams of soap per 15 liters of water). Dipping, which kills some of the young leaf hoppers, should be done in the latter part of June and again one month later. This is the period when the maximum number of nymphs will be found on the trees. Many adult leaf hoppers can be captured as they fly away by placing a shield covered with a sticky substance close to the tree.

Four generations of leaf hoppers are produced each year. The eggs, laid in blisters under the bark of the tree, winter over. Eggs laid in the summer are placed in leaf veins and petioles.

The *grape-vine leaf hopper* (*Typhlocyba comes*) is a small yellow-colored insect, sometimes mistakenly called thrips. This pest is most prevalent in the western United States. It sucks sap from the underside of leaves, causing the leaves to become brown and brittle. In treating for this hopper, care must be taken to thoroughly reach the underside of the foliage.

LEAF INSECT (*Insecta, Orthoptera*). Large insects of the Old World tropics related to the walking-stick insects. They have leaf-like wings and in some species the body and legs are extended in flat processes which also resemble leaves.

LEAF MINER (*Insecta, Lepidoptera*). Larval insects which work in the soft tissue of leaves between the upper and lower epidermis. They are necessarily small and are sometimes able to complete their development on a very small part of the food available in a single leaf. The burrow or mine shows as a brownish or transparent patch in the leaf and its form is characteristic of the insect making it. The larvae of many of the smallest moths and of some sawflies are leaf miners.

LEAF ROLLER (*Insecta, Lepidoptera*). Also sometimes referred to as the *leaf tyer* or *leaf sewer*, these small moths are usually members of the family *Tortricidae*. Their descriptive name derives from their practice (in caterpillar stage) of rolling all or part of a leaf into a cylindrically-shaped case, tying the case with natural gum threads, and then lining the case with silk and thus forming a cocoon wherein the insect transforms into the pupa stage. Researchers have observed that several larvae may work cooperatively to form a common "nest." There are several species, each of which builds a characteristic nest.

The adults of the *apple leaf roller* (*Archips argyrospila*) are brown moths with light markings on the wings and having a wingspan of about $\frac{3}{4}$ inch (18–19 millimeters). The larvae are from pale yellow to a dirty green in color, with brown or black heads, and ranging up to $\frac{3}{4}$ inch (18–19 millimeters) in length. Light yellow, green, or grayish eggs are laid on branches in masses of 10–19. The red-banded leaf roller has a broad, reddish-brown band across the wings. The larvae feed on buds, fruit, and leaves. The leaves are webbed together to form a tent or cocoon as previously described. The larvae eat irregular holes in leaves and fruit.

Distribution is throughout the United States. The red-banded leaf roller is confined to the eastern United States and ranges as far west as the Mississippi Valley.

A number of parasites and predators attack the leaf rollers. Toads eat many caterpillars that drop from the trees; birds also prey upon the caterpillars. Since the insect overwinters in the egg stage and deposits its eggs on the twigs and bark of the tree, it is possible to control the first brood by spraying with a dormant fruit tree oil spray. Also, the folded leaves can be pinched by hand to destroy the caterpillars inside. Debris should be burned after picking.

The *avocado leaf roller* (*Amorbia emipratella*) is a yellowish-green caterpillar with a pinkish-brown stripe approximately an inch (2.5 centimeters) long when fully mature. The insect rolls the leaves and eats small holes into the fruit, making it unmarketable.

The *strawberry leaf roller* (*Ancylis comptana*) has similar habits, with the larva, usually less than $\frac{1}{2}$ inch (12 millimeters) long, feeding and folding the leaves. The insect produces two broods per year.

LEAKAGE CURRENT. This is the current which flows or "leaks" along the surface or through the body of an insulator. Except under abnormal conditions such as dirty or moist surfaces or in electronic circuits having very minute currents the leakage is usually negligible.

LEAKAGE REACTANCE. This is the inductive reactance caused by the flux which links only one coil of a transformer. The useful flux, of course, links both windings and is the medium of transfer of energy between them. Leakage reactance is one of the major internal impedance components of the transformer.

LEAK DETECTION. See **Mass Spectrometry.**

LEAPING MAMMALS. See **Rabbits and Hares.**

LEARNING DIFFICUTY. See **Dyslexia.**

LEAST ENERGY PRINCIPLE. A principle relating to stable equilibrium, and having very wide application. If a system is in stable, equilibrium, any slight change in its condition or configuration requiring the performance of work will put it out of equilibrium, so that, if the system is now left to itself, it will return to its former state and in so doing will give up the energy imparted when it was disturbed. Consider, for example, a block of wood floating in a pail of water. If the block is lifted slightly, work is done and the center of mass of the wood-water system as a whole is raised, so that it now has more potential energy. The same would be true if the block were pushed a little farther into the water. In either case, when the block is released, it resumes its former level and the potential energy of the system diminishes to its former minimum value. This illustrates the general principle, which is that a system is in stable equilibrium only under those conditions for which its potential energy is at a minimum.

The principle of least energy is one aspect of the principle of virtual work.

LEAST SQUARES. Suppose that it is required to fit an equation of functional form $y = f(x_1, x_2, \ldots, x_p)$ to a series of observations on y and the x's. If the number n of observations exceeds the number of constants in the functional form, no exact fit is, in general, possible. The method of least squares determines a good fit by minimizing the sum of squares of residuals $\Sigma\ (y - f)^2$ over the observations.

The method is clearly reasonable in all cases. In some it has optimal properties; for example, if f is linear and the model is of the type $y = \beta_0 + \beta_1 x_1 + \cdots + \beta_p x_p + \epsilon$ where ϵ is a random residual normally distributed, the estimators of the β's derived by least squares are unbiased and have minimum variance. See **Regression.**

LEATHER-JACKET. 1 *Insecta, Diptera*. The tough-skinned larvae of some species of crane flies. They live in the ground in pastures, hay fields, and grain fields and are sometimes serious pests. Since they come to the surface at night they can be destroyed by the use of poison baits. 2 *Pisces*. File fish related to trigger fish. Coastal; frequenting reefs and rocks, mostly poisonous but two or three Australian species said to be good as food.

LEAVENING AGENTS. The generation of carbon dioxide for use as dough leavening is produced by reacting sodium carbonate (baking soda) with one of several leavening acids. In the case of an acidic phosphate salt (with two replaceable hydrogen atoms), the reaction is:

$$MH_2PO_4 + 2NaHCO_3 \rightarrow MNa_2PO_4 + 2H_2O + 2CO_2$$

where M can be a hydrogen or an alkali metal ion. Claims for use of acidic phosphate salts, in addition to formation of carbon dioxide, are the buffering effects for providing an optimal pH for the baked product, as well as interactions with protein constituents of flour, with resulting optimal elastic and viscosity properties of the dough batter.

Other leavening acids used in modern bakeries include sodium aluminum sulfate, $Na_2SO_4 \cdot Al_2(SO_4)_3$; sodium aluminum phosphate hydrate (and anhydrous); potassium acid tartrate, $KHC_4H_4O_6$ (cream of tartar); and glucono-delta-lactone. The baker is concerned with (1) *dough rate of reaction* (DRR), a measure of the rate at which the leavening acid reacts with the baking soda during both the mixing stage and the holding period after mixing (bench action); and (2) *neutralizing value* or neutralizing strength, i.e., the weight of leavening acid required to neutralize a given weight of sodium bicarbonate. This value is used to compute the amount of leavening acid required to yield the needed amounts of leavening gas as well as its effect upon the pH of the baked goods.

Properties of the principal leavening acids are given in the accompanying table which shows the most appropriate baking applications for each.

PROPERTIES OF LEAVENING ACIDS

Chemical Name and Formula	Abbreviation	Relative Speed at Room Temperature	Neutralizing Value[1]
Sodium aluminum phosphate (anhydrous)	SALP	Medium	110
Sodium aluminum sulfate, $Na_2SO_4 \cdot Al_2(SO_4)_3$	SAS	Slow	100
Monocalcium phosphate (anhydrous), $CaH_4(PO_4)$	MCP	Slow	83
Monocalcium phosphate (monohydrate), $CaH_4(PO_4) \cdot H_2O$	$MCP \cdot H_2O$	Quite fast	80
Sodium acid pyrophosphate, $Na_2H_2P_2O_7$	SAPP	Medium	72
Glucono-delta-lactone, $C_6H_{10}O_6$	GDL	Slow	55
Potassium acid tartrate, $KHC_4H_4O_6$	—	Medium to fast	50
Dicalcium phosphate dihydrate, $CaHPO_4 \cdot 2H_2O$	DCP	Very slow	33
Sodium aluminum phosphate hydrate	$SALP \cdot H_2O$	Slow	100

[1]Values in this column indicate the parts of sodium bicarbonate that will be neutralized by 100 parts of the leavening acid under nominal conditions. Values vary with composition of dough.

Baking powders, as prepared for the home baker and for use in premixes, usually incorporate, along with sodium bicarbonate, one of the following leavening acids: (1) potassium hydrogen tartrate (2 parts for 1 part sodium bicarbonate; (2) tartaric acid (infrequent), 1 part; (3) calcium hydrogen phosphate (crystallized), 1.5 parts; (4) sodium aluminum sulfate or ammonium aluminum sulfate, 1.8 parts. With 7 parts by weight of this finely powdered mixture, there is usually mixed about 3 parts by weight of starch to diminish the effects of moisture in storage. In some cases, dry powdered egg albumin is added to decrease the loss of carbon dioxide upon wetting the flour and baking powder mixture when used. For some purposes, ammonium carbonate can be used alone, since upon heating this material furnishes both ammonia and carbon dioxide gases to make the product light. These gases escape from the product during the baking process. In selecting a baking powder, one must keep in mind the speed with which the components react at room temperature: alum-containing baking powders act slowly; phosphate baking powders have a medium speed; and tartrate baking powders act quickly to produce carbon dioxide. Hence, when using the latter type, it is necessary to bake quickly after mixing to eliminate the loss of too much gas.

Within the last several years, advantage has been taken of mixing different leavening acids in premixes and household baking powders. Because the use of emulsifiers in most cake mixes reduces the need for early leavening action, it is common practice to use combinations of slow-acting leavening acids which retain much of their leavening reaction for the baking stage. In mixes, the leavening process must be regarded as a system because, in addition to gas generation, the leavening system controls the pH of the finished product and thus affects crumb and crust color, the intensity of flavor, as well as other properties. For various cakes, the optimum pH values are: white cakes, 6.9–7.2; yellow cakes, 7.2–7.5; chocolate or devil's food cakes, 7.1–8.0. Monocalcium phosphate (anhydrous) and sodium aluminum phosphate are frequently used together in white and yellow cake mixes; monocalcium phosphate and sodium acid pyrophosphate or dicalcium phosphate dihydrate are used in chocolate cake mixes. Generally, the combination will be comprised of 10–20% fast-acting leavening acid and 80–90% slow-acting leavening acid.

For pancake and waffle mixes, a common blend of leavening acids is 20–30% monocalcium phosphate monohydrate or monocalcium phosphate (anhydrous), combined with 70–80% sodium aluminum phosphate. A batter of this type can be prepared several hours in advance if retained under refrigeration. It has been observed that such a batter will sour before a serious loss of leavening power occurs.

Prepared biscuit mixes made of flour, shortening, and salt usually contain 30–50% monocalcium phosphate (anhydrous) and 50–70% sodium aluminum phosphate or sodium acid pyrophosphate. Self-rising flours and corn meals usually contain flour or corn meal, salt, soda, and leavening acid. Usually used in these products are combinations of sodium aluminum phosphate and monocalcium phosphate (anhydrous).

Refrigerated doughs available for preparation of biscuits, dinner rolls, and various sweet rolls, usually contain flour, water, shortening, nonfat milk solids (or dried whey solids), sugar (or corn sugar), salt, soda, and a leavening acid. Long-term refrigerated storage requires that only slow-acting leavening acids be used, frequently the sodium acid pyrophosphates. The latter have the disadvantage of possibly producing orthophosphates under certain conditions. The orthophosphates have a rather disagreeable, astringent flavor.

Unleavened Products. The principal unleavened bakery product is pie crust, which is low in moisture and high in fat content. The ingredients and method of preparation prevent the formation of a continuous gluten network through the dough mass. The porosity associated with leavened products is not desirable because the crust literally acts as a container and requires some strength.

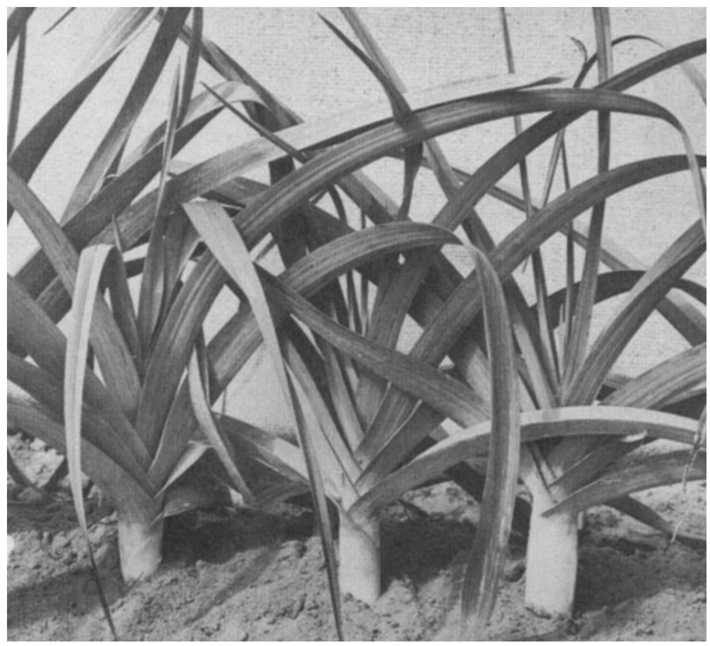
LE CHÂTELIER'S PRINCIPLE. Let us perturb a system which is initially in stable equilibrium to a neighboring nonequilibrium state. Since the initial equilibrium is supposed to be stable, the system will return to an equilibrium state.

Theorems governing the behavior of perturbed systems are often known as *theorems of constraint* or *theorems of moderation*. The best known thermodynamic theorem of moderation is that of Le Châtelier-Braun which in the form stated by Le Châtelier is:

"Any system in chemical equilibrium undergoes, as a result of a variation in one of the factors governing the equilibrium, a *compensating* change in a direction such that, had this change occurred alone it would have produced a variation of the factor considered in the *opposite direction*."

However, this principle suffers from a number of important exceptions. It is therefore preferable to study the "moderation" starting from the usual thermodynamic formalism without invoking a special principle.

LEEK. Of the family *Amaryllidaceae* (amarylis family), the leek (*Allium porrum*) is related to a great number of other species of the genus and of similar odor and taste. Closely related species are chive, garlic, onion, shallot, and Welsh or Japanese onion. The leek resembles the onion in its adaptability and cultural requirements. Instead of forming a bulb, the leek produces a thick, fleshy cylinder that has the characteristics of a large, green onion. See accompanying illustration.



Leek (*Allium porrum*), closely related to the onion. (*USDA photo.*)

Usually, the seeds are sown in a shallow trench so that the plants can be more easily hilled up as growth proceeds. Leeks are ready for use any time they reach a proper size. Under favorable conditions, they will grow to 1.5 inches (4 cm) in diameter or more, with white parts from 6 to 8 inches (15 to 20 cm) in length. They may be lifted in the fall and stored like celery in a dry, cool place.

LEEWAY. The difference between the actual direction in which a ship is moving relative to the surface of the water and the direction in which the keel of the ship is pointing. Leeway is usually produced by the pressure of the wind against the side of the vessel and is much more pronounced in the case of sailing vessels than in internally powered ships. The amount of leeway can best be determined by observing the angle between the wake of the vessel and the line of the keel.

In determining the true course of the vessel, the leeway is treated in the same manner as a compass correction. If the wind is blowing against the left side of the vessel, the vessel is said to be on the port tack, and the true course will be to the right of the course indicated by the keel. Hence, for a ship on the port tack, leeway has the same effect as an east or positive compass correction; on the starboard tack, leeway is applied as a west or negative correction.

See also **Course;** and **Navigation.**

LEFT-HANDEDNESS (System). See **Handness (Right-and Left-).**

LEGENDRE DIFFERENTIAL EQUATION. A second-order equation

$$(1 - x^2)y'' - 2xy' + n(n + 1)y = 0$$

It is a special case of the associated Legendre equation

$$(1 - x^2)y^n - 2xy' + \left[n(n + 1) - \frac{m^2}{1 - x^2}\right]y = 0$$

which in turn is a special case of the Gauss hypergeometric equation. Both Legendre equations have singular points at $x = \pm 1, \infty$ and if m, n are integers the solutions are Legendre polynomials. For nonintegral values of these parameters, the solutions are Legendre functions. These differential equations occur in the quantum mechanical problems of the rigid rotator and the hydrogen atom.

The associated polynomials may be defined by the expression

$$P_n(x) = (-1)^m(1 - x^2)^{m/2}\frac{d^m P_n(x)}{dx^m}$$

and the special case $m = 0$ is the Legendre polynomial of degree n

$$\begin{aligned} P_n(x) &= \frac{1}{2^n n!}\frac{d^n}{dx^n}(x^2 - 1)^n \\ &= \frac{(2n)!}{2^n(n!)^2}\left[x^n - \frac{n(n-1)}{2(2n-1)}x^{n-2}\right. \\ &\quad \left. + \frac{n(n-1)(n-2)(n-3)}{2 \cdot 4(2n-1)(2n-3)}x^{n-4} \pm \cdots\right] \end{aligned}$$

The first definition, in terms of the nth derivative, is called the *Rodrigues formula*. A second set of polynomials, linearly independent of P_n, is composed of polynomials of the second kind, Q_n. The general solution of the Legendre equation is then $y = AP_n(x) + BQ_n(x)$, where A, B are arbitrary constants. Many other definitions and relations for these polynomials are known.

One integral representation is the *Schläfli formula*

$$P_n(z) = \frac{1}{2\pi i}\int_C \frac{(t^2 - 1)^n}{2^n(t - z)^{n+1}}\,dt$$

where the contour encircles the point z in the counterclockwise direction in the complex plane. Another such representation is the *Heine formula*, which for the associated polynomial becomes

$$\begin{aligned} P_n^m(x) &= (n+1)(n+2)\cdots(n+m)(-1)^{m/2} \\ &\quad \times \frac{1}{\pi}\int_0^\pi x + \sqrt{x^2 - 1}\cos\phi]^n \cos m\phi\, d\phi \end{aligned}$$

See also **Generating Function.**

LEGENDRE SYMBOLS. See **Number Theory.**

LEGIONELLOSIS. Initially called legionnaire's disease, this illness is a form of pneumonia and is caused by *Legionella pneumophils*, a filamentous gram-negative bacillus. The disease was first noted in Philadelphia in 1976, at which time an outbreak of 82 cases occurred. Since that time, several outbreaks have occurred in the United States and a few in Europe. Because the causative agent had not previously been identified, statistics on the disease still are rather meager. It does appear that the number of cases in the United States peaks in late summer and early fall and is at a minimum during the winter months. At least six distinct serogroups of *Legionella pneumophilia* have been identified by immunofluorescence studies. Several cases of pneumonia that clinically resemble legionellosis have been shown to be caused by previously uncharacterized gram-negative bacilli that have several phenotypic characteristics of *L. pneumophila*, but represent *Legionella* species separate from *L. pneumophila*.

Shortly after infection, the patient presents a predrone of myalgias, malaise, and slight headache lasting for about 24 hours, which is the start of an incubation period which lasts from 2 to 10 days. Then rather abrupt high fever, shaking chills, dry cough, and usually pleuritic pain are present. In some cases, there is abdominal pain, diarrhea, and vomiting. In the first-noted Philadelphia outbreak, 16% of cases exhibited rapidly progressing symptoms, ending with respiratory failure, shock, and death. Although not yet fully determined, it appears that erythromycin is the drug of choice in the treatment of the disease.

Legionellosis is a notifiable disease in the United States. Approximately 1200 cases were report in 1992, making it one of the least prevalent of the notifiable diseases, such as measles, rubella, tuberculosis, and so on. The largest number of cases has occurred in the east-north central, mid-Atlantic, and south Atlantic reporting areas.

Although legionella generally is regarded as a pulmonary pathogen and has been responsible for several outbreaks of pneumonia in both community and hospital settings, extrapulmonary legionella infections have been reported. P. N. Lowry and researchers (Stanford University) reported in 1991 of instances of sternal wound infections attributable to inoculation of legionella into fresh wounds as the result of bathing patients with contaminated tap water during dressing changes.

Since legionella were identified during the initial investigations of legionellosis, the potable water supply has been implicated as a principal reservoir for *Legionella pneumophila*. R. S. Bhopal (University of Newcastle upon Tyne) and R. J. Fallon and M. R. C Path (Ruchill Hospital, Glasgow), in their studies of the epidemiology of Legionnaires' disease, have shown a strong association between the proximity of patients' homes to a cooling tower and an elevated risk of Legionnaires' disease (a threefold greater risk among persons living within 0.5 km (0.3 mi) of a cooling tower). C. W. Hoge (Armed Forces Research Institute of Medical Science, Bangkok, Thailand) observes, "The prevalence of legionellae in sources of potable water (particularly hot water) and cooling towers not associated with clinical disease is high, ranging from 11 to 52 percent. In addition, the diagnosis of Legionnaires' disease is frequently based only on seriologic examination (such as indirect fluorescent antibody testing or urinary-antigen determination), which precludes a comparison between clinical and environmental strains."

Additional Reading

Addis, D. G., and J. P. Davis: "Sporadic Cases of Legionnaires' Disease," *N. Eng. J. Med.*, 1699 (June 18, 1992).

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LEGUMINOSAE. The Pea Family is second only to the Composite Family among the dicotyledons, with respect to the number of species it includes. Of its more than 10,000 species in early 500 genera, many are trees or shrubs, especially those in tropical regions. Herbaceous species are numerous in temperate regions. Many climbing plants, also, are found in the family. Leguminous plants are found in all sorts of environments and climates.

Nearly all the plants in this family have pinnately compound leaves. The stipules present in the leaves are sometimes modified to persistent spines. The flowers are either regular or irregular. When regular, the flowers have five sepals, commonly more or less united, five petals, a varying number of stamens, and a single pistil. Irregular flowers are of the type known as papilionaceous, a name given because of the fancied resemblance of the flower to a butterfly. In flowers of this type, the calyx has five unequal, more or less united sepals, which frequently persist during development of the fruit, five separate petals, showing very constant difference in form. The upper one, called the standard, is large and showy; the two lateral to this, called wings, are smaller in size; and the two lower ones are more or less united into one unit, called the keel or carina. Within this keel are the ten stamens, which may be separate but in many genera are united in groups, the nine lower ones having their filaments more or less completely joined, while the tenth stamen remains free. The pistil has a somewhat flattened ovary, a long style, and a terminal stigma. The ovary contains several ovules. The mature fruit is called a pod or legume, which when mature often splits open with sufficient force to eject the seeds to considerable distances. The seeds in most cases have large food reserves stored in the thick cotyledons.

There are three subfamilies in the *Leguminoseae*. The *Mimosoideae* have regular flowers and valvate corolla. The *Caesalpinioideae* have irregular (zygomorphic) flowers. These two subfamilies are essentially tropical. The *Papilionoideae* have irregular (papilionaceous) flowers, and include most of the important cultivated forms.

Many members of this family supply man with important foodstuffs, such as beans, peas, and peanuts, while others are important forage crops for domestic animals. The high protein content of the plant is the principal reason for its importance as a food source. Clovers and alfalfa are not only valuable as forage plants, but furnish an excellent hay. Legumes are also of immense value because of their association with nodule-forming bacteria, resulting in a considerable accumulation of nitrogenous substance, which is later liberated into the soil, greatly enriching it. Other members of the family yield valuable dyes, gums and resins, and oils; many are sources of timber.

The leaves of many genera of legumes are interesting because of their ability to move. In many of them, the leaflets fold together at night, so that the blade of the leaflet is vertical. Of particular interest in this connection is the Sensitive plant, *Mimosa pudica*, the leaves of which respond very quickly to external stimuli. A light blow will cause the many leaflets to fold together, and the whole section of the compound leaf to bend down. A sudden breeze or change of temperature will produce the same result. Recovery from the shock is gradual. When stimulated by a series of successive shocks, the plant recovers more and more slowly each time.

See also **Locust Trees.**

LEHR. See **Glass.**

LEISHMANIASIS. There are two major subtypes of this disease—*visceral* and *cutaneous* leishmaniasis. Visceral leishmaniasis (kala-azar; tropical splenomegaly) is a chronic, systemic disease caused by the flagellate protozoan *Leishmania donovani.* Discrete foci are distributed in South and Central America, Africa, Asia, the former U.S.S.R., and Europe. Kala-azar is basically a domestic disease where poor hygiene, moderate heat and vegetation favor the activities of the sand-fly (*Phelbotomus*) vectors. There are marked geographic variations in epidemiological patterns: in India only about 40% of cases are infants and the disease is largely urban, but no extra-human reservoirs are known. In China the disease is both urban and rural, children are most affected, and dogs are an important reservoir. Clinical disease usually appears 2–4 months after the infectious sandfly bite. The tiny (2–3 Mm) amastigote stage (LD body) multiplies as an intracellular parasite within reticuloendothelial cells, leading to cell destruction. In established infections there is irregular, undulant fever, emaciation and diarrhea, anemia and leukopenia, hepatosplenomegaly, and lymph node enlargement. Untreated disease shows high mortality, from 1 week to 3 years following onset. Recovery is followed by lasting immunity. Diagnosis is based on a suggestive clinical picture coupled with demonstration of LD bodies in biopsies of bone marrow or lymph nodes. Inoculation of culture media or hamsters may be necessary. After adequate treatment with antimony (Pentostam®) or diamidine compounds, prognosis is excellent. Control measures should include residual spraying and personal protection measures against sandflies, clearing of decaying vegetation, and reduction of the dog population.

Cutaneous leishmaniasis, also spread by sanddlies, involves several *Leishmaniae* and takes a number of clinical forms. The amastigotes are restricted to the area of the small skin ulcers, producing a cutaneous rather than systemic disease. *Leishmania tropica* produces Old World cutaneous leishmaniasis, also known as Oriental sore, Baghdad boil, or Delhi boil. It occurs around the Mediterranean, in the Middle East, southern former U.S.S.R., northwest India, and parts of Africa. There are two subtypes of Old World cutaneous leishmaniasis: (1) The moist or rural form (*L. tropica major*) is a zoonosis of desert gerbils. In humans, it has an incubation period of 6–10 weeks and produces cutaneous ulcers which may spread. (2) *L. tropica minor* produces the dry or urban form, harbored by both humans and dogs. The incubation period is longer (6–10 months), the dry skin lesion does not spread, and usually heals spontaneously within a year. New World cutaneous leishmaniasis likewise includes (1) *L. tropica mexicana* (Chiclero ulcer) and (2) L. *braziliensis* (mucocutaneous leishmaniasis or espundia). The first of these is a zoonosis of rodents in Central America, producing spontaneously healing ulcers on the pinna of the ear in humans. *Espundia* is a zoonosis, especially prevalent where jungle has recently been cleared. After inoculation the parasite spreads, especially to the oro-nasal areas, where it may produce progressive, destructive, disfiguring lesions, Espundia may persist for years, followed by death from septicemia or bronchopneumonia.

Diagnosis of cutaneous leishmaniasis is, by demonstrating the amastigotes in smears from ulcers or in cultures on NNN medium, or by the Montenegro skin test. Effective drugs include pentavalent antimonials (Pentostam®), pyrimethamine (Daraprim®), and amphotericin B (Funizone®). The prognosis is good, except in untreated mucocutaneous leishmaniasis. Infection often confers lifelong immunity. Control measures include covering individual lesions, reduction of natural reservoirs, and personal protection against sandflies.

R. C. V.

LEMMA. A proposition which is stated for the purpose of later use in the proof of another proposition. Strictly, the proposition is assumed. In current use, a short proof is often given.

LEMMING. See **Rodentia.**

LEMNISCATE OF BERNOULLI. A special case of a general type of higher plane curves known as lemniscates (L. *lemniscus*, loop) and also of the oval of Cassini, when $k = a$. Its equation in rectangular coordinates is

$$(x^2 + y^2)^2 + 2a^2(y^2 - x^2) = 0$$

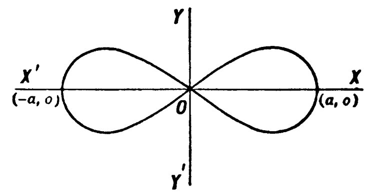
or, in polar coordinates,

$$r^2 = 2a^2 \cos 2\theta$$

The curve is symmetrical about both X- and Y-axes and the node at the origin has tangents given by $(x^2 - y^2) = 0$. If $\cos 2\theta$ is replaced by $\sin 2\theta$ in the polar equation for the curve it is rotated about the origin by 45° and is then called a two-leaved rose lemniscate (see **Rose Curve**). A more general type of curve, known as Booth's lemniscate, has the equation

$$(x^2 + y^2)^2 = a^2x^2 \pm b^2y^2$$

The curve $x^4 = a^2(x^2 - y^2)$ is known as the lemniscate of Gerono.



Lemniscate.

See also **Curve (Higher Plane).**

LEMON TREE. See **Citrus Trees.**

LEMUR (*Mammalia*). Lemurs are the most primitive animals of the order of Primates. They have a well-developed thumb and great toe, like other primates, but the second toe bears a sharp claw instead of a nail. The body is generally more like that of a squirrel than like the apes and monkeys and the face in many species is peculiarly expressionless, with large staring eyes. Lemurs constitute a family *Lemuroids*, containing, in addition to the species named as lemurs, the indri, the sifakas or propitheques, the galagos, the awantibo, the pottos, the lorises or slow lemurs, and the avahi. They center in Madagascar, but a few species occur in eastern Africa, southern India, and on the islands of the Oriental region. The tarsiers and the aye-aye are closely related to the true lemurs. A ring-tailed lemur is illustrated in the accompanying figure.

Ring-tailed lemur. (*New York Zoological Society*.)

The African slow lemur is named the *potto* and resembles the lorises of Asia in many ways. The potto is remarkable for the rudimentary index finger, which is a stub without a nail or joints, and for the very short tail. The two species are Bosman's potto and the awantibo, *Perodicticus calabarensis*. The sifaka is the genus *Propithecus*, with several species found in Madagascar. They are related to the indri lemur, but have long tails and shorter muzzles.

The galago is an African lemur, also sometimes called bush-baby. The several species have long bushy tails, large ears, large staring eyes, their thumbs and big toes opposed, and thick woolly soft fur. Their nearest relatives are the mouse lemurs of Madagascar.

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LEMUR (Flying). See **Dermaptera.**

LENGTH OF A CURVE. The length of a straight line is interpreted experimentally to mean the number of times another straight line of unit length can be superimposed on the given line. Since this operation cannot be conveniently applied to a curve the concept is generalized to the limit of the sum of chords to the curve. As the number of chords increases without limit each chord separately approaches zero as a limit.

In rectangular coordinates, if the curve is described by $f(x, y) = 0$, its length from the point (a, c) to the point (b, d) is given by either one of the definite integrals

$$s = \int_a^b \sqrt{1 + y'^2}\, dx = \int_c^d \sqrt{1 + x'^2}\, dy$$

where $y' = dy/dx$ and $x' = dx/dy$. If t is a parameter, $x = f_1(t)$, $y = f_2(t)$, the arc length between $t = a$ and $t = b$ is

$$s = \int_a^b \sqrt{x_t^2 + y_t^2}\, dt$$

where $x_t = dx/dt$ and $y_t = dy/dt$. In polar coordinates, if $r = f(\theta)$, $r' = dr/d\theta$ and $\theta' = d'/dr$,

$$s = \int_{\theta_1}^{\theta_2} \sqrt{r^2 + r'^2}\, d\theta = \int_{r_1}^{r_2} \sqrt{r^2\theta'^2 + 1}\, dr$$

See also **Circular Curves;** and **Coordinate System.**

LENSES. See **Mirrors and Lenses.**

LENS (Eye). See **Vision and the Eye.**

LENS (Field). See **Field Lens.**

LENTICELS. The young stems of plants are covered with a single layer of cells known as the epidermis. As the stem grows older, this epidermis is lost and replaced by a thicker protective tissue known as cork. The cork cells have walls which are suberized and impervious to gases. The living cells within the stem require an exchange of gases with the outside atmosphere. This exchange occurs through lenticels. See Figs. 1 and 2. A lenticel is a mass of thin-walled parenchyma cells loosely arranged so that air spaces are numerous. Through the lenticel gases pass readily. Lenticels appear on the surface of the stem as rough masses, usually protruding somewhat, and either circular or somewhat elongate in shape. They are very irregularly distributed.

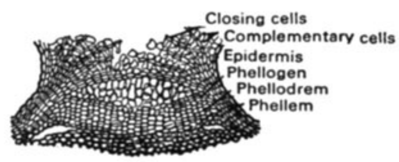


Fig. 1. Section through lenticel of cherry.

Fig. 2. Stems of apple tree, showing external view of lenticels.

LENTICULARIS. See **Clouds and Cloud Formation.**

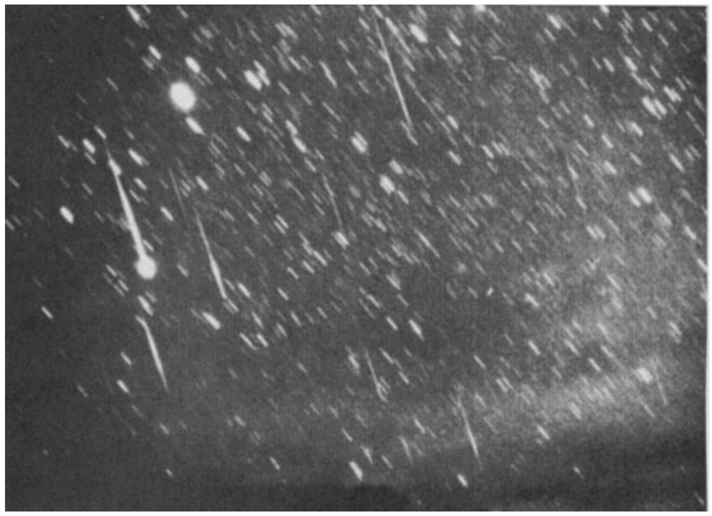
LENZ' LAW. A general law of electromagnetic induction, stated by H. F. E. Lenz in 1833. It points out that the electromotive force induced by the variation of magnetic flux, with reference to a conductor, in the manner discovered by Faraday, is always in such direction that, if it produces a current, the magnetic effect of that current opposes the flux variation responsible for both electromotive force and current. An outstanding illustration is the drag on a generator armature; if the armature circuit is closed, the rotation is opposed by a torque arising from the reaction between the field and the current in the armature conductors. Power must therefore be applied to drive the machine; and the greater the armature current, the more power is required. The effect known as magnetic damping also depends upon Lenz' Law. A copper disk, when spun between the poles of a strong magnet, quickly comes to rest because of the opposing torque. This arrangement serves as a speed regulator in watt-hour meters.

LEOPARD-CATS. See **Cats.**

LEO (the lion). One of the most easily distinguished of all the zodiacal constellations. The "sickle" of Leo, the fifth sign of the zodiac, is known to all watchers of the spring and early summer skies. The brightest star in the group, Regulus, is a double star, but cannot be resolved with telescopes having smaller than a 3-inch aperture because of the fact that, in small instruments, the bright star masks the fainter one. Gamma Leonis is one of the finest of the double stars, having two components of approximately the same magnitude, one yellow and the other orange in color.

This constellation is also noted for the location of the radiant point of the Leonids, one of the best known meteor showers. (See map accompanying entry on **Constellations.**)

LEONIDS. A name applied to a meteor shower that has probably attracted more attention than any other. Each year, about the 12th of November, a number of meteors are observed coming from a radiant point in the constellation of Leo. Records of the appearance of this shower are found back as far as 585 A.D. The Leonid shower is one in which meteors are distributed all along the orbit, so that a radiant point may be determined practically every year. There is, also, a very strong condensation of the meteors into a swarm through which the earth used to pass every 33 years. Probably the greatest display was in November, 1833. In Silliman's Journal of that year, we find: "To form some idea of the phenomenon, the reader may imagine a constant succession of fireballs, resembling rockets, radiating in all directions from a point in the heavens." One observer counted 650 during 15 minutes. During the interval between 1833 and 1866, a great deal of computing was done on the Leonid shower, and November 13 was predicted as the date of passage of the earth through the main swarm. The prediction was fulfilled, and at Greenwich, England, eight observers actually counted 8000 meteors, 4860 of them being counted between one and two o'clock in the morning. However, brilliant as the shower was at that time, it apparently was not as striking as the display in 1833. In 1899, there was a moderately good display of the Leonids, but nothing comparable to the showers of 1866 and 1833. The newspapers had promised so much to the general public that the failure of the shower to come up to the expectations proved a rather serious blow to astronomy. The explanation for the failure of the shower to live up to its prediction is to be found in the fact that Jupiter passed very close to the swarm during 1899 and deflected it from the earth's orbit. Further perturbations have so deflected the orbit that no real shower was observed in 1932, 1933, and 1934, although enough meteors were seen during each November to permit a determination of the radiant point. In November, 1966, however, perturbations again brought the main swarm into such a position that the earth passed through it, giving rise to showers comparable to the one of 1833. (See photograph.)



Leonid shower as seen from Kitt Peak near Tucson, Arizona.

LEPIDOLITE. This member of the mica group of minerals is a silicate of potassium, lithium and aluminum, sometimes with sodium, fluorine, or rarely rubidium. A general formula is K(Li, Al)$_3$ (SiAl)$_4$O$_{10}$(F, OH)$_2$. Crystals of lepidolite are monoclinic but often pseudo-hexagonal; cleavage, basal and perfect, being susceptible of splitting into thin laminae; hardness, 2.5–4; specific gravity, 2.8–3.3; luster, pearly; color, reddish to violet, grayish-blue, gray to white. A variety carrying rubidium is yellowish-gray; translucent. It usually is found as granular to scaly masses, in short stocky prisms or less often in easily cleavable sheets. Lepidolite is characteristic of pegmatite veins, frequently being associated with other lithium-bearing minerals such as tourmaline, spodumene, amblygonite, and others. It occurs in the Ural Mountains, the Czech Republic and Slovakia, the Island of Elba, and Madagascar, where it is often found in large sheets. In the United States, it is found in the pegmatites of New England, California, South Dakota, and New Mexico. The name lepidolite is derived from the Greek, meaning scale. See also **Lithium.**

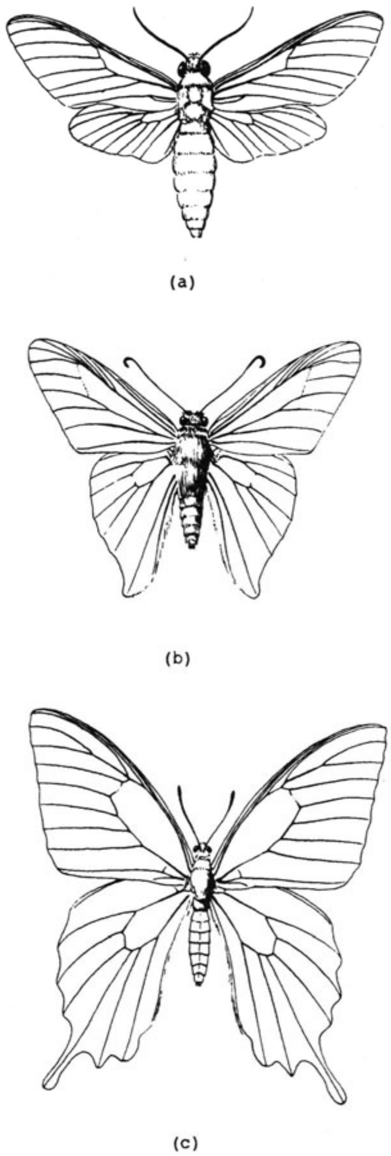
LEPIDOPTERA. The butterflies, skippers, and moths. An order of insects characterized by sucking mouths in the adult stage with two pairs of wings covered at least in part with a vestiture of flattened scales, complete metamorphosis and a larva with biting mouth-parts. The second order of insects in size, with about 90,000 known species.

Butterflies and moths are widely distributed and because of their bright colors are among the animals known to everyone. The butterflies are diurnal and so are readily observed, but most moths are nocturnal, hence many beautiful species are rarely seen unless they are sought. Skippers are an intermediate group more nearly like the butterflies. The most magnificent species of all three forms are tropical, but representatives are found even in the Arctic regions.

The adults visit flowers for nectar or take no food. In the larval stage most species are plant feeders, but a few carnivorous forms are known and some are scavengers. The order includes many economic species, among them the clothes moths, the bee moth, and the cut worms.

The main families of *Lepidoptera* include:

Aegeriidae (also called *Sesiidae*)	Clear-winged moths
Arctiidae	Tiger moths
Citheroniidae (also called *Ceratocampidae*)	Royal moths
Coleophoridae (also called *Haploptiliidae*)	Casebearers
Cossidae	Carpenter moths
Eucleidae (also called *Limacodidae* or *Cochidiidae*)	Slug-caterpillar moths
Gelechiidae	Gelechid moths
Geometridae	Measuring-worm moths
Gracilariidae	Leaf blotch miners



Typical members of Lepidoptera: (a) moth; (b) skipper; (c) butterfly. (*USDA.*)

Hepialidae	Swifts (swift-flying)
Hesperiidae	Common skippers
Incurvariidae	Yucca moths
Lasiocampidae	Lappet moths, tent caterpillars
Lycaenidae	Blues, hairstreaks, gossamer wings
Micropterygidae	Mandibulate moths
Nepticulidae	Nepticulid moths
Noctuidae (also called *Phalaenidae*)	Owlet or cutworm moths
Notodontidae	The prominents
Oecophoridae	Parsnip webworm
Olethreutidae (also called *Eucosmidae*)	Codling moths
Papilionidae	Swallowtail butterflies
Pieridae	White and sulfur butterflies
Psychidae	Bagworm moths
Pyralididae	Pyralid or snout moths
Saturnidae	Giant silkworm moths
Sphingidae	Hawk or sphinx moths
Tineidae	Clothes moths
Tortricidae	Leaf-roller moths
Yponomeutidae	Miscellaneous grouping, including diamondback moth and apple fruit moth.

LEPROSY. A chronic infectious but only mildly contagious disease caused by *Mycobacterium leprae*. The disease presents a great variety of signs and symptoms, depending on what tissue or organ of the body is involved.

Leprosy is a disease of antiquity, and there is evidence that it has existed at least since 2000 B.C. References to the disease are found in the Old Testament. Essentially, leprosy is a tropical disease. While it has been common in the Orient for several thousands of years, it appeared as a scourge in Europe in the eleventh and twelfth centuries, and did not subside until the sixteenth century when segregation of the victims was carried out on a large scale. At present the disease occurs endemically and sporadically, chiefly in the Orient, Australia, Asia, in Mediterranean countries, and in Central and South America. There are various other foci of sporadic cases such as some parts of northern and central Europe, the West Indies, Louisiana, Minnesota, and South Carolina in the United States, and several in Canada. Occasional cases are encountered in the larger seaports, both Atlantic and Pacific.

Some 12 million cases exist in the world, of which 3.5 million are found in India. The approximate number of 25,000 cases seen now in Europe and the United States have been imported since World War II from tropical countries by way of immigrants.

The mode of infection by the causative agent is not definitely known. The nasal mucosa, the gastrointestinal tract, and the skin have been considered as possible portals of entry. The disease is definitely contagious, but years of exposure and contact seem to be necessary for its transmission.

The organism causing leprosy is similar to the tubercle bacillus in appearance and staining characteristics. It is found in great numbers in the nodules occurring under the skin, in discharges from the nose and throat, and in discharges from ulcers. It was first discovered in 1873 by Hansen.

Much difficulty has been encountered in culturing, or rather maintaining, the organism, since no artificial medium has been found which will sustain it. Instead, mouse foot-pad inoculations have been the major source of organisms for study. More recently use of the immunodeficient mouse and the armadillo have provided almost adequate supplies of *M. Leprae*. To a great extent, leprosy may be considered an immunological disease. The leprosy bacillus is virtually nontoxic and most symptoms of the disease are due to immune reactions against antigenic constituents of *M. leprae*.

The period of incubation is variable. It may be from a few months to 20 to 30 years. There are two general types of the disease: (1) nodular leprosy in which the skin is primarily involved; and (2) maculoanaesthetic leprosy, in which there is an involvement of nervous tissue. A mixed form also occurs showing symptoms of both forms.

In the first stages of nodular leprosy, brownish-red spots appear on the skin, usually on the limbs and face, covering large and small areas of the skin. Later nodular thickenings appear at these sites. The face may show the so-called leonine appearance, due to the thickening of the skin in the region of the forehead, eyes, lobes of ears and around the nose and mouth. The entire skin assumes an unhealthy, dusky appearance. Some of fhe thickened areas ulcerate and fingers and toes may atrophy. Ulceration also appears in the nose and throat and the voice becomes hoarse. The eyes are affected similarly, and blindness may result. This form of the disease may last 10, 20 years, or longer without treatment. Many of the patients die of complicating disorders such as pneumonia, nephritis, tuberculosis and malnutrition.

Maculo-anaesthetic leprosy is characterized by flat, red to brown lesions on the skin, distributed symmetrically. These lesions gradually become insensitive to pain (anaesthetic); trauma, even burns, may occur without the patients feeling pain. Ulceration of the area and contractions will produce deformities.

At one time, chaulmoogra oil was widely used in the treatment of leprosy. Cortisone drugs have been used successfully in the treatment of certain manifestations and for certain phases of the disease, notably for relief of acute lesions of the eye. Derivatives of diaminodiphenylsulfone (Dapsone), administered orally or by injection, have brought about marked improvements, particularly in advanced cases, and for ridding the body of *Mycobacterium leprae* in from two to four years. Para-aminosalicylate and streptomycin also have been used successfully.

Most recently, Dapsone resistance has been seen in the disease organisms so that emphasis is being placed upon the development of a vaccine. This is proving to be a difficult problem, but meanwhile, what appears to be a successful intermediate mode of treatment, has been devised in Venezuela. Purified *M. leprae* bacillus is combined with *Bacillus Calmette-Guerin* (BCG). When this is used to vaccinate leprosy victims, the bacilli were cleared from the patients' bodies in 18 months and progress of the disease was halted. Further work is being done to confirm the value of this form of prophylaxis.

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LEPTONS. The electron, muon, and two kinds of neutrino are collectively called *leptons.* The leptons are considered to be pointlike particles without structure and thus truly elementary. Leptons can interact with other particles through the weak interactions. Electrons and muons also can interact through electromagnetic and gravitational forces, but they appear to be without capability of interaction through the strong (nuclear) forces. The neutral members, the electron neutrino, the muon neutrino, and their antiparticles have extremely weak interaction with matter and do not participate in electromagnetic interactions. Leptons make excellent probes in particle physics experiments. The other major family of subatomic particles is referred to as *hadrons.* See also **Electron; Muon; Neutrino;** and **Particles (Subatomic).**

The name *lepton* from its derivation means "light," referring to the fact that the masses of the leptons are all lighter than that of the lightest meson. The properties of the electron are discussed in that entry; here it will merely be noted that the term *electron* is used to denote the negative electron (often called the "negatron" when ambiguity might arise). Its antiparticle is the positron (also called positive electron).

LEPTOSPIROSIS. Caused by pathogenic spirochetes (*Leptospira interrogans*), leptospirosis may take a number of forms. The serotype *icterohemorrhagiae* causes the syndrome known as Weil's disease; the *autumnalis* serotype causes pretibial fever (Fort Bragg fever); serotypes *canicola, icterohemorrhagiae*, and *pomona* may cause aseptic meningitis (leptospiral meningitis), which differs from viral meningitis. In recent years, infections with the *canicola* serotype have occurred most frequently in the United States.

Humans may be infected through contact with the urine or affected tissues of an infected animal and the entry portal may be the skin (hands, etc.), the conjuctiva, or oral mucous membrane. Contaminated soil and water also can be infective. The 1964 outbreaks (approximately 150 cases) in the United States largely reflected water-related infections. Leptospiras can survive for several weeks in a warm, neutral or alkaline medium. Transmission of the disease between humans is rare. At one time, it was believed that leptospirosis was mainly associated with farming, husbandry, veterinary work, meat packing, and sewage facility work, but in recent years outbreaks have occurred among persons engaged in recreational activities, particularly involving contaminated water. The disease is usually seen during the summer months and shows no geographical patterns. Since 1950, the average number of cases per year has been rising slowly, but in a cyclic manner. A peak of 76 cases was reported in 1964. Over the last few years, average cases per year have numbered about ninety.

The incubation period ranges from 1 to 3 weeks. Frequently, the disease is self-limiting. The more serious cases involve secondary developments during the course of the disease. Symptoms usually include moderate to high fever, chills, headache, and prostration. Onset is usually quite abrupt. Fever and chills may persist for about one week. In some patients, cough and chest pain are also present. Muscle tenderness (calves, thighs) and stiff neck are common complaints. Conjunctivitis is present in some patients. This initial phase of the disease abates in about one week, followed by a period of a few days of apparent recovery, during which period the leptospira disappear from the blood. Then the second phase commences, usually with somewhat milder symptoms. In very serious cases, this period of apparent remission may not occur. Generalizations are difficult because of the numerous courses which the disease can take.

In the Weil's disease syndrome, the second phase features high fever and the liver enlarges and becomes tender. There may be purpura and gastrointestinal bleeding. Renal damage may occur, with accompanying jaundice. Oliguria (reduction in urine flow) may be serious. Weil's disease is the most serious of the leptospirosis infections, with a mortality of 15–40 percent.

The use of antimicrobials is usually not effective unless their administration is commenced within the first few days after onset. Both penicillin G and tetracycline antibiotics have been used. Further treatment depends upon the course of the disease, with each feature addressed specifically. In cases of renal damage, hemodialysis or peritoneal dialysis may be indicated.

The vaccination of dogs, sometimes implicated in leptospirosis, is not fully protective because even healthy animals can shed the leptospiras in their urine.

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LEPUS (the hare). A southern constellation located near Canis Major and Orion.

LESCH-NYHAN SYNDROME. A sex-linked genetic disorder characterized by random, uncontrollable movements, mental retardation, and a self-destructive psychotic behavior. An enzyme deficiency results from the absence of a gene (or presence of a defective gene) on the X chromosome. The disorder affects only males because males have only one X chromosome. Because females have two X chromosomes, the presence of a normal gene on one of these suffices and there is no resulting disorder.

LESION. An injured or otherwise impaired tissue of the body.

LESSER CATS. See **Cats.**

LETTUCE. See **Rose Family**

LEUCITE. The mineral leucite is a metasilicate of potassium and aluminum corresponding to the formula $KAlSi_2O_6$. It is isometric at a temperature of about 600°C (1112°F) and pseudoisometric at lower temperatures, at which leucite is tetragonal but retains an external isometric crystal form, usually trapezohedral. It has a conchoidal fracture; is brittle; hardness, 5.5–6; specific gravity, 2.47–2.50; luster, vitreous; color, white or some shade of gray; translucent to opaque. It is commonly found in the more recent lavas of high alkali content. Leucite is seldom reported from plutonic rock types. It is a relatively rare mineral. It is found plentifully at Vesuvius and Monte Somma and elsewhere in Italy, and Germany in the Tertiary volcanic district of the Eifel. In the United States, leucite has been found in the Leucite Hills of Wyoming, the Highwood Mountains of Montana, and as pseudomorphs (pseudoleucites) representing a mixture of nepheline, orthoclase, analcime, and aegerine from New Jersey, Arkansas, and Montana. Its name is derived from the Greek word leukos, referring to its white color.

LEUCOCYTES. See **Blood.**

LEUCOPLASTS. See **Plastics.**

LEUKEMIAS. The leukemias, like the anemias, represent a group of disorders of the blood cells rather than a single disease and they rank as a significant cause of death in children.

Normally, precursor (hematopoietic) blood cells follow a fixed sequence in their development and differentiation into erythrocytes (red blood cells) and leukocytes (white blood cells) and produce the complement of blood cells required by the human body. In the leukemias, however, the development and differentiation do not cease and the group of diseases is characterized by the undifferentiated proliferation of malignant cells derived from the hematopoietic precursors with resultant replacement of the normal bone marrow and often infiltration of other organs. The cell clone has (1) poor responsiveness to normal regulatory mechanisms, (2) a tendency to have a diminished capacity for normal cell differentiation, (3) an ability to expand at the expense of normal myeloid or lymphatic tissue, and (4) an ability to suppress or impair normal myeloid cell growth.

The cause of this erratic behavior of hematopoietic cells is unclear. Many factors have been suggested in the etiology; these include RNA viruses, ionizing radiation, and a number of drugs and chemicals which have been shown to have adverse effects upon the body. Genetic effects also appear to be involved, especially those conditions associated with chromosomal damage.

About 1–2 kg of leukemic cells in the body ($1–2 \times 10^{12}$ cells) appear to be sufficient to cause death from leukemia. This number would be about the total marrow volume of an average adult. Because the diagnosis of leukemia cannot usually be made unless the cell burden is 10^9 cells, it is apparent that only ten tumor cell doublings separate the smallest detectable number from a potentially lethal cell number. When a patient is said to be in remission it means only that the leukemic cell number is not clinically detectable and is, therefore, less than 10^9.

From a standpoint of terminology, the traditional basic classification of leukemia has been the acute and chronic forms of the disease. In acute leukemia, the predominant hematopoietic cell is a primitive precursor which may be undifferentiated or have the features of lymphoblast, myeloblast, monocyte, or erythroblast. In the chronic disease there is more obvious differentiation into myeloid or lymphoid cells. Apart from this general classification, the leukemias are classed according to the kind of cell primarily involved.

The several different kinds of leukemia differ in symptoms and in life expectancy of the patient. The acute form occurs most frequently in young children; the chronic form usually in persons over 35 years old. *Lymphocytic leukemia* principally involves the white blood cells which arise from the lymph nodes and spleen. In *granulocytic leukemia*, one or more of the three types of granulocytes, originating in the bone marrow, are involved. *Monocytic leukemia* is typified by the appearance of excessive numbers of monocytes derived from connective tissue. In lymphocytic leukemia, an early symptom is enlargement of the lymph nodes and acute leukemia is often first detected by prolonged hemorrhage following a minor surgical procedure, such as tooth extraction. Fever, arthralgia and anemia are also early symptoms. In some acute leukemias the total white cell count may be normal or even below normal, although the particular type of cell is found in excessive numbers in other tissues.

Three conditions characterize chronic leukemias: (1) enlargement of the spleen; (2) increased numbers of white cells, and (3) anemia. Because of the numerous symptoms and variation in degree, the diagnosis of leukemia is confirmed only by microscopic examination of blood and bone marrow which exemplifies severe anemia and reduced blood platelet counts.

Some of the clinical manifestations of acute leukemias can be related to bone marrow failure, leading to infection, anemia, and bleeding, together with the metabolic effects of a large tumor mass and infiltration of various organs by malignancy.

Treatment of leukemia dates back several decades. In 1948, Farber initiated the folic acid antagonistic route involving daily injections of drugs, such as prednisone, l-aspariginase, vincristine, methotrexate, cytabarine, or cyclophosphamide. Often, within a few weeks, the patient was free of measurable signs of the disease, but when therapy was discontinued, relapse occurred and symptoms recurred within three months. Modern drug treatment follows along similar lines. It is a long-term maintenance procedure, requiring from 3 to 5 years and continued maintenance as required if death is to be averted. Probably less than 50 percent of patients survive after stopping all therapy beyond the initial period of 5 years or longer. Ironically, the chemotherapy used to arrest other cancers has become an important causation of leukemia. .

Treatment-Related Leukemias. The treatment methodologies used for primary cancers and the resulting effect upon the incidence of leukemia have been established, but require further elucidation. Much remains to be researched pertaining to the biologic nature of chemotherapy and radiation treatment effects that bring about leukemia. Past studies have indicated cytogenetic abnormalities, including the loss of entire chromosomes 5 and 7, produced by treatment of cancers as well as from a variety of environmental toxins, such as benzene. C. A. Coltman, Jr. and S. Dahlberg observe, "This consistent picture of resistant leukemia associated with defects of chromosomes 5 and 7, regardless of the underlying disease process and the type of carcinogen, implies a cause-and-effect relation. The deletion of these genes, which are critical to the proliferation and differention of hematopoietic (blood forming) tissues, must have a central role in the basic biology of treatment-related leukemia. The exact relation remains to be elucidated, but it is clearly a common pathway of carcinogenesis for ionizing radiation, chemotherapy, and environmental toxins in a wide variety of benign and malignant conditions."

It has been demonstrated that patients who have received chemotherapy with alkylating agents (cyclophosphamide, chlorambucil, melphalan, and other leukemogenic substances) for ovarian cancer have an increased risk of acute leukemia, particularly of the myeloid (resembling bone marrow) type. J. M. Kaldor (International Agency for Research on Cancer, Lyon, France) reported on an extensive study in 1990 and concluded, "The extent to which the relative risks of leukemia are offset by differences in chemotherapeutic effectiveness is not known." A further observation of the Kaldo report: "The trend in treatment for ovarian cancer is toward more use of chemotherapy and in particular an increasing reliance on combinations of drugs. Despite some promising recent findings (Tropé reference), there is limited evidence so far that the effectiveness of such chemotherapy improves with the intensity or number of agents, and our study clearly shows dramatic increases in the long-term risk of leukemia at high dosages."

The Kaldor group also investigated the effect of different treatments for Hodgkin's disease on the risk of leukemia. The conclusion: "We conclude that chemotherapy for Hodgkin's disease greatly increases the risk of leukemia and unaffected by concomitant radio therapy. In addition, the risk is greater for patients with more advanced stages of Hodgkin's disease and for those who undergo splenectomy." However, in another observation, "In the case of Hodgkin's disease, a substantial risk of leukemia following combination chemotherapy is clearly offset by enormous gains in survival. However, after ovarian cancer, the extent to which the increased risk of leukemia is offset is still unclear."

Other Approaches to Leukemia Treatment. Distinct from chemotherapy, red cell transfusion has been used for severe anemia at the outset of treatment, and although granulocyte transfer is not clearly understood, bone marrow transplantation is recognized as a definitive modality for aplastic anemia, acute leukemia, and chronic myeloid leukemia. The effect of such treatment becomes apparent after 2 to 4 weeks and is marked by a rise in circulating granulocytes and later by an increase in the platelet count. However, because allogenic grafts of marrow contain immunocompetent cells as well as marrow stem cells, the engrafted cells may mount an attack against the host. This becomes apparent in approximately half of the cases so treated.

A major goal in the treatment of acute leukemia is to reduce the amount of therapy and to minimize the adverse side effects without cutting the effectiveness of established treatments. For related topics, see **Blood; Bone;** and **Cancer**.

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LEUKOCYTOSIS. An abnormally high level of leukocytes (white cells) in the blood, a common manifestation of infection.

LEUKOPENIA. A decrease in the normal number of leucocytes in the bloodstream. This is a usual accompaniment of certain stages of some infectious diseases, while in other infections it is of grave prognostic significance, and indicates a failure of one of the lines of defense of the body. In agranulocytosis, leukopenia is the chief manifestation of the disease.

Drugs and chemicals sometimes implicated in leukopenia include arsenic, chloramphenicol, cimetidine, 5-fluorocytosine, phenytoin, sulfopyridine, thiazides, and trimethoprim. Leukopenia is sometimes associated with Colorado tick fever, meningococcemia, pneumococcal pneumonia, and salmonelolosis. Leukopenia is also a result of starvation.

LEVEL OF ESCAPE. Sometimes called *critical level of escape*, that level, in the atmosphere, at which a particle moving rapidly upwards will have a probability of $1/e$ of colliding with another particle on its way out of the atmosphere (where e is the number **e**, the base of natural logarithms). It is also the level at which the horizontal mean free path of an atmospheric particle equals the scale height of the atmosphere. The critical level of escape is the base of the exosphere, which is the outermost, or topmost, portion of the atmosphere. See **Atmosphere (Earth).**

LEVEL (Surveyor's). An instrument for determining differences of elevation. It is of use in obtaining comparative elevations of two points, or in defining the profile of a certain path, such as a roadway, drainage ditch, etc. A level of this type has an accurately made bubble level which is attached to and made exactly parallel with a telescope. In the wye level this telescope rests in Y-shaped supports. These supports are held in turn by the instrument base, which may be adjusted by hand, and which is attached, usually by screwing, to the top of a tripod, upon which the instrument rests when in use. The bubble tube being parallel to the center of the telescope, the latter will automatically be leveled, ready for a horizontal sight, when the bubble tube is level. The dumpy level has the telescope and supports cast in one piece or rigidly connected. Lasers are also used in levelling operations, particularly for grading irrigated areas. See also **Irrigation.**

LEVEL WIDTH (Excitation Energy). A measure of the spread in excitation energy of an unstable state of a quantized system. In emission or absorption spectra, variations in the intrinsic line widths show that some energy levels in atomic or nuclear systems are broad and others are narrow. In nuclear physics, level widths have been observed chiefly in connection with neutron and charged-particle resonances, which are found to have nonuniform breadths in energy. The level width is related to the mean life of the level by the expression

$$\Gamma = \hbar/\tau$$

where Γ is the level width, h is $\hbar/2\pi$ (h is Planck's constant), and τ is the mean life. Level widths usually show themselves as the widths of resonance peaks observed when the cross section for the particular reaction is plotted as a function of the energy of the incident particle. The quantitative value of the level width is usually taken as the full width at half maximum of the resonance peak.

If a system at a given level has several alternate modes of disintegration, there is associated with each a partial level width proportional to the probability of disintegration by the particular mode. The total level width is the sum of the partial level widths. See also **Broadening of Spectral Lines.**

LEVERS. See **Machine (Simple).**

L'HOSPITAL RULE. If two functions $f(x)$ and $g(x)$ together with their derivatives up to order $(n - 1)$ vanish at $x = a$, and if their deriva-

$$\lim_{x \to a} \frac{f(x)}{g(x)} = \frac{f^{(n)}(a)}{g^{(n)}(a)}$$

tives of nth order do not both vanish there or both become infinite, then See also **Indeterminate Form.**

LIBRA (the scales or the balances). A small constellation, best known as the seventh sign of the zodiac; its symbol is taken for the autumnal equinox (i.e., the point where the sun apparently crosses the celestial equator from north to south). The brightest star in Libra is a wide double star, which can easily be resolved with a field glass. This star carries the Arabic name Zuben el Genubi, the southern scale. (See map accompanying entry on **Constellations.**)

LIBRATIONS. A term applied in astronomy to many periodic oscillations; in particular, to slight apparent oscillations of the moon, whereby observes on the earth are enabled to observe somewhat more than 50% of the moon's surface. There are three principal librations of the moon: a libration in lunar latitude, a libration in lunar longitude, and a diurnal (or daily) libration.

The libration in lunar latitude arises from the fact that the orbit plane of the moon is inclined at about 6.5° to the plane of the moon's equator. This produces an effect relative to the earth similar to the terrestrial effects relative to the sun, which produce the seasons. During one-half of the month, the north lunar pole is directed slightly toward the earth, and during the remainder of the month the south lunar pole is toward the earth.

The libration in lunar longitude is due to the fact that, although the rotation of the moon is quite uniform, with a period equal to the period of revolution about the earth, the orbital motion is not uniform, but is in accordance with the second Keplerian law of planetary motion. Suppose that a certain point on the surface of the moon is directly toward the earth at a time when the moon is in perigee. Since the moon is moving more rapidly in its ellipse at perigee than at any other part of its orbit, by the time the elongation has increased to 90%, the selected point has not completed one-quarter of a rotation and is not directly toward the earth. The result is that the observer will see a slightly different hemisphere of the moon than he did at perigee. By the time the moon has reached apogee, the selected point will again be toward the observer, and he will have the same view of the moon as at perigee. At this point, the moon's orbital motion is a minimum; the selected point will complete a quarter rotation before a quarter revolution is completed, and another slightly different hemisphere of the moon will be visible.

The diurnal libration is due to the fact that the observer sees slightly "over the top" of the moon when the moon is rising, and slightly under the bottom when the moon is setting. This is really a libration of the observer rather than a libration of the moon, but is classed with the latter.

The combined result of the librations is that about 41% of the moon is always visible from the earth (or would be if the sun were shining upon it), 41% is never visible, and the remaining 18% is either visible or invisible, depending upon the particular position of the moon relative to the earth.

The term libration is also applied to certain periodic perturbations in the orbits of members of the solar system.

See also **Perturbation (Astronomy).**

LICHEN. Perennial plants which are a combination of two plants growing together in an association so intimate that they appear as one. Either of the component plants alone possesses none of the characteristics shown by the two in combination. Lichens often are cited as examples of symbiosis, i.e., two organisms living together of benefit to each other. The components of a lichen are always an alga and a fungus. The algal constituent is usually one of the simple green algae, or, more rarely, a blue green one. The alga can live perfectly well by itself, and is often found growing free on rocks or tree trunks in regions where the lichen would exist. The fungal component is usually a member of the ascomycetes. In lichens growing in cooler regions it is always one of this group. There are certain lichens found in the tropics, however, in which the fungal component is a basidiomycete. In the lichen, the fungus alone is capable of fruiting, although the algal cells do divide and so increase in number. It is very difficult to see how such an association came about, and to determine whether it is really a case of symbiosis or whether it is not parasitism, one of the plants living on the other. Unquestionably the fungus benefits from the presence of the alga, since the latter carries on photosynthesis, making food materials which are used by the fungus. The latter, lacking chlorophyll, cannot manufacture its own food. It is possible that the alga benefits by the added moisture gathered by the fungus, that the latter protects the alga against desiccation. Certainly the association is well established, and seemingly has been so for a long period of time. Lichens have been "made" artificially, that is, the two components have been grown separately in pure cultures, and when brought together have produced a lichen. So lichens can be formed anew, always with a very constant appearance characterizing the particular form considered.

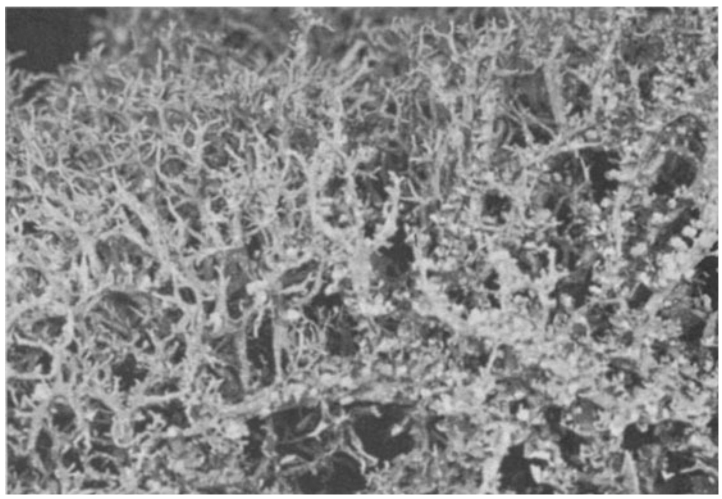
Lichens are found nearly everywhere where civilization has not killed them. They are found on the surface of rocks and soil; they occur on the bark of trees; in the tropics they may be found on the surface of thick evergreen leaves of trees; a few species even grow on rocks submerged by the tides.

The shape of the lichen body or thallus is very diverse. Some species are flat crusts growing on or even in the surface of the substratum, whether the latter be trunk of tree or barren rock. Lichens of this type are called crustose. In others the thallus is split up into many radiating divisions, and is called a foliose lichen. Many others have an erect, often much-branched thallus and are called fruticose lichens. The color of the thallus may be yellow, orange, brown, gray, or black.

The greater part of the lichen thallus is composed of fungus hyphae, which form a compactly tangled mass. The algal cells occur in an irregular loosely arranged layer near the outer surface of the thallus. Short irregular branches from the fungus hyphae grow tightly around each algal cell, sending into it short absorbing structures called haustoria. The surface of the lichen is composed of enlarged thick-walled fungus cells which form a compact layer over the more loosely arranged central portions. In many of the crustose and foliose lichens there are many rhizoids which anchor the plant firmly.

One of the ways in which a lichen reproduces is by means of soredia. These are minute bits of lichen, formed on the surface, and composed of one or more of the algal cells together with a small mass of closely associated hyphae. Often these soredia are so numerous as to give to the lichen a powdery appearance. Either through disintegration of the lichen body, or because the continuity of the hyphae breaks down, soredia become free from the thallus. They are then easily spread by wind or by water. They may even be carried about unintentionally by the many small insects and other animals which feed on lichens. Lichens also reproduce by means of spores; that is, the fungus component forms special reproductive structures very similar to those formed by similar fungi not forming lichen thalli. In the lichens composed of ascomycetes these reproductive structures are open cups or mounds, called apothecia. Commonly these apothecia are of a different color from that of the thallus.

Reindeer moss, *Cladonia rangiferina*, is the principal food of the reindeer, and may also be used as fodder for other animals. Reindeer moss, not a moss at all, is an erect much-branded lichen which grows abundantly over wide stretches of barren soil. The dense grayish-green tufts grow continuously at the tops, becoming 6–10 inches (15 to 25 centimeters) tall and attaining great age. Another lichen, *Cetraria islandica*, or Iceland moss, may be used as stock food. In habit it resembles reindeer moss but is coarser and less branched. See accompanying illustration.



Lichen (reindeer moss). (*A. M. Winchester.*)

Litmus paper is prepared from *Lecanora tartarea*, a common lichen found in the Netherlands.

LICORICE. This herbaceous native plant of southern Europe (*Glycyrrhiza glabra* L.) of the family *Leguminosae* is the source of popular flavorings used in the food industry. Stolons or roots (at least 2 years old) are extracted with hot water. The taste is sweet and rich; the essence is sometimes described as slightly spicy. The principal constituent of licorice is the potassium-calcium-magnesium salt of glycyrrhizic acid, which upon hydrolysis produces glycyrrhetic acid and 2 moles of glucuronic acid. Licorice roots also contain triterpene, flavonoids, and B vitamins. The powerful sweetening power of licorice (estimated at 50 times that of sucrose) is derived from the glycyrrhizin. In food and drug applications, licorice is used both to enhance and also to mask or subdue flavors, particularly of a bitter nature.

The licorice plant ranges from 4 to 5 feet (1.2 to 1.5 meters) tall and has many pinnately compound, pale-green leaves and purplish flowers resembling those of the perennial pea. Cultivated in southern Europe, the plant also is found wild in eastern Europe. Licorice root and extract are used in nonalcoholic beverages (up to 130 parts per million); in ice cream (200 ppm); in candies (as high as 2500 ppm); in baked goods (200 ppm); in gelatins and puddings (5 ppm); in syrups (50 ppm); and in some chewing gums (as high as 22,000 ppm). Licorice is also frequently used in tobacco products. A synthetic licorice flavoring is made and is the ammonium salt of glycyrrhizic acid.

LIE GROUP. A group which is also an analytic manifold in which the group operations, multiplication and formation of inverse, are analytic. Historically, the first Lie groups were continuous transformations of the points of a manifold. Thus, consider the set of transformations of the points of a Euclidean plane,

$$x_1 = \phi(x, y, a), \quad y_1 = \psi(x, y, a)$$

for some range of values of the parameter a, given by functions ϕ and ψ such that, if two transformations of the set are carried out in succession, the result is again a transformation of the set. An easily visualized example is the group of rotations

$$x_1 = x \cos a - y \sin a, \quad y_1 = x \sin a + y \cos a$$

Lie groups of transformations owe their practical importance partly to their usefulness in systematizing the solutions of differential equations, both ordinary and partial, and partly to the fact that many of the standard groups of linear transformations are Lie groups. For example.

The *full linear group* is the group of all non-singular matrices with complex numbers as elements.
The *real linear group* is the group of all non-singular $n \times n$ matrices with real numbers as elements.
The *unimodular group* is the group of all complex matrices with determinant equal to unity.
The *real unimodular group* is the group of all real matrices with determinant equal to unity.
The *unitary group* is the group of all unitary matrices.
The *unitary modular group* is the group of all unitary matrices with determinant equal to unity.
The *real orthogonal group* is the group of all real orthogonal matrices.
The *rotation group* is the group of all real orthogonal matrices with determinant equal to plus one.

See also **Group.**

LIGAMENT. See **Tendon.**

LIGAND. Any atom, radical, ion, or molecule in a complex (polyatomic group) which is bound to the central atom. Thus, the ammonia molecules in $[Co(NH_3)_6]^{3+}$, and the chlorine atoms in $[PtCl_6]^{2-}$ are ligands. Ligands are also complexing agents, as for example, EDTA, ammonia, etc. See also **Chelates and Chelation.** Examples of common configurations of complex ions in various geometric arrangements are shown by the accompanying diagram.

Ligand field theory incorporates elements from the valence bond theory of Pauling and the molecular orbital method of Hund, Mullikan, and others. As pointed out by Mortimer, the chemists of the late 19th Century had difficulty in understanding how "molecular compounds" or "compounds of higher order" are bonded. The formation of a compound such as $CoCl_3 \cdot 6NH_3$, was baffling, particularly in this case since simple $CoCl_3$ does not exist. In 1893, Alfred Werner proposed a theory to account for compounds of this type. Werner wrote the formula of the cobalt compound as $[Co(NH_3)_6]Cl_3$. Werner assumed that the six ammonia molecules are symmetrically coordinated to the central cobalt atom by "subsidiary valencies" of cobalt, while the "principal valencies" of cobalt are satisfied by the chloride ions. Werner devoted over 20 years preparing and studying coordination compounds and perfecting and proving his theory. Although modern work has amplified his theory, it has required relatively little modification.

In ligand field theory, one is concerned with the origin and the consequences of splitting the inner orbitals of the central metal by the surrounding ligands. The most satisfactory correlations have been demonstrated with the first transition series, in which the $3d$-orbitals are split into different energy levels. To appreciate the effect of a ligand field, imagine that a symmetrical group of ligands is brought up to a charged ion from a distance. First, the electrostatic repulsions between the ligand electrons and those in the d-orbitals of the metal will raise the energy of all five d-orbitals equally. Then, as the ligands approach to within bonding distances, the repulsion interactions will take on a directional character that will vary with the particular d-orbitals under consideration. This arises because of the different shapes and orientations of the five d-orbitals in space along a Cartesian coordinate system. The splitting of the orbitals for a given central metal ion is dependent on the set of ligands.

Applications of ligand field theory to many transition metal complexes have played an important role in the interpretation of visible absorption spectra, magnetism, luminescence, and paramagnetic resonance spectra.

LIGASES. See **Enzyme.**

LIGATURE. A threadlike material or wire used for tying off blood vessels or other structures of the body during surgical operations. The material may be absorbable (catgut) or nonabsorbable (silk, nylon or linen) or metal wire. Ligatures are made in various grades of thickness and tensile strength.

LIGHT. Although the use of the word *light* has been broadened over the decades, general usage still refers to the visible portion of the electromagnetic spectrum. The wavelengths of visible light extend approximately from 4000 to 7000 Å (1 Å = 10^{-8} centimeter). The speed of light in vacuum (symbol c) is generally published as 299,792,500 meters per second ($\sim$186,292 miles per second). The direct determination of the velocity of light, usually performed in air, conventionally has been based upon the measurement of the time for a light pulse to cover a known distance. (See reference to Michelson-Morley experiment in entry on **Laser.**) Such a pulse means an increase, followed by a decrease, of the amplitude of the light vibrations. What is observed is energy exchange associated with the amplitude changes and, as a result, the propagation velocity of light is obtained. A change of amplitude is, however, equivalent to an interference among a series of adjacent wavelengths, inasmuch as that change is created by just such an interference. The light pulse, therefore, consists of a whole group of adjacent wavelengths interfering with each other. Interference is a sum-product. If the participating waves have different velocities, one can find by simple addition of two sine oscillations that the group formed has a velocity different from those of the waves creating the group. An interesting experiment in the use of a laser to measure the speed of light was made in 1962. This experiment was later followed by a number of other, even more sophisticated laser measurements.

A review of the numerous experiments to measure the velocity of light makes fascinating reading for the student of science. An excellent starting review of the experiments of Galileo (1676), Bradley (1725), Foucault (1850), Kohlrausch (1856), Blondlot (1891), Michelson (1879 and 1926), Karolus and Mittelstaedt (1928), Anderson (1940), Jones and Conford (1947), Alaskon (1949), Essen and Gordon-Smith (1947–1950), Bergstrand (1950), Froome (1950–1958), Rank, Bennet and Bennet (1955), Mackenzie (1953), Kolibayev (1958–1963), Mockler (1961), and Cohen (1972) is given in "The Encyclopedia of Physics," (R. M. Besançon, editor), Van Nostrand Reinhold, New York.

Light has a physical character similar to that of radio waves. However, the frequency of light waves is almost a billion times higher and the wavelengths a billion times shorter, than the waves of standard radio broadcast bands. The perception of color depends upon the distribution of the electromagnetic energy over the visible wavelengths. White light is a superposition of waves at many frequencies. It can be decomposed into its monochromatic spectral components by a prism or other spectral apparatus. The violet end of the spectrum is near 4000 Å. The red end is near 7000 Å. Whereas light in its narrow definition should be confined to this relatively narrow portion of the electromagnetic spectrum, in recent years it has become customary to extend the definition to the ultraviolet and infrared portions of the spectrum. One sometimes speaks loosely of ultraviolet and infrared light, although electromagnetic waves at these frequencies are not detectable by the human eye. Instruments and photographic films, however, can be made sensitive to both the shorter and much longer wavelengths. See **Infrared Radiation; Photography and Imagery; Spectrochemical Analysis; Spectro Instruments; Spectroscope;** and **Ultraviolet Radiation.**

The study of the human eye as a detector of light is the task of *physiological optics*. The impression of light is not necessarily always connected with the simultaneous presence of electromagnetic energy at the retina. The eye is capable of creating false images, as when one "sees stars" from a heavy mechanical blow in the dark. The impression of light is retained for about 0.1 second after the light source is shut off. This fact is made use of in motion pictures to create the impression of motion through use of a series of still images. The eye is a detector with a relatively long response time. Photoelectric cells can react more than a million times faster. Color vision is also subject to physiological peculiarities which are quite complex. See also **Vision and the Eye.**

The quantum theory of radiation applies to light, the energy quanta of which are called photons. Some of the optical phenomena so readily interpreted on the wave theory, such as reflection, refraction, interference, diffraction, and polarized light, offer difficulties when studied in terms of quanta. The laws of photoelectric phenomena, photoconductivity, and the spectrum, on the other hand, appear more readily understandable when studied in terms of quanta.

The property of light which is most immediately accessible to observation is its propagation along straight lines. If light rays pass from one medium to another, their direction is changed according to the law of refraction. See also **Mirrors and Lenses; Refraction;** and **Refractive Index.** If light in medium I propagates with velocity v_1 and makes an angle v with the normal to the boundary between media I and II, the direction v in medium II, with a velocity of propagation v_2 is given by Snell's law, $\sin v_1/\sin v_2 = v_1/v_2 = n$. The constant n is called the relative index of refraction of medium II with respect to medium I. These laws are the basis of *geometrical optics*. This branch of the science of light describes the paths of light rays, the formation of images by mirrors and lenses, the action of telescopes, microscopes, prisms, and numerous other optical instruments.

The wave character of light becomes apparent by more refined observations. The phenomena of diffraction, interference and polarization are the subjects of *physical optics*. Diffraction describes how waves are bent around obstacles. They represent corrections to the deviations from the laws of geometrical optics. These effects become pronounced only when the material has a characteristic dimension comparable to the wavelength of the wave. When light waves reach the same point along different paths, the resulting intensity may be smaller than that produced by each individual wave separately. The relative phases of the waves may be such that they interfere destructively, when the arrival of one wave with maximum positive deflection coincides with that of another wave with maximum negative deflection. Observations of light in crystals of calcite (Iceland spar) first showed that there are two different modes of vibrations for each direction of propagation. These are called the two transverse modes of polarization.

All phenomena of geometrical and physical optics are described consistently by Maxwell's equations of electromagnetic theory. Optical phenomena are, therefore, closely related to other electric and magnetic phenomena. Very early in the present century, the prevailing opinion was that the wave character of light was unambigously established and the nature of light well understood.

There was, however, a mathematical difficulty with the intensity of radiation of ultraviolet and higher frequencies. The photoelectric effect could also be interpreted only by considering light to have a quality of particles. The number of electrons emitted from a photosensitive surface is proportional to the intensity of the light. The energy of the individual electrons is, however, determined by the light frequency. This led to the postulate of *light quanta* with energy $h\nu$, where h is Planck's constant. This duality in nature, in which wavelike and particle-like properties are combined, is described without internal contradiction by quantum mechanics. The combined particle-and-wave character of light is revealed by the combination of properties of the light sources, the electromagnetic field describing the light waves, and the detectors.

The combination of the laws of quantum mechanics and electromagnetic theory gives a consistent description of the generation, propagation and detection of light. Since these same laws also describe many other properties of matter, such as electronic structure, chemical binding, electricity, and magnetism, it may be said that the nature of light is well understood. In this context, it is not necessary and not even desirable to pose the question, "What is it, precisely, that vibrates in a light wave in vacuum?" The electromagnetic fields acquire meaning only through their relationships with detectors and sources. Human knowledge or understanding is here used in the operational sense that a relatively simple framework of physical concepts and mathematical relationships exist, which gives an accurate description of the wide variety of optical phenomena at present accessible to observation or verification in experimental situations.

The study of the interaction of light waves with matter in the sources and detectors is the subject of *spectroscopy*. This is a wide field which encompasses atomic and molecular spectroscopy, parts of solid-state physics, and photochemistry. The quantum theory was largely developed on the basis of spectroscopic data. A light quantum is emitted by an excited atom, molecule, or other material system when an electron in such a particle makes a transition or "quantum jump" from a state with higher energy to a state with lower energy. The energy difference between these states is equal to the quantum energy $h\nu$. Similarly, the absorption of light quanta is accompanied by an electronic transition from a state with a lower energy to a state with an energy higher by an amount $h\nu$. In this manner, the frequencies of spectral lines are characteristic for the electronic energy levels in each material. The frequency of the light may be said to correspond to the frequencies of the vibrating charges or oscillators, which are represented by electrons.

Light sources are thus bodies with a sizeable population of electrons in excited states. This may be accomplished by raising the temperature of the material. The most important source of light is the sun. The moon and other planets are visible only because they reflect sunlight, just as all objects on earth which we can see by daylight, but not at night.

Human-engineered light sources range from primitive fire, candles, and oil and kerosene lamps to electric light bulbs, fluorescent-gas discharge tubes, arcs, and many others. See also **Illumination.** In early sources, the material particles of smoke or wick were heated by the chemical reaction of oxidation or burning; in incandescent lamps, a wire is heated to a very high temperature by an electric current. There are so many energy levels in these luminous solid materials or gases at high pressures that the emitted light is essentially white and contains all frequencies. The higher the temperature, the more radiation is emitted and the higher the average frequency of radiation. It should be realized that most of the energy is emitted as invisible (infrared) radiation, even in the better incandescent lamps. Hot gases in flames may also emit sharp spectral lines characteristic of the atoms occurring in the flame. The yellow color which arises when sodium chloride is sprinkled in a flame is due to the characteristic yellow spectral line of sodium atoms.

In gas discharge tubes, atoms or molecules are excited by collisions with electrons in ionized gas. The energy is provided by the generator, which provides the voltage necessary to maintain the discharge current. An arc is a discharge in air or in a high-pressure vapor. Mercury and sodium discharges are used for street lighting. Fluorescent tubes use a gas discharge with a substantial ultraviolet component. This ultraviolet light excites electrons in fluorescent centers on the walls of the tube. The electrons drop immediately from the highly excited state to an intermediate state with a lower energy. From this state, they finally drop down to the original ground energy level, with emission of visible light. Gas discharges at relatively low pressure may serve as spectroscopic sources to study the emission spectra of atoms, ions, and molecules. From the relationship between the energy levels and the frequency of radiation, it follows that a material, when heated, can emit precisely those frequencies which it absorbs when it is in the lower energy level at low temperature.

All of these light sources are incoherent in the sense that there is no phase relationship between the light waves emitted by the different atoms in the source. This is quite different from the property of the usual sources of electromagnetic radiation at lower frequencies. In electronic oscillators used in radio and microwave transmitters, all electrons move and vibrate in step with each other. Unlike the light sources previously mentioned, lasers emit a coherent beam of light. In lasers, the original, spontaneously emitted light forces the other excited atoms to emit their radiation in step, or coherently. If stimulated emission thus dominates the spontaneous emission, a laser results. This requires a high concentration of excited atoms and a sufficient feedback mechanism of light by mirrors. In its simplest form, a laser consists of a gas discharge in a tube of suitably chosen dimensions and gas pressure between a set of parallel mirrors. Because the atoms in the laser source all act constructively in step, these sources provide a more efficient means to transmit light energy.

The high light intensities available in focused laser beams have led to the development of the branch of *nonlinear optics.* The optical properties of materials are different at high intensities, because the electronic oscillators are driven so hard that enharmonic properties become evident. A typical effect is the harmonic generation of light in which red laser light is converted into ultraviolet light at exactly twice the frequency when the high-intensity beam traverses a suitable crystal, such as quartz. It thus becomes feasible to duplicate at light frequencies all nonlinear effects known from the field of radio communications, such as modulation, demodulation, frequencing mixing, among others. It is no longer correct to say that the propagation of a light wave is independent of the presence of other light waves. At high intensities, there is a noticeable interaction between light waves of different frequencies.

Optical Phase Conjugation

Even though the laser enjoys very wide application (see **Laser**), its full potential has not been realized because of distortion that occurs in light waves when they pass through optical systems where inhomogeneities are present. There is a real need for means to reduce or compensate for static and dynamic distortions (noise) which frequently occur. Highpower lasers, tracking systems, atmospheric communication networks, and photolithographic systems, among others, are degraded by such noise. A relatively new area of optical system research, known as *optical phase conjugation*, promises to provide at least a partial solution to this problem.

That light beams possess the property of reversible propagation has been known for many years. For example, assuming a *perfect* optical system, a coherent light beam introduced at point *A* in a system and traveling through a complex of components will exit at point *B* undistorted. If a light beam of exactly the same characteristics were introduced at point *B*, it would travel the same exact path and exit, undistorted, at point *A*. In other words, under perfect conditions, the propagation is reversible. What was not known concerning reverse propagation until the early 1970s was that a distorted (noisy) light beam, if propagated in reverse through a given optical system, will during the course of its backward track remove all the distortions introduced into it during its prior forward track, and thus exit at the point of origin as a "clean" beam, free of distortion. Thus, the concept of reversible propagation holds not just for a theoretically perfect system, but for a practical imperfect system as well.[1] The backward-traveling wave, in essence, is a "time-reversed" replica of the original incident wave.[2]

Producing a Phase-Conjugated Wave. Two general approaches have been investigated thus far for implementing the concept of optical phase conjugation in a practical way for the purposes previously mentioned. These two approaches employ different physical principles, but both rely on the laser light itself to interact with the nonlinear optical properties of a specific medium to initiate phase conjugation or a turnabout of the distorted light waves: (1) *stimulated Brillouin scattering* (SBS) and (2) *degenerate four-wave mixing* (DFWM). When any material—gas, liquid, or solid—is penetrated by light of intensity great enough to compete with the atomic forces that bind the material together, the material is modified, as is also the light penetrating it. This nonlinear interaction generates the SBS or DFWM time-reversed waves. The success in the reversal of wave motion is due to an extreme simplification of the problem: the quantum-mechanical and thermal motions of atoms and electrons that radiate light do not need to be reversed. It is sufficient for practical purposes, as observed by Shkunov and Zel'dovich, to reverse the temporal behavior of macroscopic parameters describing the averaged motion of a large number of particles.

In SBS, the modified material generates sound waves that serve as an appropriate reflective surface to produce the time-reversed waves.[3] In DFWM, the interaction uses a holographic process in a nonlinear material to generate the conjugate, or reversed light waves.

Examples of nonlinear mediums include semiconductors, crystals, liquids, plamsas, liquid crystals, aerosols (as in the atmosphere), and atomic vapors. The term *nonlinear*, as used here, pertains to media that are altered or affected by light. Linear materials, in contrast, are not so affected.

Prospective Uses of Optical Phase Conjugation. Currently, a number of scientific laboratories, in addition to those in Russia, have been conducting intensive research in this area in attempts to refine the technology and to find practical applications. Among the active laboratories are the California Institute of Technology, the AT&T Bell Laboratories, Philips Research Laboratories (the Netherlands), the University of Southern California, the University of Waikato (New Zealand), the University of Arizona, the National Institute of Standards and Technology, IBM, and Hughes Aircraft Research Laboratories. A number of scientists envision several important applications within the next decade. Admittedly, the field remains in an investigative state. Some potential applications as reviewed by Pepper (see reference) and others are reviewed briefly here.

Some scientists have suggested that phase-conjugated systems may be used for image transfer in photolithography. Researchers at IBM have demonstrated image transfer. Light from a laser passes through the mask pattern, a semitransparent mirror, and then an amplifier. The intensity of the beam increases, but at the cost of introducing distortions into the beam. When the image is returned through an amplifier by a phase-conjugating mirror, the "time-reversed" beam is both powerful and free of distortions. Conventional methods, such as compensating for optical aberrations, are no longer required. Another way in which lensless-imaging schemes may be used include fiber-optic communications and associated memory as used in pattern-recognition devices. As

[1]In an experiment at the P. N. Lebedev Physical Institute (Moscow) in 1972, Boris Ya Zel'dovich and coworkers observed a curious phenomenon while doing an experiment. The researchers intentionally distorted an intense beam of red light from a pulsed ruby laser by directing it through a frosted glass plate. The degraded beam was directed down a long tube filled with methane gas under high pressure. Interactions occurred between the beam and the molecules of the gas (stimulated Brillouin scattering) and, acting as a mirror, the gas reflected the beam backward. The investigators were surprised to find that once the reflected wave passed back through the same piece of frosted glass, a nearly perfect, undistorted optical beam emerged. Thus, they found that the distortions introduced by the glass during the forward passage were, in essence, canceled out during the backward passage. (The phase difference between any two points of the reversed beam has a sign opposite to that of the phase difference between the same points of the original beam. As described by Shkunov and Zel'dovich, the mathematical operation of changing a phase sign is known as conjugation, and thus the coining of the term optical phase conjugation.)

[2]During the early phases of explaining this unexpected phenomenon, a number of homely analogies were developed, the most common being that of comparing the retracing of light waves back through the distorting media as "making a film run backward." As described by Shkunov and Zel'dovich, "The relation between the wave fronts of two mutually reversed waves is anlaogous to the relation between the positions of two opposing armies on a military map. The front line of each army coincides with that of the other, and the directions of desirable movement are opposite. One can say that the front lines are mutually reversed: a convex part of one army's front corresponds to a concave part of the other."

[3]Traditionally, the frequency of scattered light has been regarded as identical to that of the incident light. In actual practice, as first predicted by Brillouin in 1914, a slight line broadening occurs due to motion of the scatterers (Doppler effect) and also due to variations in the directions or magnitudes of their polarizability tensors (due to chemical reactions). The Brillouin effect, simply stated, is as follows: upon the scattering of monochromatic radiation, a doublet is produced, in which the frequency of each of the two lines differs from the frequency of the original line by the same amount, one line having a higher frequency, and the other having a lower frequency.

stressed by Pepper (see reference), the emergence of optical phase conjugation has unified many areas of applied and fundamental optical physics. Spectroscopy, the study of the interaction of matter and radiation, has particularly benefited. The concepts, techniques, and basic applications of optical phase conjugation can, in principle, be applied to most other areas of the electromagnetic spectrum. Microwave phase conjugation would have major applications in radar, millimeter-wave imaging systems, and high-frequency temporal signal processing, as well as microwave spectroscopy. Researchers are planning experiments in the acoustic-wave area. Acoustic signal-processing devices and sonar may benefit from such research.

Squeezed Light. Although research has been proceeding for over a decade, the study of "squeezed" light is still essentially in an experimental stage. For a number of years, scientists have recognized that a beam of light is not free from random fluctuations, but in electronic terms it is noisy. Light is used frequently to make measurements and to observe numerous instrumental phenomena. Thus, faulty light contributes to errors, if ever so small. This situation is explained by the quantum theory, which implies that light must be accompanied by a certain minimal amount of light fluctuation.

A beam of light has been defined as an oscillating electromagnetic field, which can be viewed as a smooth wave. The shape of the wave can be foretold with absolute certainty, but its slope must fit within a particular "envelope" of uncertainty. Even in darkness, quantum physicists would allow that the wave exists but is flat, with some degree of uncertainty.

Researchers Slusher and Yurke (AT&T Bell Laboratories) have observed, "In practical terms, this means that even in a vacuum, with no external light sources, there must still be small fluctuations in the electromagnetic field. Noise limits the precision of spectroscopy, in which the frequency and intensity of the radiation emitted by atoms or molecules yield information about their properties." It also is envisioned that quantum noise will also limit those technologies concerned with optical computing and communications.

In early investigations, researchers at several laboratories had experimented with so-called *squeezed light*. This field of study involves quantum fluctuations, and largely stemmed from studies of the coherent light generated by a laser. A number of theoretical physicists nearly 30 years ago studied the statistical properties of coherent light. Even the presumed perfectly coherent light of an ideal laser was found to have a Poisson distribution of photons rather than a single, well-defined number. Thus, although not so noisy as a conventional incoherent light beam, the laser is also noisy and, as lasers are employed more frequently in very sophisticated research and practical applications, optical devices could ultimately be limited by such noise.

As related by Robinson (see reference), for squeezed states, the statistical variance of the photon number is not so important as the variances of the amplitude of the electric field, which is related to the photon number, and of its phase. Researchers at AT&T Bell Laboratories have successfully "squeezed" noise or unwanted fluctuations in *phase* at the cost of increasing it in amplitude and, conversely, they have also squeezed noise in *amplitude* at the cost of increasing it in phase. In essence, these researchers generated the squeezed light by using a technique known as *four-wave mixing*, previously mentioned under optical phase conjugation. The researchers directed two laser beams at each other from opposite directions. The laser beams met in a material (nonlinear medium). In their experiment, the nonlinear medium was a beam of sodium atoms oriented at right angles to the laser beams. The research team found that the laser beams interacted with the Na atoms in such a way that two output beams emerged in opposite directions along an axis tilted at a slight angle with respect to the axis of the two input laser beams. Attributed to the properties of the nonlinear medium, when the output beams were combined by mirrors, the result was a single beam of squeezed light. By varying the position of the mirrors that direct the beams emerging from the nonlinear medium, it was found possible to reduce the noise in either the phase or the amplitude of the squeezed light. The degree of success of the experiment was relatively small—about a 7% drop in noise. More recently, other researchers (University of Texas) have reported a 42% reduction in noise.

Practical applications of squeezed light must await further research findings.

Subwavelength Illumination. Superresolution light microscopy permits researchers to optically study specimens without being limited by the diffraction properties of visible light. This technique, however, requires the efficient emission of light from *subwavelength* light sources. In an experimental setup, an electromagnetic wave is generated to emerge from an aperture. First, the wave is highly collimated to the aperture dimension. As pointed out by K. Lieberman and coworkers (Hebrew University, Jerusalem), "It is only after the wave has propagated a finite distance from the aperture that the diffraction that limits classical optical imaging takes effect. Thus, in the near-field region, a beam of light is present that is largely independent of the wavelength and is determined solely by the size and shape of the aperture." In a 1990 paper (see reference listed), an approach for producing sources of light with subwavelength dimensions is described.

The methodology is based on the packaging of photons as molecular excitons, effectively reducing the volume of the light beam by 10^9 and making possible propagation through dimensions of 1 nanometer. The researchers further observe, "Molecular microcrystals are grown in the tips of micropipettes that have inner diameters of 100 nanometers or less. Measurements are presented that demonstrate this improvement in transmission for pipettes of various diameters. The ultrasmall dimensions of these light sources, the wavelength range (UV to IR) of their emission, their ease of production, and their expected unique abilities for high efficiency excitation-imaging of surfaces portend significant applications for this methodology."

Light Reflections in Computer Graphics. The technology of computer graphics has enabled the production of objects in realistic three dimensions from two dimensions, a technique that frequently has replaced the need for clay models in a number of fields, notably in the automotive and appliance fields and for creating realistic backgrounds in simulators as, for example, in aircraft flight training. To achieve realism of painted, metal, glass, and other reflective surfaces, "reflection" algorithms can be created.

Reflection patterns vary markedly from one type of product surface to the next. These patterns are unique for given materials or material families, including, for example, steel, chromium, rubber, and plastics, because of the wide range of diffuse and specular reflections for given materials. Through careful experimentation, appropriate algorithms can be developed. Considerable detail pertaining to these methodologies is given in the D. P. Greenberg reference listed. See also Hurlbert Poggio reference listed.

Extending the Light-Sensing Range. Since so-called "night vision" sensors were developed several decades ago, the practical use of the infrared (IF) portion of the light spectrum has continued in its importance, particularly in military operations such as tank and helicopter maneuvering. Traditionally, these sensors have depended upon mercury cadmium telluride (*mer-cal*) semiconductors. These sensors pick up wavelengths between 3 and 5 and 8 and 12 microns. A greater sensitivity over the whole IR range has been sought. The materials for these sensors also have been difficult to process. After some years of research, a group at AT&T Bell Laboratories has developed ways to produce gallium arsenide optical crystals that are claimed to approach atomic perfection. The new detectors, which have greatly increased sensitivity, feature what is known as a quantum well, which precisely controls the flow of electrons. Experimental arrays with more than 16,000 pixels have been built and tested. Such detectors can sense temperature differences as small as a ten-thousandth of a degree. Cool objects appear dark, with warmer objects ranging over a gray scale. The quantum-well infrared photodetectors (QUIPs) now are being studied in terms of mass producing them.

Sky Colors. As early as 1899, Rayleigh claimed that the sky is blue on a clear day because of scattering and color separation by the molecules contained in the air of the atmosphere. Rayleigh viewed the phenomenon as a composite of all the colors in the visible spectrum. See also **Scattering**. He proposed that short-wavelength light (blue) is scattered more than red light, which has a longer wavelength. Based upon the ratio of the blue and red wavelengths (~ 1.68), Rayleigh reasoned that the blue-scattered light is approximately eight times that of the red-scattered light.

Sunsets are reddened because the light reaching the observer has passed through a much longer path, allowing domination of the red end of the spectrum in contrast with the situation when the sun is viewed high in the sky.

Other investigators have attributed sky color to layers of ozone in the upper atmosphere. It has been postulated that, in these layers, molecules have absorption bands that favor the absorption of light at the red end of the spectrum. A number of postulations have emerged pertaining to the development of several colors, including the purple sky coloration. These are explored in more detail in the Meinel reference listed.

Some correlation of sunset color with volcanic eruptions has been made. As examples, the explosion of Krakatau (near Java) in 1883 produced brilliant sunsets over a span of about 5 years; The explosion of Agung on Bali in 1963 affected sunset coloration for about 3 years.

Optical Computer

The pace of research to develop an optical computer based upon beams of light rather than electric currents have increased markedly during the past decade. An optical analogue of the transistor was demonstrated in the early 1980s. An optical computer that might operate a thousand times faster than contemporary electronic computers is indeed a powerful incentive. Theoretically, the operations carried out by a computer (logical and arithmetic) could be effected by numerous means and, in fact, this already has been demonstrated when one considers that early computers depended exclusively on vacuum tube switches. Although the transistor that replaced the tube was a great step upward in electronics, the transistor was a significant hardware departure—thus indicating that computer designers will quickly adapt to new hardware when there are significant demonstrable advantages. Thus, among some scientists, there is an attitude that tends to regard the optical computer as an inevitable development for the future. Some forecasts indicate that an optical computer may be capable of a trillion operations per second. Even though the current electronic computers are undergoing significant changes in the way they are organized to process information, such reorganization today probably would be easier to achieve if the system were basically optical rather than electronic—as considered by some contemporary scientists. See also **Digital Computer.**

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LIGHT-ACTUATED CELL. See **Photoelectric Effect.**

LIGHT-COUPLED SWITCH. A switch in which the switching signal is transmitted to the device in the form of light energy. The switching element may be a phototransistor, a photodiode, or photofield-effect transistor (FET). The receipt of light energy by such a device changes the transmission characteristics of the device, permitting conduction between two terminals. When no light is present, the resistance of the device is high. When excited by photon energy, the resistance drops to a much lower value. Various light sources are used in connection with light-coupled switches. Although common incandescent sources may be used, gas-discharge sources, such as neon lamps and solid-state devices, such as gallium arsenide light-emitting diodes, are more commonly used in instrumentation systems. As compared with incandescent sources, the other sources produce less heat and possess higher speed and reliability.

A major advantage of the light-coupled switch is isolation that can be obtained between signal and drive source. Signals can be controlled without introducing errors due to the drive source. Further, isolation helps in maintaining a high common-mode rejection ratio in analog signal-handling equipment.

See also **Analog Switch.**

LIGHT CURVE (Astronomical). In the study of variable stars, and in kindred problems in astronomical research, it is desirable to represent graphically the variation of radiation intensity with time. A diagram in which light intensity, on any convenient scale, is plotted as ordinate against time as abscissa is known as a light curve. As the number of observations increases, it frequently becomes possible to detect a periodic variation in the light intensity. After a provisional period has been determined, some convenient epoch is selected, and all the observations are reduced to the cycle of variation embracing the selected epoch by the use of the provisional period. In order that the resulting points may fall on a regular curve, it is frequently necessary to apply a number of corrections to the provisional period. The curve drawn through the plotted points, all reduced to the selected epoch by means of the repeatedly corrected period, is known as the mean light curve. Examples of light curves will be found in the articles on **Cepheids; Eclipsing Binary;** and **Variable Star.**

LIGHT (Effects on Plants). See **Photoperiodism.**

LIGHT-EMITTING DIODE (LED). See **Automotive Electronics; Luminescence; Telephony.**

LIGHTER-THAN-AIR SHIP. See **Dirigibles and Airships.**

LIGHTING. See **Illumination.**

LIGHTNING. Under favorable circumstances, large electrical potential differences that are generated in the earth's lower atmosphere may be neutralized within an extremely short span of time in the form of lightning. The lightning may occur fully within the atmosphere, as by intra- or intercloud lightning, or between the atmosphere and the earth's surface, particularly the higher surfaces (mountains) or objects such as buildings and trees which extend upward from the ground for short distances. Favorable paths of electrical conductance predominantly determine the geometry of lightning. It has been observed that conductivities and air-earth current densities may be as much as ten times greater in the vicinities of high mountains than at sea level. This correlates well with other observations which indicate that lightning at sea occurs with a frequency only about 10% of the frequency over land. Conductivity decreases when humidity increases. Considering the earth as a whole, lightning is very common, occurring with a frequency of about 100 lightning events per second. Scores, hundreds, and even thousands of lightning events may occur in connection with a given thunderstorm. The frequency and variation of lightning characteristics ranges widely— with location, season of the year, and local time of day. Lightning over land occurs with the greatest frequency during late afternoon and early evening; and over the seas and oceans during late evening to an hour or two after midnight local time. There is, however, no time of day or night when lightning may not occur. Lightning can strike the same object or location not just once, but many times even during a given storm. This is particularly true of high structures. Trees do not provide protection against lightning, but represent an unsafe location for persons to be during a lightning-productive storm. Some authorities have suggested that a metal-bodied automobile may be one of the safest places to be during a storm of this type if other hazards (falling trees, flying debris, etc.) are minimal. For persons caught in open fields, golf courses, etc., lying prone on the ground is considered best in the interest of protection from lightning.

Although lightning is an extremely common phenomenon of the lower atmosphere, it remains rather poorly understood, particularly in quantitative terms. In recent years, lightning has been found to be much more complex and variable than previously believed. High-speed camera techniques, notably with improvements that permit the analysis of lightning during daytime hours, have produced much new evidence both of a qualitative and quantitative nature. The unaided eye or still camera, incapable of breaking down the individual events which occur during what appears to be a single stroke of lightning, at best can provide very rough data—estimates of patterns, brilliance, and distances and heights. In recent years, concerted programs of research in this field have contributed much toward expanding the knowledge of lightning. Such efforts include work done in connection with the Thunderstorm Research International Program (TRIP), in which several organizations such as the Kennedy Space Center (NASA), the University of Florida, the University of Arizona, the New Mexico Institute of Mining and Technology, Rice University, and the State University of New York at Albany, among others, have participated. Uman et al. (1978) reported on the physics and meteorology of a lightning flash at Kennedy Space Center; Orville and Lala (1978) reported on the development and use of a streaking camera for producing daylight time-resolved photographs of lightning.

Characteristics of Lightning

From the standpoint of geometry, the fundamental types of lightning are: (1) cloud-to-cloud (intercloud); (2) between two portions of the same cloud (intracloud); (3) cloud-to-earth; and (4) earth-to-cloud. The last two types often occur as part of what visually appears to be a single event. Lightning also may be classified as (a) long time span and low current; and (b) short time span and high current. The first of these is generally the more damaging to objects, such as igniting forest and structural fires.

Rarely do the long-time-span lightning events persist up to or over one second, although it was reported by Godionton in 1896 that a single lightning flash lasted up to 15 or 20 seconds. If this observer was accurate, it is indeed an exceedingly rare case. Observations since then have not included events of this time magnitude. Most lightning events persist for fractions of a second, ranging from 10^{-2} second through an average of about 0.2 second, up to 1 second (unusual).

In the usual cloud-to-ground lightning event, the phenomenon commences with what is called a *stepped leader*. High-speed camera techniques have been invaluable in confirming this characteristic. See the very schematic sketch in Fig. 1. The stepped leader may be of relatively low luminosity and possibly one to two meters in diameter. The leader sets up the conditions for the much more dramatic *return stroke* from earth to cloud. The time interval between steps of the stepped leader may range from a minimum of 30 microseconds to 50 microseconds (typical) to a maximum of 125 microseconds. The minimum length of a step may be as low as about 3 meters, ranging up to 50 meters (typical) to a maximum of about 200 meters. The average velocity of propagation of the stepped leader, at a minimum, may be 1.0×10^5 meters/second, ranging up to 1.5×10^5 meters/second (typical) to a maximum of about 2.6×10^6 meters/second. The charge deposited on a stepped-leader channel may be as low as 3 coulombs, ranging up to 5 coulombs (typical) to a maximum of about 20 coulombs. The return stroke may have a diameter measurable in terms of centimeters rather than meters, with a peak current ranging from 10 to 20 (typical), but up to a maximum of 110 kiloamperes. The channel length may range from 2 kilometers up to 5 kilometers (typical) to a maximum of about 14 kilometers. The velocity of propagation may range from 2.0×10^7 meters/second to 5.0×10^7 meters/second (typical) up to a maximum of about 1.4×10^8 meters/second. Temperatures within the return stroke range up to 30,000 K at pressures up to one MN/m^2 stroke.

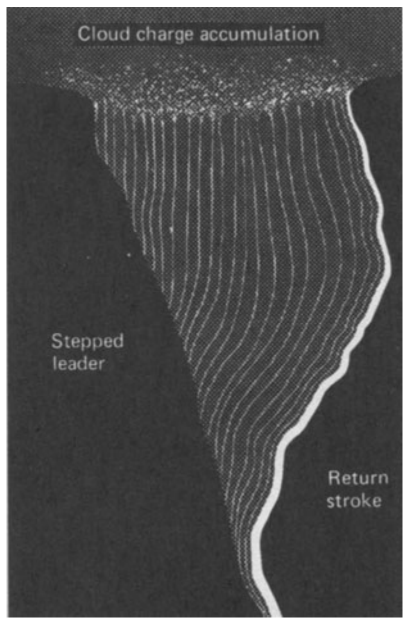

Fig. 1. Stepped leader and return stroke in cloud-to-earth lightning event.

Upon completion of the return stroke, a dart leader may again move downward because of residual potential differences, causing a second return stroke. As many as 40 secondary discharges (return strokes) may occur in what is called a multiple stroke flash. See various lightning configurations in Figs. 2 and 3.

Energy Sources of Lightning

Numerous theories have been advanced pertaining to the accumulation of electric charge in the lower atmosphere. Advanced as early as 1885, the influence theory suggested that the earth's field is usually negative with relation to positive cloud charges, thus setting up the right conditions for lightning. This concept was probably in keeping with the general scientific knowledge of that time. It was also generally proposed that particles (hydrometeors, such as rain, snow, hail, fog, etc.,

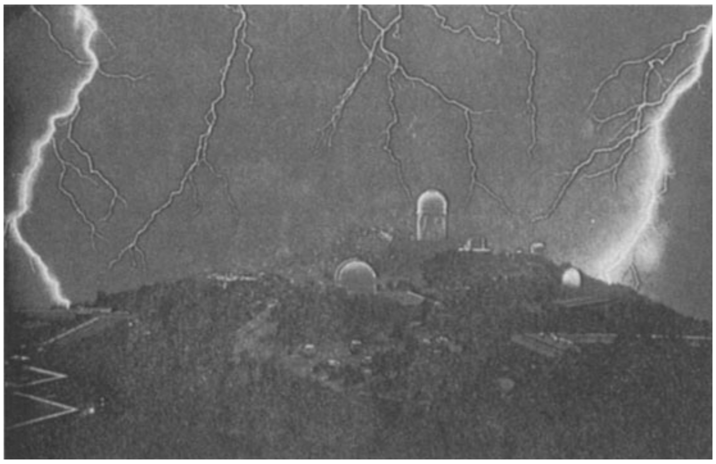

Fig. 2. Dramatic display of several lightning strokes in vicinity of Kitt Peak, Arizona. Buildings of the National Optical Astronomical Observatories are shown in the light of the lightning. *Nikkormat* FTn with 28 mm lens with tripod and cable release; *Kodachrome II* film; *f*3.5; exposure, about 1 minute. (*Copyright Gary Ladd, 1972.*)

Fig. 3. Massive display of lightning over city in western United States. (*Electric Power Research Institute.*)

associated with storms) developed a charge as the result of frictional forces with the atmosphere. In the free ionization theory, it was suggested that droplets of different sizes selectively collect available ions, thus establishing large potential differences. It is now generally believed that electric conduction current in the atmosphere is carried almost exclusively by fast ions. Positive ions include $(H_3O)^+(H_2O)_n$ and negative ions include $O_2^-(H_2O)_n$ or $CO_4^-(H_2O)_n$, where the probable values of n for positive ions in the troposphere are 6 or 7; in the stratosphere, 4; and in the mesosphere, 3 or 2. The value of n is dependent upon the water vapor pressure and temperature of the atmosphere, as well as the lifetime of the ion.

A few highlights pertaining to the energy in lightning would include:

1. Typical voltage drop in ground or other conductors after lightning impact in the neighborhood of 10 kV (dangerous!).
2. Energy delivered to an average stroke of lightning, about 100 kJ/m. Intracloud lightnings have been observed up to 100 km in length. Height from which a lightning stroke points at a target estimated several dacameters, depending on conductivity distribution in ground.
3. Long-lasting, low-current (hundreds of amperes) flashes are more dangerous to people and objects (for example, forest fires) than short, high-current flashes.

Damage Wrought by Lightning

The natural high-voltage phenomena occurring during lightning are not only of general scientific interest and value. A greater knowledge of these phenomena is of practical value as well, in particular because of the danger from lightning to life and to susceptible structures, notably electric power systems and communications equipment. For many years, the direct and induced effects of lightning discharges have been simulated in industrial laboratories by means of high-voltage impulse generators. These frequently use the Marx method of first slowly charging a number of condensers in parallel and then suddenly connecting them in series by spark-gap switches which, at the same instant, impress the multiplied voltage upon the test circuit. A typical voltage wave produced by such impulse generators rises to its peak value of several million volts in one microsecond and then diminishes exponentially, reaching half voltage in about ten microseconds.

The Electric Power Research Institute (Palo Alto, California) has for a number of years conducted research on the characteristics of lightning, means of predicting lightning, and means of protecting electric utility equipment from damage, particularly overhead transmission lines. Although improvements in calculating lightning performance of transmission lines have been achieved, little information has been available for accurately predicting the lightning flashover performance of multiple-circuit transmission lines. See Fig. 4. In most double-circuit lines, when one circuit flashes over, the second circuit will simultaneously flash over 40–60% of the time. EPRI has generated a computer program, known as MULTI-FLASH and available to electric utility operators, that enables the transmission line designer to accurately predict the lightning performance of lines containing up to 12 ac phases, 12 dc poles, or any combination thereof on the same transmission tower, with a variety of tower shapes and insulator strings. All transmission voltage and significant corona effects are included, as well as the statistical distribution of footing resistance. The output includes an analysis of expected shielding failure performance, followed by a detailed tabulation of the expected flashover frequency of each of the phases for dc poles that are involved. Using this computer analysis, the designer can explore new and innovative shapes of materials and accurately predict the degree of improvement that may be achieved.

Research also has been conducted pertaining to arresters. These devices play a vital role in protecting substation equipment from high-voltage surges that originate from either substation switching equip-

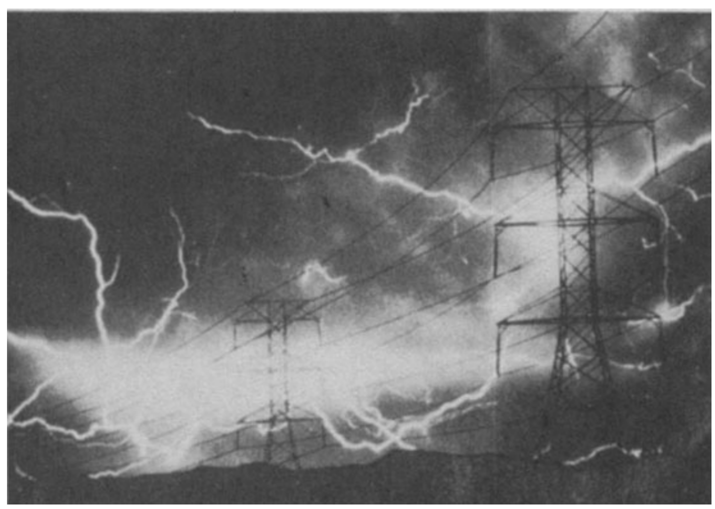

Fig. 4. Much research in recent years has been conducted to understand and to reduce the effects of lightning flashover in multiple-circuit transmission lines. (*Electric Power Research Institute.*)

ment or lightning strikes. Through the proper selection of arresters, undesirable overvoltages can be limited. New materials having an extremely nonlinear voltage/current characteristic have been used in the construction of gapless surge arresters. Materials include zinc oxide (ZnO). The new designs have been under evaluation, including some at Tennessee Valley Authority substations. See Fig. 5.

A few years ago, some utility system designers concluded that surge protection devices could not be further optimized without an improved database of lightning stroke characteristics. The EPRI thus established a project along these lines. In this study, lightning stroke current and energy magnitude were estimated by comparing the gaps of 2800 surge arresters that accumulated 32,000 arrester years of service on utility lines with gaps that had discharged known currents. The stroke energy statistics obtained are of particular interest to surge protection engineers because little information along these lines has been available. Under present standards, arrester durability is determined in part by tests using 65-kA and 8 × 20-microsecond current discharges which have a charge of 1.25 coulomb. The study showed that arresters in service are more frequently exposed to longer-duration, lower-current waves having charges of up to 4.2 coulombs, than to higher-current, lower-energy waves.

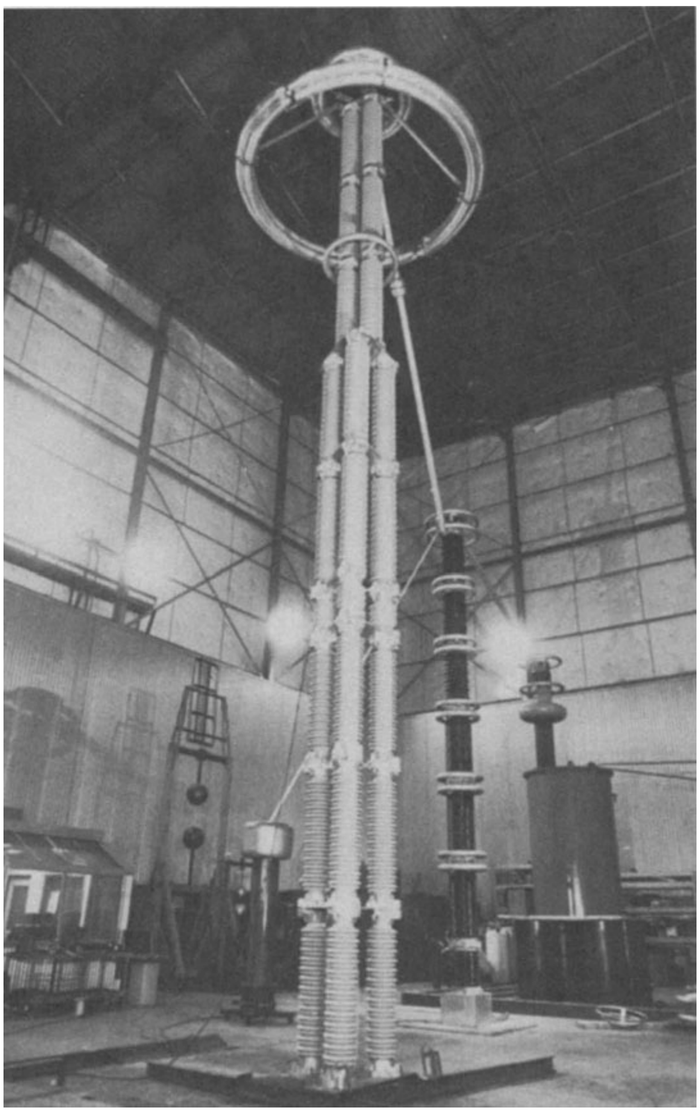

Fig. 5. New materials, including metal oxides (ZnO), have been used in innovative surge arresters to protect electric power substation equipment from high-voltage surges that originate either from substation switching equipment or lightning strokes. The new materials have extremely nonlinear voltage/current characteristics that lend themselves well to gapless surge arrester designs. (*Electric Power Research Institute.*)

In another, associated study, the geographic density of lightning flashes was probed. As a starter, EPRI has mapped the location of lightning flashes in the eastern United States. Such data will provide surge protection engineers with information on the number of lightning flashes that can be expected to strike distribution and transmission lines directly or nearby. Data on charge polarity, number of strokes per flash, and peak field strength radiated from the first stroke are being gathered.

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LIGHTNING BUG. See **Firefly.**

LIGHT (Phosphorylation). See **Phosphorylation (Photosynthetic).**

LIGHT (Polarized). See **Polarized Light.**

LIGHT POLLUTION. Today, people who live near large cities have lost much of their view of the universe. The spectacular view of the night sky that their ancestors had above them on clear dark nights no longer exists. The great increase in urban population has caused an ensuing rapid increase in urban sky glow due to outdoor lighting, which has brightened the heavens to such an extent that the only view most people have of the Milky Way or most stars is when they are well away from cities. This excess light in the sky has had an adverse impact on the environment and seriously threatens to remove forever one of mankind's natural wonders—the dark sky.

The extent of vast illuminated cosmopolitan areas of the United States is dramatically illustrated by the "map" of Fig. 1.

Effects on Professional Astronomy. While this increased urban sky glow brightens the night sky for the general public and for amateur astronomers, it is a special threat to professional astronomy. Advances in frontier astronomy require observations of very faint objects that can be studied only with large telescopes located on prime observing sites, well away from sources of air pollution and from urban nighttime sky glow. For example, most observations of cosmological interest deal with extremely remote sources—galaxies or quasars at such distances that the light has traveled several billion years, sometimes twice the age of our solar system, before reaching us. This light is then often lost in the glare of anthropogenic sky glow.

Observations of such extremely faint extragalactic sources, and even of many objects of interest in our own galaxy, can be done only during the dark moon period. The sky background is much too bright when moonlight is present. Artificial lighting of the sky due to cities has, unfortunately, the same adverse effect on nighttime sky brightness in limiting observations. An interesting contrast in the situation at Tucson, Arizona, which is located near the Kitt Peak National Observatory, is given in Fig. 2. The worsening of the problem over a 20-year period is demonstrated.

This increased sky glow which adversely affects the environment and compromises astronomical research is called *light pollution*, for it is

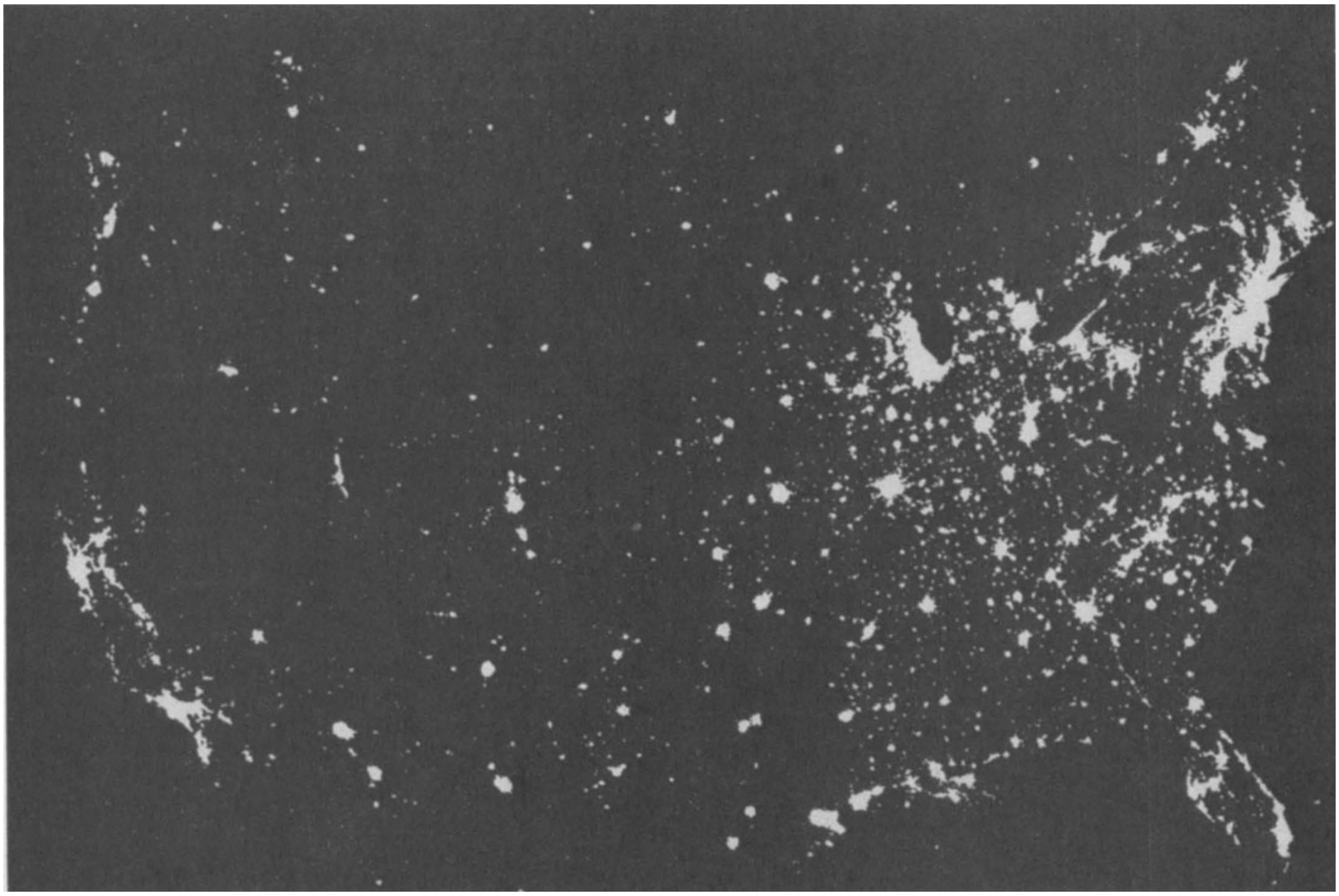

Fig. 1. Satellite view of the United States at night, on a night when most of the country was clear. The upward light produced by the urban areas is evident in the photograph. Some of the light is direct-up light; the rest is light reflected from the ground. (*United States Air Force.*)

wasted light that does nothing to increase nighttime safety, utility, or security. It only serves to waste energy and money. An example of an improperly designed billboard that pollutes the night sky is given in Fig. 3.

Ground-Based Astronomical Observations Still Much Needed. The argument that all astronomy can be done from space is not correct; the largest telescopes will continue to be ground-based for a long time because of cost factors. It does not make sense to do in space, at much higher cost, what can be done from the ground. There are many things that can only be done in space and the demand for that type of research is severe. The experience of more than two decades of space astronomy, however, has greatly increased the demand for ground-based facilities. Planning for several ground-based telescopes much larger than anything now in existence is already underway. There are exciting times ahead for astronomers, using present and future ground-based telescopes, which *complement* the telescopes in space. See **Telescope.**

Solutions to the Light Pollution Problem. Control programs are underway now in a number of communities to reduce the effects of light pollution of the night sky. Programs like these are critical to the long-term success of astronomical research, and to preserve people's view of the universe. Unlike dealing with the problems of water and air pollution, vast sums of money are *not* required to alleviate light pollution.

At present, the lack of awareness rather than resistance is generally the biggest problem in controlling light pollution. Educating the public, government officials and staff, and lighting professionals has been the major thrust of the current programs. These efforts have helped. The increase of light pollution near major observing sites is moderating. More can and must be done. Astronomers, amateurs and professionals, and many others are urging such cooperation.

Astronomers are *not* against night lighting. They have the same needs for quality lighting as everyone else. They advocate the best possible lighting for the task, with lighting designs that allow for all the relevant factors, such as glare control, efficiency, and the need for dark skies. An important added advantage is that everything that is done to minimize light pollution also saves energy by improving the efficiency and utility of the nighttime lighting.

There are other adverse effects of poor quality lighting. Light that comes out of a fixture essentially horizontally does little or nothing to light the ground, but it does cause glare. Such glare is annoying to the eye of the beholder, and it even can cause discomfort or disability. Its blinding effects have often even caused accidents. Glare never helps; it is always a sign of poor-quality lighting.

Confusion or clutter is another adverse effect of poor lighting. Nighttime lighting should provide guidance, providing help and safety, not confusion. Some installations mislead a driver, for example, leading to accidents. In addition, the clutter of outdoor lights that we see today in most cities is often just as trashy a sign of poorly controlled urban growth as is the litter of garbage we see.

Light trespass is another adverse effect of poor outdoor lighting. Light from a fixture that falls in someone else's yard or in through their window is usually unwanted. It is indeed "trespass." This results in black paint on one side of the fixture, irate phone calls to the owner of the light or to the police or to the sports park owners. Such trespass is the sign of a poorly shielded light fixture, not of a quality lighting design or installation. There is no need for such an adverse effect.

It is a sad state that many people are not aware of quality lighting. Many even think that lighting does not exist if it does not exhibit the adverse effects, for they are so used to the associated glare and light trespass of the all too common poor lighting. Quality lighting does exist. It should always be used. There are many examples of it being used,

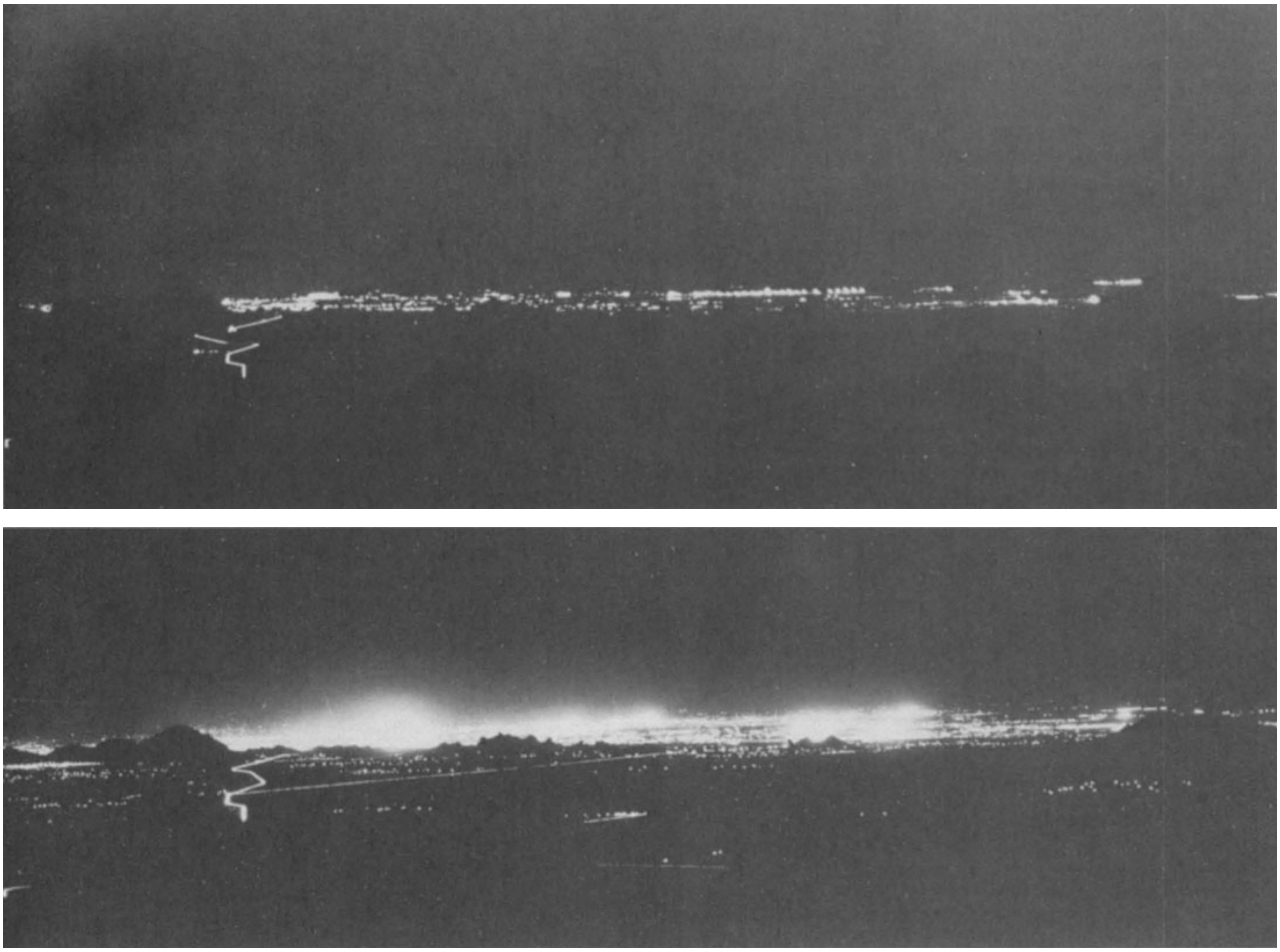

Fig. 2. Two photographs of Tucson, Arizona as seen from Kitt Peak, (about 35 miles 56 kilometers) west of the city. The growth of the city over 21 years (top view, 1959; bottom view, 1980) is evident. It is this type of growth, with its associated growth in outdoor lighting, that is comprising the research done at most of the major observatories located near cosmopolitan areas. (*National Optical Astronomy Observatories.*)

and it should be used for *all* installations. Professional lighting designers are well aware of quality lighting, and use it whenever possible. Unfortunately, so much of today's outdoor lighting is not done by lighting professionals, or by people aware of or sensitive to quality lighting. That is what must be changed.

Specific Corrective Measures. What aspects of quality lighting can be used to help solve the light pollution problem? Following are some solutions that will greatly minimize light pollution without in any way compromising nighttime safety, security, or utility:

1. Use night lighting only when necessary. Turn off lights when they are not needed. Timers can be very effective. Use the correct amount of light for the need, not overkill. Until recently, when energy conservation became an important issue, it often seemed that the only "design" criterion used for outdoor lighting was: "If a certain amount of light is good, double it and things will be better." That is not good design, as any professional lighting designer will agree.

2. Direct the light downward, where it is needed. The use and effective placement of well designed lighting fixtures will achieve excellent control. Shielding the light source to avoid the upward light helps. When possible, retrofit present poor fixtures, ones which spray light everywhere, especially directly up into the sky. In all cases, the goal is to use fixtures which control the light well, minimizing glare, light trespass, and light pollution. All of this also minimizes the energy waste. Light is used when and where needed, and not wasted.

3. Use low pressure sodium (LPS) light sources whenever possible. This is the best possible light source to minimize adverse sky glow effects on astronomy. LPS lamps are the most energy efficient light sources that exist. Areas where LPS is especially good are street lighting, parking lot lighting, security lighting. There are some applications where LPS should not be the sole lighting source, for applications where color rendering is critical. But, for most applications, LPS should be considered. It is an excellent, low-cost light source and helps greatly to minimize the adverse sky glow.

4. Avoid growth near the observatories, and apply rigid controls on nighttime lighting when such growth is unavoidable. Such controls do not compromise safety, security, or utility. Lighting ordinances have been enacted by many communities to enforce quality, effective lighting. These communities have found that the quality of lighting has improved, usually at lower cost.

All of these solutions to the problem of adverse nighttime lighting say, really: "Do the best possible professional lighting design for the task. Include all relevant factors, such as glare, light trespass, and light pollution." All the solutions needed for protecting astronomy have positive side benefits of maximizing the quality of the lighting, and of saving energy. See also **Illumination.**

The American Astronomical Society has a Committee on Light Pollution. The Illuminating Engineering Society of North America has a Committee on Light Trespass and another one on the Environmental Impact of Outdoor Lighting. Other groups are responding in a similar matter. Fortunately, lighting technology is advancing and there is an increasing interest in quality lighting. Such advances in technology, and increasing awareness of the problems of light pollution, will help

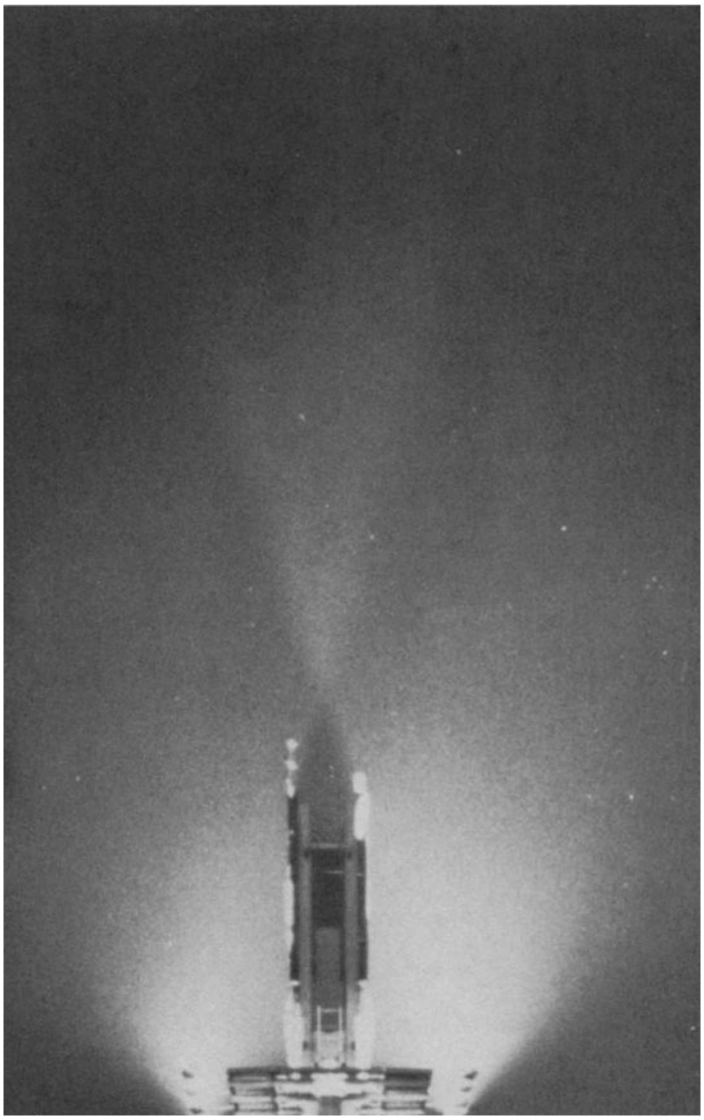

Fig. 3. Photograph of a billboard, showing a typical example of poor lighting. Most of the light output by the fixture seems to miss the billboard. Better quality fixtures would put most of the light onto the board, thus less waste and less light pollution, all at less cost. To minimize light pollution, the quality fixtures should be mounted at the top of the board, thus minimizing the direct and reflected upward light. Also, the lights should be time-clocked to go off at 10 or 11 PM. There are few potential viewers after then, and the sky would be darker after those hours. Energy and money would be saved. (*National Optical Astronomy Observatories.*)

greatly in promoting quality outdoor lighting, and thus minimizing light pollution.

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

LIGHT-YEAR. A popular method of expressing large distances; specifically, the distance that light will travel in the course of one year. The velocity of light is established (International Astronomical Union (Hamburg 1964)) at 299,792.5 kilometers per second, or about 186,282 statute miles per second. There are approximately 31,560,000 seconds in a mean solar year. Accordingly, a light-year represents a distance of approximately 9.454×10^{12} kilometers; 5.875×10^{12} miles. See also **Astronomical Unit**; and **Parsec.**

LIGNIN. Approximately 25% of the content of most woods is lignin. Lignin concentration in wood substance is greatest in the middle lamella (the zone around each individual fiber cell), decreasing in concentration through the cross section of the fiber, reaching a concentration of about 12% at the inner layer of the fiber adjacent to the fiber cavity, or lumen. Lignin and hemicellulose cement the fiber cells together, providing rigidity to the fibrous wood structure. In the destructive distillation of wood, the methanol produced is derived from the lignin. In the manufacture of paper pulp, it is necessary to remove the lignin, usually accomplished by treatment of the wood fibers with such agents as sulfur dioxide, calcium bisulfite, and sodium sulfate/sodium sulfide solutions. Sodium hydroxide is sometimes used. An important by-product of the paper pulp industry is dimethyl sulfoxide, $(CH_3)_2SO$ which is produced from the lignin released during wood pulping by the Kraft process. Dimethyl sulfoxide has a number of industrial uses—as an intermediate in organic syntheses, as a solvent in spinning synthetic fibers, and in some pharmaceuticals.

The wall material of plant cells is one of their distinguishing characteristics. As a result, lignin, cellulose, and other wall constituents have been studied in many plant tissue cultures. Phenylpropanoids, for example, have been shown to be precursors of lignin formation in white pine, *Sequoia*, lilac, rose, carrot, and geranium tissue cultures. Moreover, the biosynthesis of lignin has been shown to be affected by kinetin, boron, and major elements, such as calcium.

Lignin is a major source of vanillin.

LIGNITE COAL. See **Coal.**

LIGNOCELLULOSE. See **Cellulose.**

LIGNUM VITAE (*Guaiacum*). The heartwood of a tree growing native in the West Indies. The tree also is found in other regions of moderate climate. It is a valuable, tough, resinous wood and very heavy, being the heaviest of all commercial woods; a cubic foot weighs 76 pounds (34.5 kilograms). Lignum vitae has been used in the making of bowling balls, pulley sheaves, and mallet heads.

The tree attains a height of about 40 feet (12 meters); the trunk may range from 2 to 4 feet (0.6 to 1.2 meters) in diameter and normally grows quite crooked. The branches are knotty.

The record *G. sanctum* (roughbark) tree growing in the United States is located in Biscayne National Park, Florida. As compiled by the American Forestry Association, this specimen has a circumference (at $4\frac{1}{2}$ feet; 1.4 meter above ground level) of 56 in. (142 cm), a height of 37 feet (11.3 meters), and a spread of 26 feet (7.9 meters).

The record *G. angustifolium* (Texas lignum vitae) tree growing in the United States is located at Alamo, Texas. This specimen has a circumference of 32 in. (81 cm), a height of 26 feet (7.9 meters), and a spread of 22 feet (6.7 meters).

LIKELIHOOD. If $P(x, \theta)$ denotes a probability function depending on one or more parameters collectively denoted by θ, the likelihood of a sample $x_1, x_2 \cdots x_n$ is defined as $L = P(x_1, \theta) \cdot P(x_2, \theta) \cdots P(x_n, \theta)$.

The method of maximum likelihood consists in estimating the parameters θ by choosing those values which maximize L (or $\log L$). Under general conditions, maximum likelihood estimates are consistent and efficient and tend to be normally distributed in large samples; further, that they are sufficient if sufficient statistics exist. The large sample variances or covariances of the maximum likelihood estimates are given by the elements of the inverse matrix to

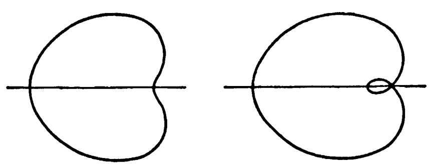
$$\left[nE\left(\frac{\partial^2 \log P}{\partial\theta_i\, \partial\theta_j}\right)\right]$$

where E denotes the expected value.

LILIACEAE. The Lily Family has representatives in all parts of the world, and more especially in the drier regions of the temperate zone. Several members of the family are important vegetables, notably asparagus and onions, while a great many more are cultivated for ornament. Among the latter are the true lilies (the genus *Lilium*), tulips, and hyacinths.

Most members of the Lily Family are herbaceous plants with a shallow fibrous root system. A few species of *Aloe* and *Dracaena* are shrubby or even small trees. Characteristic of the family are underground rhizomes or bulbs, storage organs which enable the plant to survive in regions where protacted dry seasons occur. As a rule, these plants have linear undivided leaves which do not show division into petiole and blade. The inflorescences of the family are very diverse. In some genera the flowers are solitary, in others they occur in racemes, while umbels occur in still others. The perianth of the flower has six separate members in two whorls of three, which are very much alike in size, shape, and color. The stamens have conspicuous anthers. The ovary is superior, 3-celled, and bears a single style with a 3-lobed stigma. The fruit is a capsule or a berry.

LIMACON. A higher plane curve, also known as Pascal's snail (named for Stefan Pascal, the father of Blaise Pascal, 1623–62, the famous French philosopher, mathematician, and physicist). Its equation in rectangular coordinates is $(x^2 + y^2 - 2ax)^2 = k^2(x^2 + y^2)$ and in polar coordinates, $r = 2a \cos \theta \pm k$. The curve is closed and symmetric to the X-axis. When $k < 2a$, there is an internal node at the origin and the limaçon is called hyperbolic. The loop disappears when $k > 2a$, so that a conjugate point exists and the limaçon is now elliptic. The case of $k = 2a$ is the cardioid, with a cusp at the origin.



Limaçon. In case of diagram at left, $k > 2a$; right, $k < 2a$.

The limaçon (its name comes from Latin, *limax*, snail) can be generated as follows. Let ODM be a circle with diameter $OD = 2a$ on the X-axis. A radius vector OM meets the circle at M and to it is added and subtracted a fixed length $MP = -MP' = k$. The locus of P and P' is the curve.

The various types of limaçons are special cases of the Cartesian oval. See **Curve (Higher Plane.)**

LIMB DARKENING. The darkening of the limb of the sun or stars is due to the line of sight passing through greater thicknesses of cooler gases at the edge. Limb darkening follows the simple relation

$$I = I_0(1 - u + u \cos \theta)$$

where I is the observed intensity at a point, making the angle θ between the observer, the center of the star, and the point in question; I_0 is the intensity at the center of the disk; and u is the coefficient of limb darkening. See also **Sun (The).**

LIME AND LIMESTONE. The term *lime* includes a variety of chemicals manufactured from limestone or derived from chemical processes which utilize calcium compounds. According to the composition of the parent limestone, lime may be designated as *high calcium lime* or *dolomitic lime*. Both *quicklimes*, CaO and CaO·MgO, and *hydrated* limes, $CaO \cdot H_2O$, $Ca(OH)_2 \cdot MgO$, and $CaO \cdot MgO \cdot 2H_2O$, are conventionally called lime. Precise terminology requires complex wording, e.g., *dolomitic quicklime* to denote CaO·MgO. The various lime oxides and hydroxides are among the lowest-cost and most widely used sources of alkali for the chemical and metallurgical industries. About 80% of the lime used in the United States is used by the chemical and related industries, mostly as quicklime. About 10% is dead-burned dolomite, and less than 10% goes into construction uses, mostly as hydrate. Very little lime is imported into or exported by the United States.

Limestone. This is a rock containing chiefly calcium carbonate and variable quantities of magnesium carbonate. Limestone is classified along the lines of lime as previously mentioned. *High-calcium limestone* contains 5% or less of $MgCO_3$ and occurs in two mineral forms, calcite and aragonite. See **Aragonite;** and **Calcite.** *Dolomitic limestone* usually contains over 35% $MgCO_3$, with the remainder $CaCO_3$. See **Dolomite.** *Magnesian limestone* is predominantly $CaCO_3$, but contains from 5 to 35% $MgCO_3$. All limestones evolve carbon dioxide and bubble in dilute hydrochloric acid. Dolomite reacts only with dilute HCl, while calcite will decompose in cold dilute HCl.

Limestones vary greatly in color and texture, the latter ranging from dense and hard limestone, e.g., marble or travertine, which can be sawed and polished, to soft, friable forms, e.g., chalk and marl. Chalk is a very fine-grained white limestone, while marl is an impure deposition product that contains clay and sand. Texture, hardness, and porosity appear to be functions of the degree of cementation and consolidation during the formation of these materials. Color variations arise from the presence of impurities. Some impurities, such as sulfur and phosphorus, make limestone unattractive for metallurgical uses.

A high percentage of all limestone is quarried; the balance is mined underground. Although limestone occurs widely, good chemical- and metallurgical-grade limestone is less plentiful. Along the seacoasts, oyster or clam shells are dredged as a source of $CaCO_3$. Limestone is normally processed through a series of crushing, screening, and grinding operations. Because of transportation costs, the proximity of limestone sources to points of use is highly desirable. The major uses of limestone are in construction (asphalt filler, road stone, riprap, and bituminous aggregate); in Portland cement; in agriculture; and in metallurgy.

Precipitated $CaCO_3$ is produced in a number of chemical processes. Sometimes it is economical to dry and calcine the by-product to regenerate CaO or $Ca(OH)_2$. Some precipitated $CaCO_3$ is made to specific particle size and shape, whiteness, and purity for use as functional filler for paper coatings, paint, and polymers. These products command a premium price as compared with pulverized limestone fillers.

Manufacture of Lime. The basic processes are calcination and hydration. Commencing with high-calcium limestone, the reactions are:

$$CaCO_3 + \text{heat} \rightleftharpoons CaO + CO_2 \qquad (1)$$

$$CaO + H_2O \rightleftharpoons Ca(OH)_2 + \text{Heat} \qquad (2)$$

If dolomitic limestone is used, the reactions are:

$$CaCO_3 \cdot MgCO_3 + \text{heat} \rightleftharpoons CaO \cdot MgO + 2CO_2 \qquad (3)$$

$$CaO \cdot MgO + H_2O_{(liq)} \rightleftharpoons Ca(OH)_2 \cdot MgO + \text{heat} \qquad (4a)$$

or

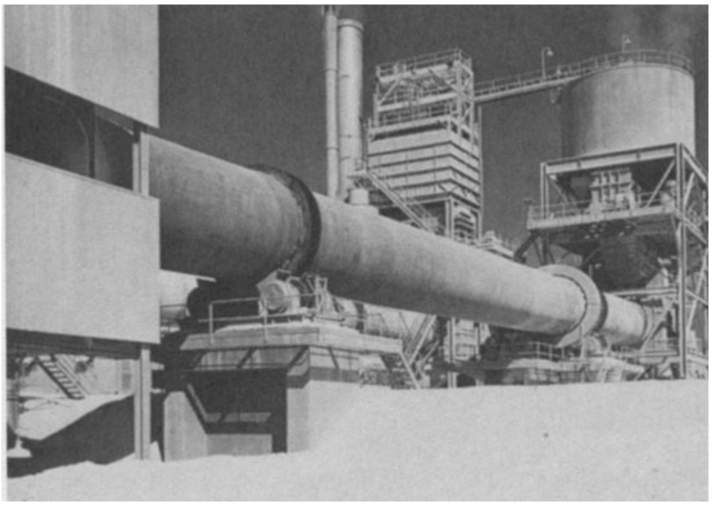
$$CaO \cdot MgO + 2H_2O_{(gas)} + \text{pres} \rightleftharpoons Ca(OH)_2 \cdot Mg(OH)_2 + \text{heat} \quad (4b)$$

High-calcium limestone dissociates at 900°C (1650°F) in 100% carbon dioxide atmosphere at 1 atm pressure. Under similar conditions, dolomitic limestone dissociates over 727–900°C (1340–1650°F). The heat of reaction required to convert $CaCO_3$ to CaO is about 2.8 million Btu per ton (0.64 million kg-Cal/metric ton) of CaO. In practice, heat input may vary from 4 to 10 million Btu/ton (0.9 to 2.3 million kg-Cal/metric ton) of lime. Calcination of limestone particles proceeds by a receding-surface mechanism. To attain reasonable rates of heat transfer into the center of the rock or pebble-sized stone, operating temperatures in lime kilns are 980–1260°C (1800–2300°F). Reaction rate is increased and opportunity for recarbonation of the oxide is decreased by rapid removal of CO_2 from the kiln.

Except for very old mixed-feed vertical kilns, lime kilns operate with countercurrent flow of raw material and heat. Modern lime kilns utilize coolers to preheat air by recuperating heat from the hot quicklime. Lime kilns may be fired directly with coal, oil, or gas.

Two major types of lime calciners are the rotary (see accompanying figure) and the vertical kiln. In North America, rotary kilns are widely used for lime calcination, whereas in Europe vertical kilns are most popular. Rotary kilns typically have higher output [up to 600 tons (545 metric tons)/day] and lower labor cost. Vertical kilns can be designed for higher fuel efficiency and lower capital investment. They handle down to about $\frac{3}{4}$-in. (19-cm) stone, but normally the rock feed is at least 3 × 6 in. (7.5 × 15 cm). Rotary kilns can handle down to $\frac{1}{4}$-in. (0.6 cm) stone.

The long-established use of lime is as a structural material in masonry mortars, wall plasters, sand-lime brick, and for soil stabilization. Double-hydrated dolomitic lime or specially processed high-calcium lime mixed with gypsum plaster is troweled on interior walls or ceilings to provide a hard, white, finished surface. It is mixed with cement and sand to make exterior plaster or stucco. Mason's mortar used to lay up



Large rotary lime kiln designed to operate 24 hours per day year around. Rugged construction is required for handling abrasive limestone rock at very high temperature.

bricks or blocks usually contains lime. Lime provides plasticity, water retention, and easy troweling. Sand-lime bricks are more popular in Europe than in the United States. About 10% hydrated lime is mixed with graded sand and water, pressed into shape, and put into autoclaves for 4–8 hours at 150–205°C (300–400°F). The reaction product, calcium silicate, results in a strong, white brick. Dead-burned dolomite, formed by calcining dolomite at about 1650°C (300°F) to convert MgO to periclase, is used as a *refractory*. Lime is used in the sulfate process for making paper. In the seawater process for producing magnesium metal, lime reacts with $MgCl_2$ to precipitate $Mg(OH)_2$. Calcium metal is made from lime by reducing CaO with coke. In water treatment, hydrated lime can be added to remove temporary hardness, or in the lime-soda process to remove permanent hardness. Dolomitic lime removes silica from boiler feedwater due to silica absorption by $Mg(OH)_2$. For acid neutralization of industrial wastes, lime is widely used.

LIME (Citrus). See **Citrus Trees.**

LIMIT. A finite number s approached by a sequence $\{s_n\}$ if, for every positive number $\epsilon > 0$, there exists a number N for all $n > N$, where $|s_n - s| < \epsilon$. The sequence is then said to converge, or symbolically,

$$s_n \to s; \quad n \to \infty$$

or

$$\lim_{n\to\infty} s_n = s$$

If the limit does not exist, the sequence diverges, but see also **Series.**

The limit of a variable or a function can be defined in a similar way. If the limit is zero, the function is called an infinitesimal. If the function increases without limit, it remains greater than any assigned number however large and it is said to be come infinite or approach infinity. A variable or function which decreases without limit approaches minus infinity.

The following definition is sometimes needed. If ϵ is a positive number, no matter how small, δ is defined by $0 < (x - a) < \delta$, and $|f(x) - A| < \epsilon$, then A, called the right-hand limit at $x = a$, is symbolized by

$$\lim_{x\to a+} f(x)$$

A left-hand limit, $\lim_{x\to a+} f(x)$ can be defined in a similar way.

See also **Sequence.**

LIMIT SWITCH. An enclosed electromechnical device which makes or breaks electrical circuits when actuated by an external force, such as by a machine member (lever, cam, or dog), or other object when a preselected position of travel or limit is reached. Circuits switched may be for safety stop, position indicating, or be part of a total sequence of automatic or semi-automatic operations. The limit switch is important in many automated systems, such as machine tools, conveyors, and other materials handling equipment. Unlike photoelectric and proximity switches, the limit switch requires physical contact for actuation. These switches are designed to operate reliably for tens of millions of actuations in industrial environments. The switches are obtainable in numerous current and voltage ratings, ranging from the switching needs of solid-state devices to motor-controlling relays and large solenoids. Voltage requirements may be as high as 600 volts.

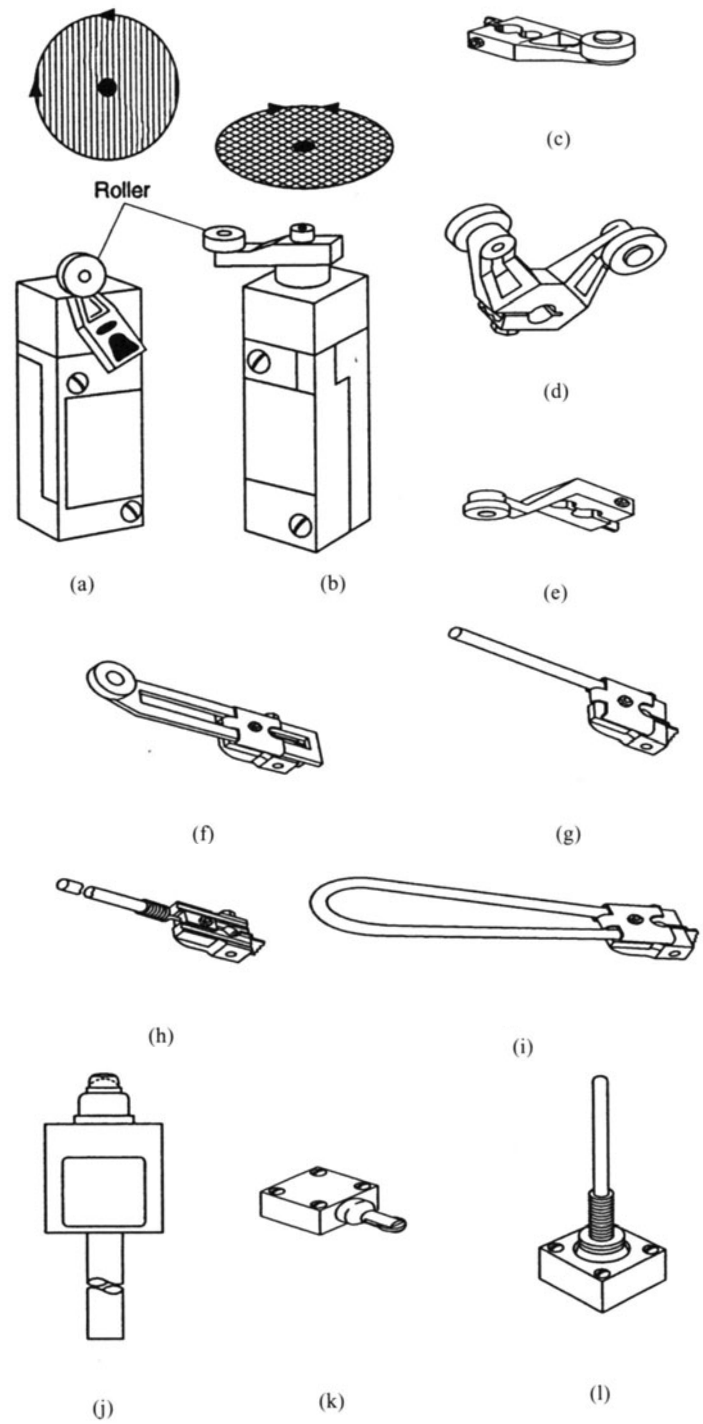


Fig. 1. Various configurations of electromechanical limit switches. (*a*) Side-mounted roller. (*b*) Top-mounted roller. (*c*) Standard roller lever. (*d*) Yoke roller lever. (*e*) Offset roller lever. (*f*) Adjustable-radius roller lever. (*g*) Rod lever. (*h*) Spring-rod lever. (*i*) Flexible loop lever. (*j*) Top plunger. (*k*) Side roller plunger. (*l*) Wobbler switch. (*MICRO SWITCH.*)

Limit switches are actuated in several ways, but the two principal methods are (1) the roller lever switch, and (2) the plunger switch. Low-force actuators, termed *cat-whisker* or *wobblestick*, also are available. See Fig. 1.

See also **Proximity and Object Detectors.**

LIMNOLOGY. The science of lakes, a synthesis of many disciplines, drawing its specialists from various scientific fields. The field includes the study of inland waters, although it is principally concerned with the physicochemical nature of lakes, their flora and fauna. Stream study has lagged behind lake investigation, although the ecological approach to rivers falls within the realm of limnology.

Lake basins owe their origins to diverse causes, many of which are geologic accidents, or catastrophic. Geologically, lakes are temporary phenomena in geomorphic evolution. Tectonic events have created some of the oldest and deepest lakes of the world: the African rift lakes and Lake Baikal of Siberia, with an ancient, largely endemic fauna. Vulcanism, glacial activities, solution of calcareous substrates, aeolian forces, and even meteoritic impact have created lake basins. In 1957, Hutchinson summarized 76 major categories and 8 subdivisions of these events which have resulted in lake genesis.

No matter what its origin, a lake is doomed to eventual extinction because of its concave nature and accumulation of autochthonous and allochthonous materials,[1] which gradually obliterate the depression. Thus, a lake passes from a youthful stage to maturity, senility, and extinction. The rate of succession depends upon various factors. For example, introduction of domestic sewage enriches the lake and accelerates the aging process. The youthful lake may be described as *oligotrophic*, the mature lake as *eutrophic*. Many intermediate stages between extreme oligotrophy and extreme eutrophy occur, and the term *mesotrophy* can be applied to them. Senility is characterized by much shallow water and the conspicuous enroachment of large aquatic plants upon the open water. Extinction often involves a marshy meadow which is later colonized by plants typical to terrestrial situations. If drainage is poor and the lake is protected from wind, a floating bog mat may close over and eventually obliterate the open water. Bog lakes are acid, or at least circumneutral, and are typified by characteristic marginal vegetation contributing to the floating mat. When calcium content is low in bog-lake waters, decay of organic matter is reduced greatly. Plant fragments from the bog mat accumulate in flocculent layers, and the water may become tea-colored from humic matter. Flocculated humic colloids contribute to bog sediments to form a characteristic deposit termed *dy* by Scandinavian researchers. Under such conditions, nutrients are not recycled by decay, and the lake approaches extinction as a *dystrophic lake*.

In recent years, pollution of lakes has been of major concern. The effect of acid rain on certain lakes, usually at relatively high elevation (alpine lakes) is described in the entry on **Water Pollution.** Accelerated eutrophication, resulting primarily from phosphorus additions due to anthropogenic activities, is generally regarded as one of the major causes of the deterioration of the Great Lakes water quality. A mathematical model of the Great Lakes total phosphorus budgets indicates that a milligram per liter effluent restriction for point sources would result in significant improvement in the trophic status of most of the system. However, because large areas of their drainage basins are devoted to agriculture or are urbanized, western Lake Erie, lower Green Bay, and Saginaw Bay may require non-point source controls to effect significant improvements in their trophic status.

Volcanic lakes are discussed under **Volcano.**

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[1]Definied in separate entries in this encyclopedia.

LIMONITE. The mineral limonite, hydrated oxide of iron, corresponds to the formula $Fe_2O_3 \cdot nH_2O$, but is often very impure due to the admixture of sand and clay. It is not found crystallized but grades from loose porous material to compact masses. Its hardness is variable but pure material is 5–5.5; specific gravity 3.6–4; usual luster, dull to earthy but may be silky to submetallic; color, various shades of yellowish-brown, sometimes nearly black; streak, yellowish-brown; opaque. Limonite is a secondary mineral from the alteration of various other iron-bearing ores or minerals; it is of widespread occurrence and used both as an ore of iron and as a pigment. Limonite has been formed in marshy and boggy areas and is frequently called bog iron ore. Limonite is an important ore of iron in Lorraine, Luxemburg, Bavaria and Sweden. It is found in Saxony, Austria and England. In the United States, limonite is found particularly in Connecticut, Massachusetts, Pennsylvania, New York, Virginia, Tennessee, Georgia and Alabama, but these deposits are of little economic importance at the present time.

LIMPET (*Mollusca*, *Gasteropoda*). Marine and fresh-water animals related to the snails, with a low conical shell, not spirally twisted. In the common limpets the shell is solid and in the keyhole limpets it is either notched in front or perforated between that point and the apex. Mollusks of the family *Capulidae*, more closely related to some of the species with coiled shells than to the true limpets, also have shells which are not spiral and are called limpets. One form, *Crucibulum*, is called the cup and saucer limpet and another, *Crepidula*, the boat limpet or slipper shell.

LINEAR. An adjective often used in mathematics to describe certain properties. Given the quantities $x_1, x_2, x_3, \ldots, x_n$, a linear combination of them is $a_1x_1 + a_2x_2 + \cdots + a_nx_n = 0$. The quantities x_i are linearly dependent provided all a_i are not zero; otherwise they are linearly independent. The test for such dependence may be made by means of the Gram determinant or the Wronskian.

A linear function is a polynomial of the first degree in its variables and it usually means a polynomial in one variable. Thus, with the linear function, $y = mx + b$, a plot of it would be a straight line of slope m and intercept b on the Y-axis. The general case of a linear algebraic equation would be $a_1x_1 + a_2x_2 + \cdots + a_nx_n = a_0$, where the x_i are variables and the a_i are constants.

A set of simultaneous linear equations is

$$\sum_{j=1}^{n} a_{ij}x_j = b_i$$

$i = 1, 2, \ldots, n$ and a_{ij}, b_i are constants.

A linear differential equation is $A_0(x)y + A_1(x)y' + A_2(x)y'' + \cdots + A_n(x)y^{(n)} = f(x)$ where the $A_i(x)$ are functions of the independent variable only and $y', y'', \ldots$ are the first, second, etc., derivatives. These equations are also inhomogeneous. If the right-hand side is zero in any case, the equation is still linear but homogeneous. Similarly, if all b_i are zero in the case of simultaneous algebraic equations, they are also homogeneous.

See also **Transformation (Mathematics).**

LINEAR ACCELERATOR. See **Particles (Subatomic).**

LINEAR ENERGY TRANSFER. The linear energy transfer of charged particles passing through a medium was defined by the ICRU in 1962 as dE_L/dl, where dE_L is the average energy locally imparted to the medium by a charged particle of specified energy in traversing a distance *dl*. The term locally imparted may refer either to a maximum distance from the track or to a maximum value of discrete energy loss by the particle beyond which losses are no longer considered as local. In either case the limits chosen should be specified. The concept of linear energy transfer is different from that of stopping power. The former refers to energy imparted within a limited volume, the latter to loss of energy regardless of where this energy is absorbed.

LINEAR GRAPH. See **Graph (Mathematics).**

LINEAR HYPOTHESIS. In statistics, generally, any hypothesis which is linear in the parameters entering into it. More specifically, the term has been used to refer to hypothesis which are linear in the means of a set of normal distributions, many of the simpler statistical hypotheses being capable of being thrown into such a form.

LINEAR INEQUALITIES. A system of relations among variables x_i, possibly including linear equations among them, but also including at least one inequality of the form

$$\sum a_i x_i \geq b$$

(In practice a strict inequality is seldom required.) Such a system may be incompatible (e.g., $x_1 \geq 0, x_2 \geq 0, -x_1 - x_2 - 1 \geq 0$), may define a unique point (e.g., $x_1 \geq 0, x_2 \geq 0, -x_1 - x_2 \geq 0$), or else will define a region in space, not necessarily bounded (e.g., $x_1 \geq 0, x_2 \geq 0, -x_1 - x_2 + 1 \geq 0$ define a bounded region, $x_1 \geq 0, x_2 \geq 0, x_1 + x_2 - 1 \geq 0$ an unbounded region).

The inequality written above can be replaced by the equivalent pair

$$x_0 + \sum a_i x_i = b_i, \qquad x_0 \geq 0$$

and, in general, it is possible to replace a system of inequalities by a system in the special form

$$\mathbf{Ax} = \mathbf{b}, \qquad \mathbf{x} \geq 0$$

where **A** is a rectangular matrix, **x** and **b** are vectors.

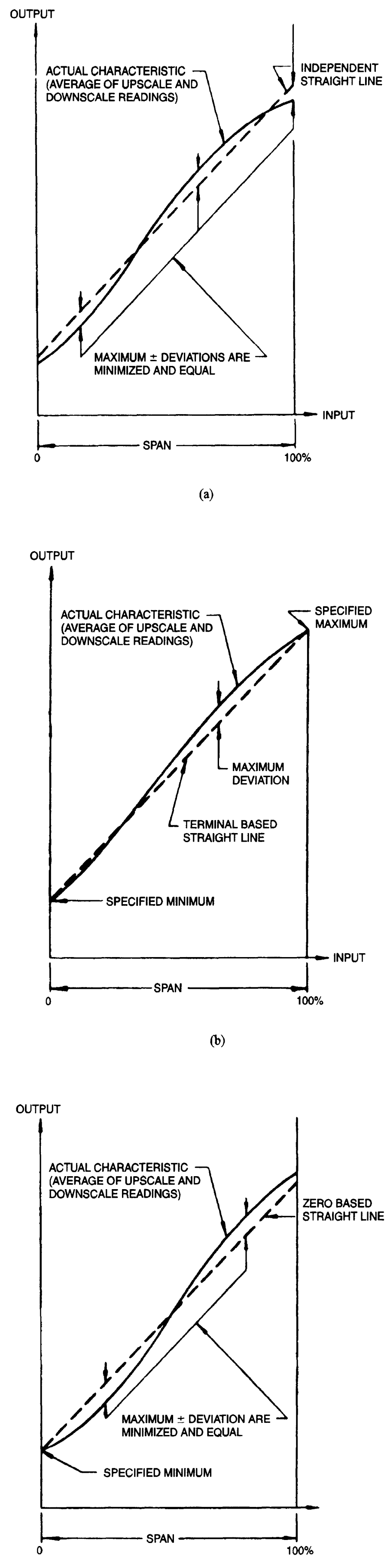
LINEARITY. With reference to industrial and scientific instruments, the Scientific Apparatus Makers Association defines linearity as the closeness to which a curve approximates a straight line. Linearity is usually measured as a *nonlinearity* and expressed as *linearity*; e.g., a maximum deviation between an average curve and a straight line. The average curve is determined after making two or more full range traverses in each direction. The value of linearity is referred to the output unless otherwise stated. As a performance specification, linearity should be expressed as *independent linearity*, *terminal-based linearity*, or *zero-based linearity*. When expressed simply as linearity, it is assumed to be independent linearity. See also **Conformity.**

Independent Linearity. The maximum deviation of the actual characteristic (average of upscale and downscale readings) from a straight line so positioned as to minimize the maximum deviation. See (a) of accompanying diagram.

Terminal-Based Linearity. The maximum deviation of the actual characteristic (average of upscale and downscale readings) from a straight line coinciding with the actual characteristics at upper and lower range-values. See (b) of accompanying diagram.

Zero-Based Linearity. The maximum deviation of the actual characteristic (average of upscale and downscale readings) from a straight line so positioned as to coincide with the actual characteristic at the lower range-value and to minimize the maximum deviation. See (c) of accompanying diagram.

LINEARITY CONTROL. 1. In cathode-ray tube equipment, an adjustment that tends to correct any distortion in the sawtooth current or voltage waves used for deflection. 2. In television, a control which varies the distribution of scanning speed throughout the trace interval.



Fundamental relationships pertaining to linearity: (a) independent linearity; (b) terminal-based linearity; (c) zero-based linearity.

LINEARLY INDEPENDENT VECTORS. The vectors $A_1, A_2, \cdots, A_m$ are linearly independent if the equation $c_1A_1 + c_2A_2 + c_2A_2 + \cdots + c_mA_m = 0$ implies that $c_1 = c_2 = \cdots = c_m = 0$. Any n-dimensional space contains n linearly independent vectors. They are a base of the vector space. Any other vector can be written as a linear combination of these base vectors. It is convenient to choose the base vectors mutually perpendicular or orthogonal and of unit length. In ordinary three-dimensional space, these vectors are denoted by i, j, k.

LINEARITY THEOREM. See **Laplace Transform.**

LINEAR MAGNIFICATION. For each optical surface in a system, the linear size of the image, I_i, is to the linear size of the object, O_i, as is the distance of the image, Q_i, to the distance of the object, P_i. Then

$$I_i/O_i = Q_i/P_i = m_i$$

The total linear magnification of the system is then the sum of the products of the magnifications of the parts.

LINEAR PROGRAMMING. A technique of mathematics and operations research for solving certain kinds of problems involving many variables where a best value or set of best values is to be found. This technique is not to be confused with computer programming, although problems using the technique may be programmed on a computer. Linear programming is most likely to be feasible when the quantity to be optimized, sometimes called the objective function, can be stated as a mathematical expression in terms of the various activities within the system, and when this expression is simply proportional to the measure of the activities; i.e., is linear, and when all the restrictions are also linear.

In 1947, mathematician George B. Dantzig devised a linear programming algorithm (the *simplex algorithm*). Long before the advent of modern computers, the method enabled machines to handle complex problems with hundreds of constraints. Over the years, the method has been improved—to the point where it can process problems with 15,000 to 20,000 constraints. Beyond this point, the simplex algorithm may become prohibitively slow and cumbersome. Many of today's problems, especially those of the telecommunications industry, are larger and may reach many tens of thousands or more constraints. A large problem may require 12 to 24 hours or longer of computer running time.

In 1984, Narendra Karmarkar (AT&T Bell Laboratories) demonstrated a new algorithm (called the *Karmarkar* or *AT&T algorithm*) that greatly reduces computation times for many practical problems and it is still being explored to determine its ultimate capabilities and limitations. In an early test of the algorithm, it was applied to the solution of a communication network planning problem involving approximately 42,500 variables and 15,000 constraints. For sake of comparison, the same problem was run with a widely used conventional simplex linear programming package. The results showed the Karmarkar method to be faster than the simplex method by more than an order of magnitude, i.e., every 10 hours of time required by the older method was cut to 1 hour for the new method.

Essence of the New Algorithm. A linear programming problem, such as an overseas communication facilities planning project, can be assembled as a matrix equation that represents variables, constraints, and an objective function (cost) to be optimized. The problem's possible solutions can be envisioned as a geometric shape (a polyhedron, referred to by mathematicians as a *polytope*). The boundary of the polytope is formed from multisided, irregular, flat planes called polygons. Each polygon corresponds to an equation describing a constraint. The optimal solution always lies on one of the vertices (corner points) of the polytope. In the simplex algorithm, the optimal solution is arrived at by starting at one of the vertices of the polytope and "hopping" to the most appropriate adjacent vertex. This vertex must be selected from many available, and must lie in the direction of optimal cost. The process is repeated many times as the computer searches from vertex to vertex for the optimal solution, using the simplex algorithm as its guide ("steering" the computer). The zig-zag path of iterations crosses the surface of the polytope until it ends at the final vertex. Large problems are difficult to solve using the simplex method because a lot of vertices need to be checked, and because movement in the direction of optimal cost is limited to only one vertex at a time. Many thousands of such movements (iterations) are required to solve the problem.

With the Karmarkar algorithm, the lengthy process is cut short, so to speak, by moving through the polytope's interior, finding a much more direct route to the solution. For this kind of problem, the solution does not consist of a single, short answer. The final result defines the optimal value for each of the thousands of variables. It also ensures that the many thousands of constraints have been satisfied as well. This massive solution is typically read onto a disk, from which the facility planner can extract all or part of the information in printed form.

Details and examples of the Karmarkar algorithm can be found in the "Record," 4–13 (March 1986), published by AT&T Bell Laboratories, 150 John F. Kennedy Parkway, Short Hills, New Jersey 07078.

LINEAR SPACE (Vector Space). A generalization of certain algebraic aspects of addition in Euclidean space.

A linear space is a set V with a binary relation called addition making it a group and a multiplication of the elements of V by scalars. The elements of V are called *vectors*. The scalars are usually the real or complex numbers.

The subset $x_1, x_2, \ldots, x_n$ of vectors in V is said to be *linearly* independent if the only scalars $\lambda_1, \lambda_2, \ldots, \lambda_n$ for which $\lambda_1x_1 + \lambda_2x_2 + \ldots + \lambda_nx_n = 0$ are $\lambda_1 = \lambda_2 = \cdots = \lambda_n = 0$. A subset $x_1, x_2, \ldots x_n$ is said to be a *basis* for V if every vector x in V can be expressed $x = \lambda_1x_1 + \lambda_2x_2 + \cdots + \lambda_nx_n$ for scalars $\lambda_1, \lambda_2, \ldots, \lambda_n$. The number of vectors in a linearly independent basis for V is said to be the (linear) dimension of V. The linear dimension of three-dimensional space is three.

If $e_1, e_2, \ldots, e_n$ is a linearly independent basis for V, then each vector in V corresponds to the n-tuple $(\lambda_1, \lambda_2, \ldots, \lambda_n)$ of scalars for which $x = \lambda_1e_1 + \lambda_2e_2 + \cdots + \lambda_ne_n$. Conversely, the set of n-tuples of scalars form a linear space.

A mapping L from V to the scalars is said to be a *linear functional* on V if it satisfies: $L(\lambda_1x_1 + \lambda_2x_2) = \lambda_1 L(x_1) + \lambda_2L(x_2)$. The set of linear functionals on V can be made into a linear space in a natural way and this linear space is called the *adjoint* or *conjugate space* of V.

Since many systems describing physical situations (or a first-order approximation) are linear, that is, satisfy the principle of superposition, the space of possible states of the system is a linear space. Also linear spaces are important in discussing linear partial differential equations such as the wave equation or the heat equation.

See also **Euclidean Space.**

LINEAR SYSTEMS. Systems such that the interrelated quantities comprising the system are related by linear differential or differentio-integral equations. Such equations and therefore such systems obey the principle of superposition, namely, the combined effect of a number of causes acting together is the sum of the effects of the several causes acting separately.

LINEAR TOPOLOGICAL SPACE. A generalization of the algebraic and topological aspects of Euclidean space to infinite dimensional linear spaces.

A linear topological space is a linear space X with scalars either the real or complex numbers along with a topology $\mathcal{T}$ so that addition of vectors and scalar multiplication are continuous. Elementary examples of linear topological spaces are the Euclidean spaces with the usual topology. More representative examples are Hilbert space and the space of infinitely differentiable functions on some Euclidean space with the topology of uniform convergence of each derivative on closed bounded subsets.

A function $x \to \|x\|$ from a linear space V to the nonnegative reals is said to be a *norm* if (1) $\|x\| = 0$ implies $x = 0$, (2) $\|\lambda x\| = |\lambda|\, \|x\|$, and (3) $\|x + y\| \le \|x\| + \|y\|$ and V is said to be a *normed linear space*. The "metric topology" defined on V by the metric $\rho(x, y) = \|x - y\|$ makes V into a linear topological space. V is said to be a *Banach space* if it is complete relative to the metric $\rho(x, y)$.

The *conjugate space* V^* of a linear topological space V is the set of all continuous linear functionals on V.

A continuous linear operator ϕ is a mapping between linear topological spaces V and W that is continuous and satisfies $\phi(\lambda_1 x_1 + \lambda_2 x_2) = \lambda_1\phi(x_1) + \lambda_2\phi(x_2)$.

The study of linear topological spaces provides the theoretical underpinning for the modern theory of distributions or generalized functions. Although used by engineers and physicists during most of this century, generalized functions were first defined in a mathematically rigorous way by L. Schwartz in the late forties to be the linear functionals on the space of infinitely differentiable functions.

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LINE (Mathematics). The path described by a moving point. If, as is generally meant, the line is straight its equation in a plane and in rectangular coordinates is $Ax + By + C = 0$, which is a degenerate conic section. Other forms of its equation are: (a) $y = mx + b$, where m is the slope and b the y-intercept; (b) $x/a + y/b = 1$, where a, b are the $x-$, y-intercepts, respectively; (c) $y - y_1 = m(x - x_1)$, where (x_1, y_1) is a point on the line; (d) $(y - y_1)/(y_2 - y_1) = (x - x_1)/(x_2 - x_1)$, where (x_2, y_2) is another point on the line; (e) $x \cos\theta + y \sin\theta = p$, where the perpendicular from the origin to the line has length p and makes an angle θ with the X-axis; (f) $rr_1 \sin(\theta - \theta_1) + r_1 r_2 \sin(\theta_1 - \theta_2) + rr_2 \sin(\theta_2 - \theta) = 0$, where (r, θ) are polar coordinates of any point on a straight line passing through two points (r_1, θ_1) and (r_2, θ_2); (g) $x = x_0 + at$, $y = y_0 + bt$, where t is a parameter for the line of slope b/a which passes through the point (x_0, y_0).

If the straight line is located in a three-dimensional rectangular coordinate system, its equation may be taken as

$$(x - x_1)/\lambda = (y - y_1)/\mu = (z - z_1)/v$$

or

$$(x - x_2)/(x_2 - x_1) = (y - y_1)/(y_2 - y_1) = (z - z_1)/(z_2 - z_1)$$

where (x_1, y_1, z_1) and (x_2, y_2, z_2) are two points on the line and λ, μ, v are its direction cosines or numbers proportional to them. A straight line in space is also determined by the equations of two planes, intersecting to form the given line. Thus, such a line can be defined by the simultaneous equations for two planes, $Ax + By + Cz + D = 0$; $A'x + B'y + C'z + D' = 0$.

See also **Coordinate System;** and **Direction Cosine.**

LINE OF APSIDES. A line that contains the major axis of an ellipse is known as the line of apsides of the ellipse. In astronomy, the term is used to indicate the line joining perihelion and aphelion points in an orbit and extending to infinity to cut the celestial sphere. See also **Orbit.**

LINE OF NODES. The astronomical term applied to a line of intersection of any two fundamental planes. The line of nodes for the moon is the line of intersection of the plane containing the moon's orbit with the plane of the ecliptic. The line of nodes for any member of the solar system, other than satellites, is the line of intersection of the plane of the orbit of the object with the plane of the ecliptic. The line of nodes for the earth is the line of intersection of the plane of the earth's equator with the plane of the ecliptic. For binary stars, the line of nodes is the line caused by the intersection of the plane of the orbit with the plane perpendicular to the line of sight and containing the center of gravity of the system.

LINE OF POSITION. In navigation any line on the surface of the earth upon which a ship is known to be located. If two or more lines of position are known, the ships must be at their point of intersection. A position determined by the intersection of lines of position is known as a fix.

Lines of position are usually circles, either great or small. In most cases, the distance of the center is so great in comparison with the length of the plotted line that the curvature is not apparent in the plot. In some cases, however, where both points of intersection of two circles of position appear on the plot, a dead-reckoning position will indicate which one is the desired fix. Lines of position are obtained by several methods. See also **Course; Navigation;** and **Pilotage (or Piloting).**

LINE WIDTH. A measure of the spread in wavelength (or energy) of radiation that is normally characterized by a single wavelength (or energy) value. In practice, the width is usually measured at one-half the maximum intensity of the line. The three phenomena that contribute to line broadening are doppler broadening, pressure broadening, and the intrinsic level width. Due to the finite resolution of measuring apparatus, a broadening due to the characteristics of the instrument also must be considered.

LINOLEIC ACID. Also called linolic acid, formula $CH_3(CH_2)_4HC$:$CHCH_2CH$:$CH\ (CH_2)_7COOH$. This is a polyunsaturated fatty acid (two double bonds) existing in both conjugated and unconjugated forms. It is a plant glyceride essential to the human diet. It is found in linseed oil, safflower oil, and tall oil. See also **Vegetable Oils (Edible).** At room temperature linoleic acid is a colorless to straw-colored liquid. Specific gravity 0.905 (15/4°C); mp −5°C; bp 228°C (at 14 millimeters pressure). Insoluble in water; soluble in most organic solvents. Combustible. Sources are the oils previously mentioned. Linoleic acid is used in soaps; special driers for protective coatings; emulsifying agents; pharmaceuticals; livestock feeds; and margarine.

LINOLENIC ACID. Also called 9,12,15-octadecatrienoic acid, formula CH_3CH_2CH:$CHCH_2CH$:$CHCH_2CH$:$CH(CH_2)_7COOH$. This is a polyunsaturated fatty acid (three double bonds). It occurs as the glyceride in many seed fats. It is an essential fatty acid in the diet. See also **Vegetable Oils (Edible).** At room temperature, linolenic acid is a colorless liquid; soluble in most organic solvents; insoluble in water. Specific gravity 0.916 (20/4°C); mp −11°C; bp 230°C. Combustible. Linolenic acid finds use in various pharmaceuticals and drying oils.

LION. See **Cats.**

LIOUVILLE EQUATION. In the statistical mechanics of an ensemble of systems, each containing N particles of mass m, it is useful to introduce a density or probability function $P_N(q_N, p_N)$, representing the probability that a system of the ensemble will have its point in phase space fall within the volume bounded by $q_1, q_2, \ldots, q_N, p_1, p_2, \ldots, p_N$, and $q_1 + \delta q_1, q_2 + \delta q_2, \ldots, q_N + \delta q_N, p_1 + \delta p_1, p_2 + \delta p_2, \ldots, p_N + \delta p_N$. If no new systems are created or destroyed, P_N will satisfy the Liouville equation:

$$\frac{\partial P_N}{\partial t} + \sum_{j=1}^{N}\left(\nabla_j \frac{P_N p_j}{N} + \delta_j P_N \dot{p}_j\right) = 0$$

where

$$\nabla_j = \partial/\partial q_j \qquad \text{and} \qquad \delta_j = \partial/\partial p_j$$

The p-terms are values in generalized coodinates for the positions of the systems, the q-terms are their momenta, the δ-terms in p and q denote small changes, $\partial P_N/\partial t$ is the partial derivative, and ∇_j and δ_j are vector differential operators. The dot over p_j denotes its first derivative with respect to time.

LIOUVILLE-NEUMANN SERIES. An infinite series

$$\phi(x) = \sum_{n=0}^{\infty} \lambda^n \phi_n(x)$$

which is a unique, continuous solution of a Fredholm integral equation of the second kind. If the *n*th iterated kernel is defined as

$$K_n(x, z) = \int\int \cdots \int K(x, y_1)K(y_1, y_2) \cdots K(y_{n-1}, z)\, dy_1\, dy_2 \cdots dy_{n-1}$$

then

$$\phi_n(x) = \int K_n(x, z) f(z)\, dz$$

The resolvent or solving kernel is given by

$$K(x, z; \lambda) = \sum_{n=0}^{\infty} \lambda^n K_{n+1}(x, z)$$

hence the solution of the integral equation becomes

$$\phi(x) = f(x) + \lambda \int K(x, z; \lambda) f(z)\, dz$$

Similar methods may be used to solve the Volterra equations.

LIPIDOSES. Disturbances of lipid metabolism in which abnormal deposits of lipids are found in various groups of cells in the body. Primary lipidoses, i.e., those due to an inborn error of metabolism include: (1) *Gaucher's disease*, in which the product accumulated is ceramide glucoside, caused by a beta-glucosidase enzyme defect. In the infantile form, there is mental retardation. In the adult form, there is hepatosplenomegaly and bone changes. (2) *Niemann-Pick disease*, in which the product accumulated is sphingomyelin, caused by a sphingomyelinase enzyme defect. In this disease, there is mental retardation and hepatosplenomegaly. (3) *Tay-Sachs disease*, in which the product accumulated is ganglioside G_{M2}, caused by a hexosaminidase A enzyme defect. In this disease, there is mental retardation and sometimes blindness. (4) *Generalized gangliosidosis*, in which the product accumulated is ganglioside G_{M1}, caused by a beta-galatctosidase enzyme defect. In this disease, there is mental retardation and hepatosplenomegaly (enlargement of liver and spleen). In secondary lipidoses there may be abnormal overgrowth of reticulum cells, in which lipoids are present; these include the Letterer-Siwe syndrome (destructive lesions chiefly in bone, fatal in infancy), Hand-Schäuller-Christian syndrome (bone lesions with fibrosis of the lungs, diabetes and stunted growth) and the eosinophilic granuloma of bone, a solitary, benign and often self-limiting tumor of children and young adults.

Additional Reading

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Figueroa, M. L., et al.: "A Less Costly Regimen of Alglucerase to Treat Gaucher's Disease," *N. Eng. J. Med.*, 1632 (December 5, 1992).
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LIPIDS. A heterogenous group of substances which occur ubiquitously in biological materials. They may be categorized as a group by their extractability in nonpolar organic solvents, such as chloroform, carbon tetrachloride, benzene, ether, carbon disulfide, and petroleum ether. Structural types within the group range from simple, straight-chain hydrocarbon molecules to complex ring structures with varying side chains. A useful classification of the lipids is: (1) fatty acids; (2) neutral fats; (3) phosphatides; (4) glycolipids; (5) aliphatic alcohols and waxes; (6) terpenes; and (7) steroids. See **Steroids.**

Many lipids, especially the phospholipids, have a strong tendency to form complexes with each other and with various substances. Complex formation is due to the electrostatic attraction of polar groups and to the mutual solubility of the long hydrocarbon chains. Thus, the lipoproteins and proteolipids are complexes of proteins and a variety of lipids, such as cholesterol, phospholipids, glycerides, and glycolipids. The lipids are linked to the proteins by several types of forces. Electrostatic forces, van der Waals forces, hydrogen bonding, and hydrophobic bonding hold these complexes together. Because of their attraction for water, the polar groups of the protein and phospholipid arrange themselves on the outside of the complex, while the hydrocarbon groups of the lipids are folded into the center. Thus, there is presented to the aqueous phase those groups which have an affinity for water. This arrangement accounts for the solubility of the complexes in water. The phospholipids, owing to their polar groups, act as water solubilizers for the nonpolar lipids. The arrangement may be different in the proteolipids of the brain and nerves, since they are not soluble in water. In these complexes, the lipids may completely envelop the protein.

Knowledge of lipid metabolism has increased at an accelerated rate during the past few decades. The detailed biochemical reactions whereby the fatty acids are synthesized and oxidized; how phospholipids, glycolipids, and cholesterol are synthesized; and how lipids are absorbed and transported have been elucidated. Fatty acids are synthesized from acetyl coenzyme A and malonyl coenzyme A thiol esters. The vitamin biotin plays a vital part in the fixing of carbon dioxide to form malonyl coenzyme A, an important intermediate in fatty acid synthesis. The hormone insulin also favors fatty acid synthesis. The oxidation of fatty acids occurs as their coenzyme A esters in the Krebs cycle of the mitochondria. Cholesterol is biosynthesized from acetyl coenzyme A. Cholesterol in humans is converted to bile acids, fecal sterols, and to steroid hormones. The synthesis of lecithin is mediated via phosphatidic acid and diglyceride precursors. Cytidine nucleotides play a role in the transfer of choline (as phosphorylcholine) to a diglyceride to form lecithin. Uridine nucleotides act to transfer sugar residues in the synthesis of glycolipids.

The transport of lipids in the blood plasma is effected by complex formation with proteins to yield lipoproteins. The liver is the major organ for the synthesis of the lipoproteins. Analysis of serum lipoprotein patterns is important in the understanding of vascular disease (atherosclerosis). The clearing of lipemic blood, such as may occur after a heavy fat meal, is brought about by enzyme known as lipoprotein lipase. This enzyme yields free fatty acids which combine immediately with the plasma albumin to form complexes known as NEFA (nonesterified fatty acids). NEFA act as important transport vehicles for transport of triglycerides and the levels of blood NEFA are very sensitive to hormonal control and neural control. Certain hormones, such as epinephrine, stimulate the membrane-bound adenyl cyclase which converts ATP (adenosine triphosphate) to cyclic AMP (adenosine monophosphate). The latter stimulates adipose tissue lipase and mobilizes depot fat.

The excess utilization of lipids and excess oxidation of fatty acids causes an increase in acetoacetic acid in the body. This condition is known as *ketosis* and can lead to *acidosis*. This situation is common in severe diabetes and can occur whenever carbohydrate utilization is severely decreased.

Research continues in a effort to gain a more thorough understanding of how lipoproteins are synthesized; how lipids are arranged and combined with proteins to form cell membranes; what specific role lipids play in transport across cell membranes; how hormones act to regulate lipid metabolism; the biochemical basis of such abnormal lipid metabolic states as Gaucher's disease, Niemann-Pick's disease, etc.; how lipids per se permeate cell membranes; and how many phenotypic lipoproteins occur in serum.

See also **Cholesterol** and list of references at end of that article.

LIPOMA. A benign tumor made up of fat cells. Lipomas occur commonly in the subcutaneous tissues about the head and neck. They cause no symptoms.

LIQUATION. 1 A process of magmatic differentiation believed to take place as a result of the separation of two immiscible liquids from the parent magma. 2 The separation of a more readily fusible substance from a less fusible one by controlled heating.

LIQUEFACTION. See **Natural Gas.**

LIQUID CRYSTAL POLYMERS. These materials (LCPs) exhibit a highly ordered structure in the melt, solution, and solid states. A tightly packed and highly ordered morphology particularly susceptible to orientation during processing is characteristic. Commercial applications for LCP resins include chemical pumps, tower packings, coil bobbins, connectors, sockets, etc. for electronic components, and various automotive parts. LCPs have an excellent combination of chemical and flame resistance, dimensional stability, and ease of processing. Their thermal stability makes them suitable for dual ovenable cookware and where thermal resistance for both conventional and microwave oven service is important.

Compared with other polymeric materials, LCPs have very high unidirectional properties. *Vectra*™ (Celanese Corp.) resins are primarily aromatic polyesters based on *p*-hydroxybenzoic acid and hydroxynaphthoic acid monomers. *Xydar*™. (Celanese Corp.) injection molding resins are polyesters based on terephthalic acid, *p,p'*-dihydroxybiphenyl and *p*-hydroxybenzoic acid. Differences in monomers are primarily responsible for the differences in specific properties and end uses. The fibrous nature of the polymers imparts good impact strengths.

LIQUID CRYSTALS. Liquids that have the structural character of cybotactic liquids (see **Cybotaxis**), but which are considerably more viscous, with viscosities extending from that of a light glue to that of a glassy solid. They also exhibit much more definite evidences of structure than the cybotactic liquids.

Liquid crystals must be geometrically highly anisotropic—usually long and narrow—and revert to an isotropic liquid through thermal action (thermotropic mesomorphism) or by the influence of a solvent (lyotropic mesomorphism). Several thousand organic compounds are now known which meet these criteria, but significant molecular features found in thermotropic liquid crystals are among the following. The molecule will be elongated and rectilinear; if "flat segments," e.g. benzene rings, are present its liquid crystallinity will be enhanced. The molecule will be rigid along its long axis and double bonds will be common in this direction. The simultaneous existence is seen in the molecule of strong dipoles and easily polarizable groups. Of lesser importance are weak dipolar groups at the extremities of the molecule.

The present day classification of thermotropic liquid crystals is three-fold. *Smectic liquid crystals*, such as *p*-ethyl azoxybenzoate, have their molecules arranged in definite strata, a variety of molecular arrangements being possible within each stratification. In *smectic Type A* crystals, the molecules may be considered to "stand on end" with their long axes perpendicular to the plane of the layer but with their centers irregularly spaced. When the molecular centers adopt hexagonal close packing, the crystals are considered *smectic Type B*, and when they adopt a titled form of Type A, they are classified as *smectic Type C*. See Fig. 1.

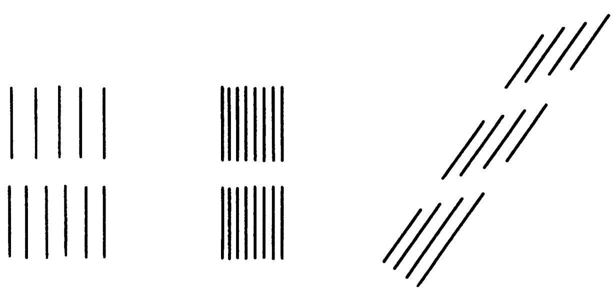

Fig. 1. Smectic liquid crystals; types A, B, and C.

In *nematic liquid crystals*, the molecular structures possess a high degree of long-range orientation order, but no long-range translational order. The molecules are spontaneously oriented with their long axes approximately parallel, but without the stratification seen in smectic crystals. Nematic liquid crystals like *p*-azoxyanisole are generally optically uniaxial, positive, and strongly birefringent, and some are composed of hundreds of molecules (cytotactic groups), the molecular centers in each group arranged in layers. See Fig. 2.

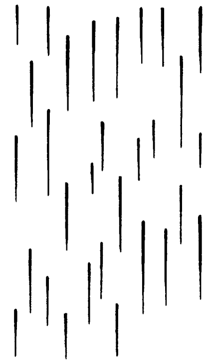

Fig. 2. Nematic crystals.

Lyotropic liquid crystals possess at least two components. One of these is water and the other is amphible (a polar head group attached to one or more long hydrocarbon chains). In the lamellar form, water molecules are sandwiched between the polar heads of adjacent layers while the hydrocarbon tails lie in a non-polar environment. Lyotropic liquid crystals have very complex structures, but occur abundantly in nature, particularly in living systems. See Fig. 3.

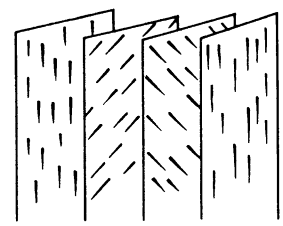

Fig. 3. Cholesteric crystals.

Polarized light is the most powerful tool for investigating liquid crystals, all of which exhibit characteristic optical properties. A smectic liquid crystal transmits light more slowly perpendicular to the layers than parallel to them. Such substances are said to be optically positive. Nematic liquid crystals are also optically positive, but their action is less definite than that of smectic liquid crystals. However, the application of a magnetic field to nematic liquid crystals lines up their molecules, changing their optical properties and even their viscosity.

Both smectic and nematic crystals split a beam of ordinary light into two polarized components whose transverse vibrations are at right angles to each other. This is the well-known phenomenon of double refraction. Cholesteric liquid crystals exhibit the phenomenon of circular dichroism. That is, they break a beam of ordinary light into two components, one with the electric vector rotating clockwise and the other with it rotating counterclockwise. The first is usually transmitted, and the second is the one to be reflected. It is this property that gives cholesteric crystals their characteristic iridescent colors when illuminated by white light.

This ability to exhibit colors is one of the most useful attributes of liquid crystals. Many cholesteric substances behave as liquid crystals only in a certain temperature range. Above it they are colorless, but as they are cooled through it they assume a succession of colors, running down the spectrum from red to violet and finally becoming colorless. However, at this final stage they still retain their molecular orientation, but it is that of smectic liquid crystals rather than cholesteric. Some cholesteric liquid crystals do not exhibit all the colors mentioned and others, which are naturally colored, simply change to another color on heating or cooling. Since the exact temperatures at which these color changes occur are invariable, these substances can be used for measuring temperatures; in fact, combinations of them cover the range from −20 to +250°C.

Useful applications have been found for the varied effects of these crystal changes. One of the first came from the property of selectively reflecting visible light; because this is temperature-dependent, the property can be used as a temperature detector, and in gel form liquid

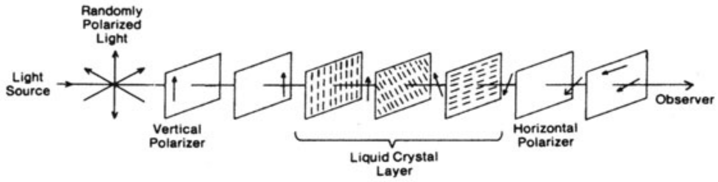

Fig. 4. Liquid crystal display operation, unactivated.

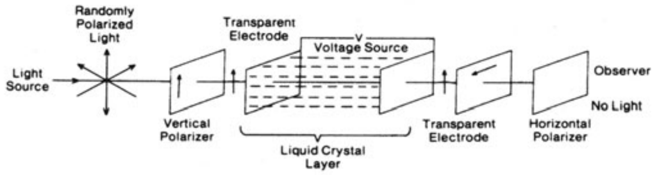

Fig. 5. Liquid crystal display operation, activated.

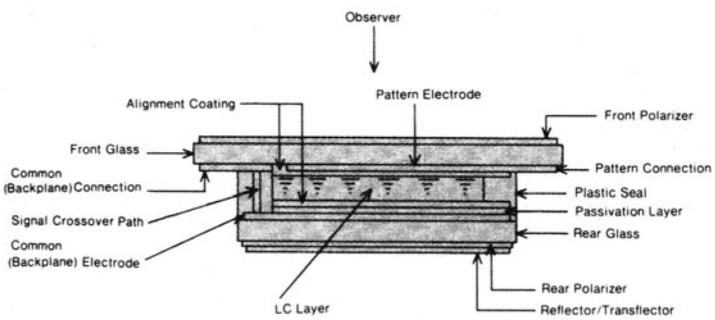

Fig. 6. Basic liquid crystal display construction. (*Hamlin, Inc.*)

crystals have been used for the early detection of those cancers which cause hot spots in the body. Applications of the smectic modifications arise from their ferroelectric properties; this phase can function as a fast-switching light-valve device with memory. This kind of application requires some control on the pitch of the polarized helix which is obtained by blending together materials. Most twisted nematic field effect liquid crystal displays make use of a 90° twist between transparent conductive electrodes and crossed polarizers, as shown in Fig. 4. As randomly polarized light enters the device, only that portion which is vertically polarized may pass through the front polarizer. This, in turn, is rotated another 90° through the rear polarizer. If a reflective surface is placed behind the rear polarizer, the light will be passed back through the cell, its polarization again being rotated. By applying a voltage to the transparent electrodes, the molecules of the crystal will leave the nematic structure and align with the field, as shown in Fig. 5. Then the incoming light is no longer polarized—so extinction occurs and the area of extinction is defined very sharply by the shape of the electrode pattern, producing a dark area on a light reflective background. The reverse can also be achieved by using parallel rather than crossed polarizers. Fig. 6 shows a cross section of a typical display. A polymeric seal contains the liquid crystal material and holds the glass substrates together and the whole is laminated to the assembly.

LIQUID-IN-GLASS THERMOMETER. This instrument consists of a glass envelope, a responsive liquid, and an indicating scale. The envelope is in two parts fused together: a bulb completely filled with the liquid, and a capillary scale section containing the liquid in excess of that required to fill the bulb. The position of the end of the liquid capillary column or index serves, by prior calibration, to indicate the temperature of the bulb. The scale may be marked directly on the capillary tube, as in the laboratory or clinical versions, or may be on a separate member mounted alongside the capillary tube, as in the domestic and industrial forms.

Invented over three centuries ago, the liquid-in-glass thermometer reached its zenith as a temperature measuring device in the early 1800s. It was used as a standard for the dissemination of the scale (Normal Thermometric Scale, adopted internationally) from the International Bureau of Weights and Measures to standardizing laboratories throughout the world, until the later adoption, in 1927, of the International Temperature Scale. Over the years, many practical applications were found for the liquid-in-glass thermometer in addition to its earlier use as a primary standard of temperature. Although still used widely in meteorology, medicine, and industry, and for domestic purposes, the glass thermometer has been replaced by various electrical and electronic temperature measurement methodologies for many other applications.

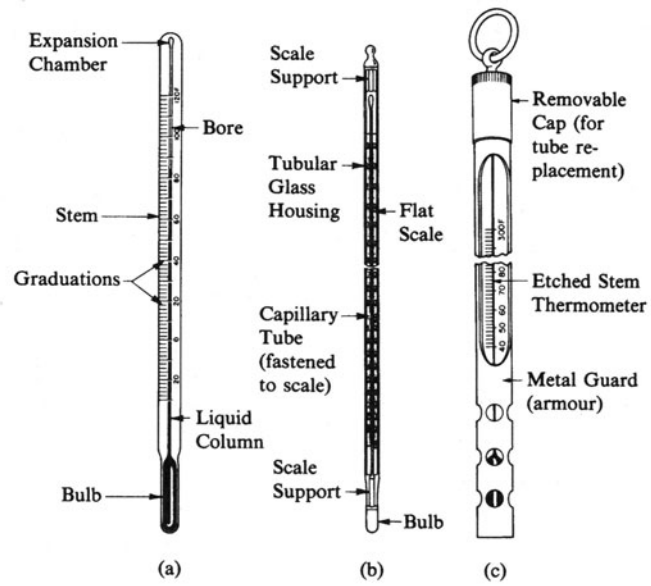

Fig. 1. Laboratory-type liquid-in-glass thermometers: (a) traditional; (b) Einschluss; (c) armored.

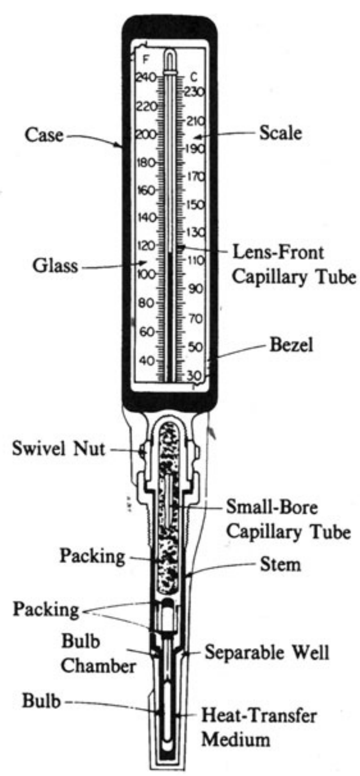

Fig. 2. Industrial-type liquid-in-glass thermometer.

Various liquids are used in liquid-in-glass thermometers. Mercury is the choice for higher temperatures or where accuracy is critical. Its advantages are a broad temperature span between its freezing and boiling points, a nearly linear coefficient of expansion, the relative ease of obtaining mercury in a very pure state, and its nonwetting-of-glass characteristic. For measurements below the freezing point of mercury, various organic liquids, such as toluene, other hydrocarbons, and organic phosphates, have been used. Representative versions of the glass thermometer are shown in Fig. 1. An industrial version is shown in Fig. 2.

LIQUID JUNCTION. To avoid the unknown liquid junction potential in measuring the potential of a half-cell against a reference electrode, the two half-cells are frequently connected via a salt bridge, usually a concentrated solution of potassium chloride. Since its anion and cation have almost the same velocity, a negligible diffusion potential is set up across the liquid junctions at the ends of the bridge.

LIQUID STATE. Because of the theoretical and practical importance to the era of electronics, which commenced nearly a half-century ago, the solid state of matter has become better known and understood than the physics of fluids (liquids and gases). Much practical engineering knowledge has been amassed pertaining to substances in the fluid state, but much research of a fundamental nature on fluids remains to be finished. Particularly, the transition of liquids to solids (and vice versa) at the theoretical level has not been fully explored and explained.

Prestigious scientists have commented on the mysteries that confront them. Russell J. Donnelly (University of Oregon) has observed, "Most flows of fluids, in nature and in technology, are turbulent. Since much of the energy expended by machines and devices that involve fluid flows is spent in overcoming the drag caused by turbulence, there is a strong practical motivation to understand the phenomenon. The study of turbulent flows, however, is one of the most formidably difficult subjects in physics and engineering. At present (1988), there is no substantial aspect of turbulent flow that can be understood fully from first principles."

Steve Granick (University of Illinois) has commented (1991), "Apart from structure, what are the dynamics of liquids in intimate contact with a solid boundary? This question has proven to be one of the most baffling aspects of liquids, in spite of long-standing interest."

Sir Samuel F. Edwards (Cavendish Laboratory, University of Cambridge) noted (1987), "Liquids are everywhere in our lives, in scientific studies and in our everyday existence. The study of their properties, in terms of the molecules of which they are made, has been the graveyard of many theories put forward by physicists and chemists. The modern student of liquids places his faith in the computer, and simulates molecular motion with notable success, but this still leaves a void where simple equations should exist, as are available for gases and solids. There is a powerful reason for the failure of analytical studies of liquids, i.e., the difficulty experienced in finding simple equations for simple liquids. We can explain the origin of the trouble and show that it does not apply to what at first might seem a much more complex system, that of polymer liquids where, instead of molecules like H_2O or C_6H_6, one has systems of molecules like $H_2(CH_2)_{10,000}$ or $H_2(CHC_6H_6)_{2,000}$ which behave like sticky jellies and yet have complex properties that can be predicted successfully."

Jacob N. Israelachvili and Patricia M. McGuiggan (University of California, Santa Barbara) observe, "The subtleties that can occur in the last few nanometers as two surfaces, particles, or solute molecules approach each other in a medium can be quite remarkable. Sometimes the forces are well described by 'continuum' or 'mean-field' theories, such as the DLVO (Derjaguin, Landau, Verwey, and Overbeck) theory, but more often they are not. Important fundamental questions remain concerning the origin of long-range attractive and repulsive hydration forces in water, the spontaneous nucleation of a bulk liquid or vapor phase between two surfaces close together, and the nature of entropic-fluctuation forces between two fluidlike interfaces. The elucidation of these interactions both at the fundamental level and when applied to specific systems (where a number of different interactions may be occurring simultaneously) present a challenge to experimentalists and theoreticians. On the purely experimental side, new techniques are constantly being introduced for extending the range and scope of surface force measurements. For example, one may anticipate that the atomic force microscope will soon provide the first direct measurements of the forces between molecules, as opposed to between surfaces."

General Properties of Liquids

A liquid is matter in a fluid state that is relatively incompressible. An *ideal liquid* offers no permanent resistance to a shear stress, but is incompressible. A liquid has a constant volume and incompletely fills any container of less than this volume. A *real liquid* is appreciably compressible, and the liquid state of a substance might be defined as the denser and less compressible phase of the two-phase fluid system that can exist in equilibrium at temperatures below the critical temperature. X-ray diffraction experiments show that, near the melting-point, the molecules of a liquid show a considerable degree of short-range order and that, in small volumes, they are arranged much as in a solid crystal. This crystalline structure persists over volumes comparable with the intermolecular distances, but cannot be traced beyond. This local or short-range order means that the average molecule is at any moment surrounded by a number of molecules occupying nearly the same relative positions as they would in the solid state. The degree of short-range order is described by the radial distribution function.

This concept of a liquid as an imperfect crystal requires that the molecules in a liquid are packed sufficiently loosely for comparatively free movement, i.e., the energy required to move a molecule from a lattice site to a vacant space is not large compared with thermal energies. Under these conditions, shear flow of the liquid resembles closely the high temperature creep of crystalline solids. A number of theories of the liquid state have this concept as their starting point.

With a few exceptions, including helium, the universal phase diagram shown in Fig. 1 applies for all pure compounds. The triple point is the single point at which all three phases (crystal, liquid, and gas) are in equilibrium. The triple point pressure is normally below atmospheric. Those substances, such as carbon dioxide, where $P_t = 3{,}885$ millimeters, $T_t = -56.6°C$, sublime without melting at atmospheric pressure. From the triple point, the melting curve defines the equilibrium between crystal and liquid, usually rising with small but positive dT/dP, and presumably always with positive dT/dP at sufficiently high P values. The line is believed to extend infinitely without a critical point (it has been followed to $T \cong 16T_c$ for helium, and calculations indicate that hard spheres would show a gas-crystal phase change). The gas-liquid equilibrium line, the vapor pressure curve, has dT/dP always positive and greater than the melting curve. The vapor pressure curve always ends at a critical point, $P = P_c$, $T = T_c$, above which the liquid and gas phase are no longer distinguishable. Since the liquid can be continuously converted into the gas phase without discontinuous change of properties by any path in the P-T diagram passing above the critical point, there is no definite boundary between liquid and gas. Two liquids of similar molecules are usually soluble in all proportions, but very low solubility is sufficiently common to permit the demonstration of as many as seven separate liquid phases in equilibrium at one temperature and pressure (mercury, gallium, phosphorus, perfluoro-kerosene, water, aniline, and heptane at 50°C, 1 atmosphere).

Stability Limits.[1] With the exception of helium and certain apparent exceptions discussed below, Fig. 1 gives a universal phase diagram for all pure compounds. The triple point of one P and one T is the single point at which all three phases, crystal, liquid, and gas, are in equilibrium. The triple point pressure is normally below atmospheric. Those substances, e.g., CO_2, $P_t = 3885$ mm, $T_t = -56.6°C$, for which it lies above, sublime without melting at atmospheric pressure.

From the triple point, the melting curve defines the equilibrium between crystal and liquid, usually rising with small but positive dT/dP, and presumably always with positive dT/dP at sufficiently high P values. The line is believed to extend infinitely without a critical point

[1] The following several paragraphs by Joseph E. Mayer are part of a large article that appears in "The Encyclopedia of Physics" (Robert M. Besancon, Editor), Van Nostrand Reinhold, New York, 1984.

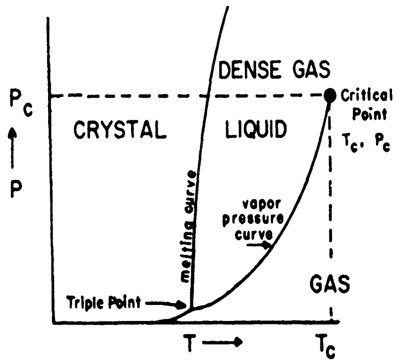

Fig. 1. Universal phase diagram.

(it has been followed to $T \cong 16T_c$ for He, and calculations indicate that hard spheres would show a gas-crystal phase change). The gas-liquid equilibrium line, the vapor pressure curve, has dT/dP always positive and greater than the melting curve. The vapor pressure curve always ends at a critical point. $P = P_c$, $T = T_c$ above which the liquid and gas phase are no longer distinguishable. Since the liquid can be continuously converted into the gas phase without discontinuous change of properties by any path in the P-T diagram passing above the critical point, there is no definite boundary between liquid and gas.

The term *liquid* is commonly reserved for $T < T_c$, and "dense gas" is used for $T > T_c$. However, certain properties, such as the ability to dissolve solids, change rather abruptly at the critical density. In many respects, the dense gas resembles the low-temperature liquid of the same density more closely than it does the dilute gas.

The slope, dT/dP, of all phase equilibrium lines obeys the thermodynamic Clapeyron equation:

$$dT/dP = \Delta V/\Delta S = T\Delta V/\Delta H \tag{1}$$

with ΔV, ΔS, and ΔH the differences, for the two phases, of volume, entropy, and heat content or enthalpy, respectively. The quantity ΔH is the heat absorbed in the phase change at constant P. Since always $S_{cr} < S_{liq} < S_{gas}$, and usually $V_{cr} < V_{liq} < V_{gas}$, one usually has $dT/dP > 0$; the relatively rare cases, including water, for which $V_{liq} < V_{cr}$ at low pressures leads to $dT/dP < 0$ for the melting curve near the triple point.

Figure 1 gives the P-T boundaries of the stable liquid phase. Clean liquids can readily be superheated or supercooled, and, in vessels having walls to which the liquid adheres, they can be made to support negative pressures of several tens of atmospheres. Thus the properties of the metastable liquid can be investigated outside the limits shown in the diagram.

Two apparent exceptions to the universality of the phase diagram of Fig. 1 deserve mention. First, many of the more complicated molecules decompose at temperatures below melting or boiling, and the diagram is unobservable. Secondly, some liquids, notably glycerine and SiO_2 and many multicomponent solutions, supercool so readily that crystallization is difficult to observe. In these cases, there is a continuous transition on cooling to a glass, which has the elastic properties of an isotropic solid. The structure of the glass is qualitatively that of the high-temperature liquid, lacking long-range order. Since glass and liquid are not sharply differentiated, the term *liquid* is sometimes used to include glasses, although common parlance reserves *liquid* for the state in which flow is relatively rapid.

Quantum Liquids. The one real exception to the phase diagram of Fig. 1 is that of helium, Fig. 2. Both isotopes, ^4He and ^3He, have no triple point, the liquid is stable to 0 K below about 20 atm for ^4He and below about 30 atm for ^3He. The liquids have zero entropy at 0 K in both cases. This is also the only case in which isotopic mixtures form two liquid phases at equilibrium, the isotopic solution separating below 1 K. The isotope ^4He has itself two phases, He I above the dotted λ-line of the diagram, and He II with remarkable properties of superfluidity, second sound, etc., below the λ-line. The phase transition along the λ-line is second order; that is, whereas S and V are continuous, heat capacity and compressibility change discontinuously across the λ-line.

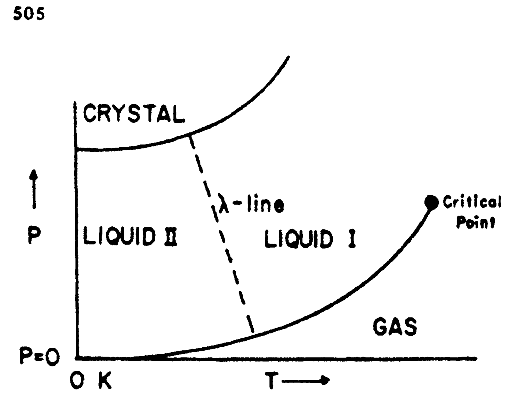

Fig. 2. Quantum liquid exception to phase diagram of Fig. 1.

Although no completely satisfactory single theory of liquid helium has yet been formulated, one can say that most of the remarkable properties are qualitatively understood and are due to the predominance of quantum effects, including the difference in the statistics of the even and odd isotopes. Thus helium is the one example in nature of a quantum liquid, all other liquids showing only minor deviations from classical behavior.

Structure. Considerable confusion in the description of liquid structure exists, due primarily to difficulties of precise formulation of verbal concepts. The geometric arrangement of any small number (say 10 to 12) of close-lying molecules resembles the arrangement in the crystal, but the order rapidly disappears as larger groups are considered. Long-range order is lacking. The fact that numerical theories based on a lattice or cell structure have some success is evidence only that most properties depend on the configuration of near neighbors alone. Insofar as the arrangement of nearest neighbors is describable in terms of that of the crystal, the structure of the normal liquid is probably characterized best by a somewhat closer spacing than the crystal of the same molecules, the reduced density arising from a considerable number of vacancies in the lattice; the coordination number, or number of nearest neighbors, is lower than in the crystal. The exception is water, in which the low coordination number, 4, of the crystal, is increased by interstitial molecules in the liquid, leading to a higher density of the liquid.

Structural descriptions of this nature usually lack the possibility of precise formulation. It is, however, possible to define for any disordered array of molecules in three-dimensional space an arrangement of contiguous cells, each containing one and only one molecule, the faces of the cells being the loci of the midpoints of neighboring molecules. The statistics of the fraction of cells with n faces and of the distances of the faces from the molecules would give the fraction of molecules having a given number of nearest neighbors and the distance distribution of these in a precisely defined manner. Neither present experimental information nor present theories lend themselves to analysis in such terms.

The only clearly defined manner of describing liquid structure in use at present involves the concept of a set of probability density functions, ρ_n, for ascending numbers, n, of molecules. The function ρ_n depends on the vector coordinates $\mathbf{r}_1, \mathbf{r}_2, \cdots \mathbf{r}_n$ of n molecules, and

$$\rho_n\, \mathbf{r}_1, \mathbf{r}_2, \cdots, \mathbf{r}_n, d\mathbf{r}_1, \cdots, d\mathbf{r}_n$$

is defined as being the probability that in the liquid of definite P and T, there will be, at any instant of time, one molecule at each position, $\mathbf{r}_i$, within the volume element, $d\mathbf{r}_i$. For a fluid, unlike a perfect single crystal, $\rho_i(\mathbf{r})$ is a constant independent of $\mathbf{r}$ and equal to the number density:

the number, ρ, of molecules per unit volume. The first significant member of the set is then the pair density function, $\rho_2(\mathbf{r}_1, \mathbf{r}_2)$, which depends only on the distance, $\mathbf{r} = |\mathbf{r}_1 - \mathbf{r}_2|$, between the two molecules. At large distances $\rho_2(\mathbf{r} \to \infty) = \rho^2$. This function can be found experimentally from the x-ray scattering intensities of the liquid (it is the three-dimensional Fourier transform of the scattering intensity at angle θ vs $(4\pi/\lambda)/\sin(\theta/2)$). A typical plot is shown in Fig. 3. The area under the ill-defined first peak integrated over $4\pi\mathbf{r}^2\,d\mathbf{r}$ is the average number of nearest neighbors, and is of order 10 to 11 for normal liquids.

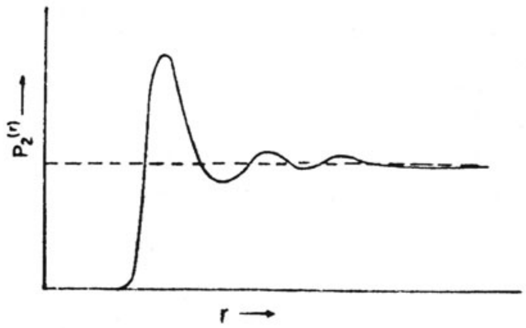

Fig. 3. Fourier transform of scattering intensity.

The quantity of dimensions of energy,

$$W_n(\mathbf{r}_1, \cdots, \mathbf{r}_n) = -kT \ln[\rho^{-n} \rho_n(\mathbf{r}_1, \cdots, \mathbf{r}_n)]$$

can be shown to be the potential of average force of n molecules located at the positions $\mathbf{r}_1, \cdots, \mathbf{r}_n$. That is, if there are n molecules at these positions there will be some average force, f_{xi}, along the x-coordinate of molecule i. This average is the sum of the direct force due to the other $n-1$ plus the average of a fluctuating force due to the others, whose average position is affected by that of the n specified ones. This average force is

$$f_{xi} = -(\partial/\partial x_i)\, W_n(\mathbf{r}_1, \cdots, \mathbf{r}_n)$$

One frequently assumes that W_n is a sum of pair forces only,

$$W_n(\mathbf{r}_1, \cdots, \mathbf{r}_n) = \sum_{n \geq i > j \geq 1} \sum{}' W_2(\mathbf{r}_{ij})$$

although this assumption is known to be only approximate. With this assumption, the pair average force potential, $W_2(\mathbf{r}_{ij})$, can be computed as the solution of an integral equation, and the solutions agree quite well with the experimental curves.

The knowledge of the complete set of functions ρ_n plus that of the intermolecular forces would permit the computation of all equilibrium properties of the liquid, and indeed if the intermolecular forces are the sum of pair forces, only a knowledge of ρ_2 at all P, T values is necessary. An adequate, although numerically difficult, theory of the transport properties also exists, using the equilibrium functions, ρ_n. At present, only qualitative success is obtained in the completely *a priori* use of the equations.

Associated Liquids. The description given above is adequate only for liquids composed of spherically symmetric molecules or molecules that are nearly so. These constitute the so-called normal liquids, which obey reasonably well the law of corresponding states, for which the entropy of vaporization at the boiling point has the Trouton's rule value of approximately 21 cal/deg. For molecules containing large dipole moments, or those forming mutual hydrogen bonds, the concept of the probability density functions must be extended to include angles or other internal degrees of freedom in the coordinates. Such inclusion is conceptually easy, but incredibly complicates the already difficult numerical evaluation of any equations. However, certain qualitative statements may be made.

Liquids composed of molecules with large dipole moments are frequently referred to as associated. Although in some instances relatively stable dimer or definite polymer units of relatively fixed orientation may exist, in many cases, notably water, it is extremely doubtful if an exact knowledge of the structure would reveal any distinguishable entities of associated molecules other than that of the whole liquid. In such cases, one would, however, expect that certain mutual angular orientations between neighboring molecules will be highly preferred, whereas in the dilute gas this will not be the case. The effect of this restriction on the internal coordinates will be to decrease the entropy of the liquid markedly compared to the gas. This effect is qualitatively the same as in association, and the properties of these liquids, particularly the high entropy of vaporization, will simulate those of a liquid composed of definite associated complexes.

Traditional Views of Forces Between Surfaces in Liquids

For many years, four kinds of forces have been recognized to operate between surfaces or particles in liquids:

1. *Van Der Waals forces*—Normally, these are monotonically attractive and occur between all molecules. See **Van Der Waals Forces.**
2. *Repulsive electrostatic (double-layer) forces*—These forces are apparent when ionizable surfaces have a net electric charge, the common case in water. See **Electrostatics.**
3. *Structural, hydration, or solvation forces*—These forces may be attractive, repulsive, or oscillatory and depend upon the structuring or ordering of liquid molecules. (Solvation may be defined as the adsorption of a microlayer of film of water or other solvent on individual dispersed particles of a solution or dispersion.)
4. *Repulsive entropic (steric or fluctuation) forces*—As defined by Israelachvili and McGuiggan, these are "forces which arise from the thermal motions of protruding surface groups (such as polymers or lipid head groups) or from the thermal fluctuations of flexible fluidlike interfaces (or surfactant or lipid bilayers)."

Although, in a vacuum, only the Van Der Waals forces are important; in liquids, all forces may operate simultaneously. In liquids, it is extremely difficult to separate the effects of each of the aforementioned forces.

In the 1950s, the DLVO theory was based largely upon forces (1) and (2) defined above. The DLVO theory became the basis for studying the properties of colloidal and biocolloidal systems.

Electrorheological Fluids

Complications continue in the theoretical exploration of liquid behavior, but, in attempting to learn about the complexities, leads toward a more fundamental understanding of liquids may emerge. One of these complexities is a class of fluids referred to as *electrorheological.*

In a normal setting, these liquids are liquid in the conventional sense, but, when they are subjected to a strong electric field, they become solids. A common example is a mixture of cornstarch and vegetable oil. The viscous, sticky starting mixture of these two components can be converted into a hardened solid material with the application of an appropriate electric field. In recent years, through random researching, investigators have found numerous combinations of materials that qualify as electrorheological fluids, but to date no satisfactory explanation of the effect from a theoretical standpoint has been developed. The effect of the ER effect can be observed readily by microscopic examination. In a normal liquid or in an electrorheological mixture not in an electric field, examination shows particles moving in random fashion throughout a container, as may be expected. But, when a field is applied, long strands of particles appear to be "solidly" linked together, thus providing rigidity to the mix. No retentivity is involved, however, because liquid normalcy is returned immediately upon cessation of the electric field. This action intrigues a number of designers of equipment, as in the electronics and valve fields, where such materials may be used in future circuits, valves, and any number of other "on-off" devices.

Artificial Magnetic Fluids. The concept of magnetizable liquids dates back over several decades. The first breakthrough occurred in the mid-1960s, with the production of stable colloids of subdomain solid ferromagnets, which variously were called magnetic fluids, magnetizable fluids, or simply ferrofluids. These may be prepared by size reduction or precipitation. It has been a rather remarkable feat to grind bulk material down to a size of 100 micrometers. Grinding is done in the presence of a dispersing agent and a solvent. In chemical precipitation, iron(II) and iron(III) ions in aqueous solution are coprecipitated in the molar ratio of about 2:1 using ammonium hydroxide. To maintain the magnetic product in a small colloidal size range, a peptization step is included in which the particles are transferred to a heated organic phase containing the dispersing agent. The behavior of these artificial fluids offers techniques for achieving efficient heat and mass transfer, drag reduction, wetting, fluidization, sealing, damping, and other process and product potential uses.

Polymer Liquids

Sir Samuel F. Edwards of the Cavendish Laboratory, previously mentioned, observes, "Polymer liquids are liquid because the temperature is high enough to change the configuration of the molecules easily." See Fig. 4.

Scientists and engineers experienced in the production and application of polymeric liquids have learned from thousands of examples of how polymeric liquids behave and can even classify some of them into behavioral families, but the complexity of these substances to date has eluded the achievement of precise designs and predictive behavior.

Mixing of Liquids

In the chemical process industries and in food manufacturing, the mixing of liquids is an important and frequently used operation. Over the years, the mixing operation has been poorly understood and essentially remains so. Progress in equipment design has been achieved mainly through the development of myriads of empirical information rather than upon the creation of precise mathematical and theoretical relationships. J. M. Ottino (University of Massachusetts) observes, "There are many fundamental questions regarding mixing in slow three-dimensional flows, and unfortunately some of the intuition we have obtained from our study of two-dimensional flows does not necessarily carry over to flows in three dimensions." The Ottino reference listed describes studies of chaotic and nonchaotic flows in laboratory setups.

Multiphase Fluid Flow

It is quite common in industrial and cross-country liquid transport problems to encounter mixtures of liquids, vapors, and gases—that is, the presence of two phases of matter. As pointed out by D. I. Koch (Cornell University), "Research programs in this area at Cornell involve studies that fall outside of the traditional realm of chemical engineering, including blood flow in capillaries, the transport of contaminants in groundwater, the dynamics of geothermal reservoirs, enhanced oil recovery, the processing of fibrous composites, melt-spinning processes, and the growth of silicon crystals."

As previously mentioned in this article, liquid behavior presents an immense variety of puzzling problems that are difficult to comprehend and hence difficult to forecast precisely. As just one example, in the study of large drops of liquid at high flow rates, the inertia of the drops and the surrounding fluids play an important role. When such drops are propelled toward one another, they may coalesce into a single drop or may rebound like a pair of elastic balls, a phenomenon that is partially (but not fully) dependent upon the comparative velocities of the two drops.

In traditional industrial two-phase flow situations, as occur in process piping, engineers classify flow patterns, as indicated in Fig. 5. This type of classification and the development of empirical data from past experimentation and practice assist much in simplifying the problems from a practical, if not from a theoretical, standpoint. Multiphase flow behavior considerations are essential in calculating pipe diameters, pumping capacities, and energy consumption. Multiphase flows also exhibit different characteristics when being pumped uphill or flowing downhill. See also **Fluid and Fluid Flow.**

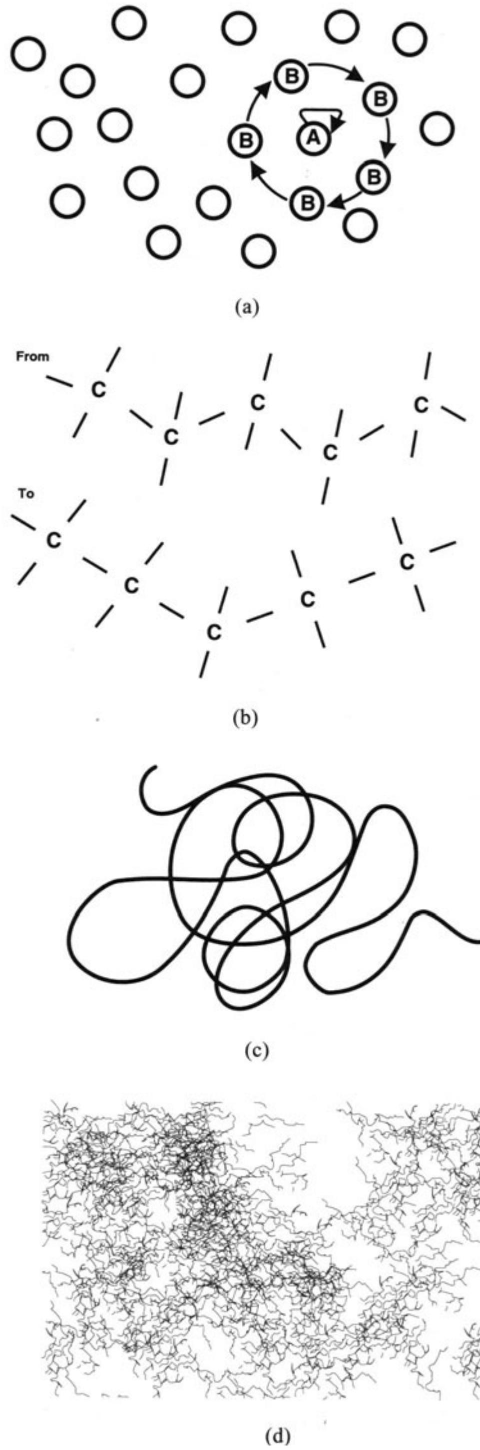

Fig. 4. Sketches by Edwards for schematically illustrating the behavior of polymer liquids. (a) Central molecule A moves around a (temporary) average position, occasionally escaping the barrier of the molecules B when some fluctuation of their positions permits it to do so. But, in addition to this single molecule motion, the molecules B can move around cooperatively, one of many cooperative motions that the observer may "invent." The numerous possibilities immediately derail the formulation of a simple theory. As explained by Edwards, "If the motion only involved one molecule at a time, quite a reasonable theory can be put together, but it will always be inadequate to describe the whole motion, and possibly, this always will be the case. The very apparatus of mathematical equations cannot handle this level of complexity, even though the human mind has no difficulty in seeing where the trouble is. The best current way to deal with this problem is to put all the molecules onto the computer and study their molecular dynamics." Continual changes are possible, and the long-chain molecule can be regarded as a flexible string in continual motion. A single coil may be drawn, as shown in (c). A computer model could be generated to show many polymers projected onto a plane, as indicated in (d). Edwards observes, "So the liquid looks like a seething pit of wriggling motion, a kind of living spaghetti, but with less smooth shapes than real spaghetti. How can one describe such a system? How do the molecules move, and how does the material flow?" (*Sketches after Sir Samuel F. Edwards.*)

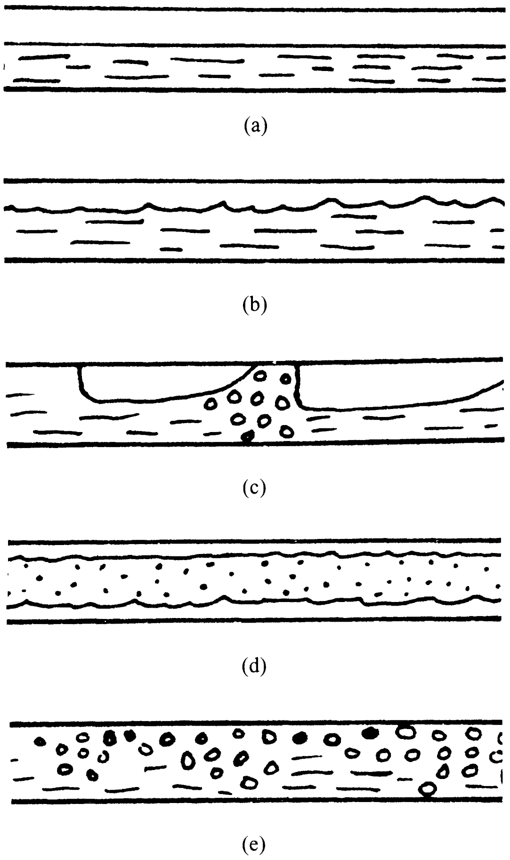

Fig. 5. Types of two-phase flow in a horizontal pipeline: (a) Stratified smooth flow where gas velocity is low. Liquid flows along bottom portion of pipelines with essentially a smooth surface. (b) Stratified flow with a wavy surface, the waviness caused by increased gas flow velocity. (c) Liquid bridges the pipeline cross section, thus causing slugs or plugs of liquid, which move at a velocity approximately that of the flowing gas. (d) Annular flow, in which the liquid essentially flows as an annular film on the pipe wall while gas flows as in a central core of the pipe. (e) Dispersed bubble flow usually results when liquid flow rates are high and gas rates are low. Because of comparative density differences, most bubbles are found above the pipe center line. Conditions vary somewhat when the pipeline is in a vetical orientation. (*After Cindric, Gandhi, and Williams.*)

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LIQUIDUS CURVE. In a temperature-concentration diagram, the line connecting the temperatures at which fusion is just completed for the various compositions.

LISTERIOSIS. A disease of animals including humans caused by *Listeria monocytogenes*, a thin Gram-positive bacillus having several serotypes. The organism is a soil saprophyte which is present in the intestines of many animals and birds. These reservoirs are potential sources of exposure to humans, but despite the ubiquity of *Listeria sp.*, human disease caused by the organism is uncommon. Most infections occur in the first month of life where, in early onset, the mortality may be as high as 40–60%; or beyond the age of 55, where there is usually an underlying predisposing illness.

The most common form of listeriosis is meningitis, with bacteremia occurring in 5–30% of cases, and endocarditis, osteomyelitis, and cholecystitis also sometimes evidenced.

Ampicillin or penicillin are the antibiotics of choice, with treatment being continued for at least ten days after the patient becomes afebrile.

R. C. V.

LITCHI TREE. Of the family *Sapindaceae* (soapberry family), the *Litchi chinensis* is probably best known for its fruit, which when dried is called the litchi "nut." A native of southern China, the tree has spread extensively in cultivation through many southern Asiatic countries. The tree has pinnately compound leaves, the leaflets of which are lanceolate and leathery. The small flowers are borne in panicles, and have no petals. The fruit is roughly spherical, 1 to $1\frac{1}{2}$ inches in diameter, with a hard, brittle rind. Within this rind is a fleshy, translucent pulp, the aril, the part which is eaten. When fresh, it is delectable; when dried into "nuts," it is much shrunken. The fruit contains a single seed.

LITHIFICATION. To lithify is, literally, to turn to stone. Lithification is a term commonly applied to the consolidation and hardening of sediments so as to form a sedimentary rock.

LITHIUM. Chemical element, symbol Li, at. no. 3, at. wt. 6.941, periodic table group 1, mp 180.54°C, bp 1342°C (at 760 torr), density 0.534 g/cm^3 (20°C). Lithium is lightest in weight of all the chemical elements that are solid at standard conditions. Elemental lithium in the solid phase has a body-centered cubic crystal structure. In comparison with other members of the alkali metal series, lithium has the smallest ionic radius, the highest ionization potential, the highest electronegativity, and the greatest heat capacity. Generally, lithium is the least reactive of the alkali metals. Lithium is a silver-white metal, harder than sodium, but softer than lead. It is tough and may be drawn into wire or rolled into sheets. The element tarnishes rapidly in air and often is preserved under naphtha. The reaction with H_2O is vigorous, producing LiOH (lithium hydroxide) and hydrogen. There are two naturally occurring isotopes, ^{6}Li and ^{7}Li. They are not radioactive. Two radioactive isotopes have been identified, ^{5}Li and ^{8}Li, both with very short half-lives, measured in fractions of seconds. Among elements occurring naturally in the earth's crust, lithium ranks 28th with an estimated average content of about 10–20 ppm. In terms of content in seawater, lithium ranks 17th with an estimated content of approximately 950 tons of lithium per cubic mile of seawater. The element was first identified by Johann August Arfvedson in 1817 in the laboratory of Berzelius. The name of the element is accredited to Berzelius.

First ionization potential 5.39 eV. Oxidation potential Li $\rightarrow$ Li^+ + e^-, 3.02 V. Other physical properties of lithium are given under **Chemical Elements.**

The main sources of lithium are pegmatites and brines. The most important pegmatite mineral is *spodumene*, $LiAlSi_2O_6$, which contains a theoretical content of 8.03% Li_2O. *Petalite*, $LiAlSi_4O_{10}$, contains between 4 and 4.5% Li_2O. *Lepidolite*, a complex mica, contains between 3 and 4% Li_2O. See also entries on *Lepidolite; Petalite;* and *Spodumene*. Brines contain normally a few hundred to a few thousand parts per million (ppm) of lithium. The only commercial source of spodumene in North America is located in North Carolina. Abundant resources of lithium pegmatites occur in Canada, the African continent, and unconfirmed sources in Russia and China. Significant quantities of lithium (as carbonate) are produced from the brines of Clayton Valley, Nevada. A recently discovered, lithium-rich brine deposit has been located in the Atacama desert of Chile. More detail on lithium resources is given in the next entry.

There are three major processes for extracting lithium from pegmatite ores: (1) An acid process, wherein the spodumene concentrate, after calcining at about 1095°C, is reacted with sulfuric acid, followed by water leaching of the resulting lithium sulfate, Li_2SO_4. The sulfate is then converted to the carbonate with soda ash. (2) An alkaline process, wherein the ore is reacted with lime or limestone at high temperatures followed by water leaching of the resulting lithium hydroxide. (3) A base exchange method, whereby the ore is reacted with an alkaline chloride or sulfate at a high temperature in an aqueous phase to yield a soluble lithium salt. The sulfuric acid leaching method is the only commercial process for extraction of lithium from spodumene in practice today.

Lithium metal was first prepared by Sir Humphry Davy in 1818 by electrolyzing lithium oxide. At about that same time, Brande also isolated the metal. In 1855, R. Bunsen and A. Matthiessen prepared gram amounts by electrolyzing fused lithium chloride. Modest commercial quantities were first made in Germany during World War I when the metal was considered as a potential alloying material. Limited production did not commence in the United States until the early 1930s. Present commercial methods were pioneered by Guntz in 1893 and involve electrolyzing a low-melting mixture of LiCl and KCl. Graphite anodes and mild steel cathodes are used. Lithium is formed at the cathode and rises to the surface, from which it is skimmed periodically. Pure lithium chloride is added to the bath as required. Chlorine gas is liberated at the anodes. The process yields a lithium metal of about 99.8% purity. The metal normally is cast into ingots of different sizes, but is also available as extruded rod, ribbon, or wire. The metal also is available as "sand"—fine dispersions in the 10–30 μm range.

Lithium in Metallurgy and Alloys

In metallurgy, lithium metal is used as a deoxidizer, desulfurizer, and degasifier in the production of a number of molten metals, notably copper and copper alloys. Lithium also is an ingredient of an increasing number of alloys, particularly with aluminum. Early alloys included aluminum alloy X2020 (1% Li), which is a structural alloy with improved high-temperature strength. In another early Li alloy, about 14% Li is alloyed with magnesium in the LA 141 alloy, designed for very lightweight structural applications, notably in aerospace applications.

In late 1989, a new proprietary (Martin Marietta Corp.) family of weldable, high-strength Al-Li alloys was introduced. With a 690-MPa (100×10^3 psi) yield strength, the material is claimed to be twice as strong as the previous leading Al-Li alloys. This alloy was developed specifically for space launch systems. The alloy is claimed to maintain a high strength under thermal conditions ranging from cryogenic to elevated temperatures. A primary use is for fuel and oxidizer tanks, where its weldability is a marked advantage. Sheet, plate, extrusion, and ingot products of the new alloy also are available.

The addition of lithium to aluminum castings has been found to be particularly advantageous. Lithium produces a lower density and higher stiffness over conventional aluminum alloys used in aerospace applications. Lithium has one of the highest solubilities of any aluminum alloying element. About 4.2% Li can be dissolved in Al at the 602°C (1116°F) eutectic temperature. The hardness of Al-Li alloys improves with aging temperature. Al_3Li precipitates are formed, producing higher hardness. Yield strength also increases with higher aging temperature and higher Li content.

Lithium alloyed with silver has been used for fluxless brazing.

Lithium Batteries. For many years, lithium has been considered for use in batteries, particularly with the growing emphasis on the electric car. See also **Battery**.

Chemistry and Compounds

Lithium has the highest ionization potential (i.e., Li $\rightarrow$ Li^+, in the vapor) of the alkali metals. However, the measured value of its oxidation potential against a normal aqueous solution of its ion is 3.02 V, which does not differ from those of the other main group I metals by as much as the difference in ionization potentials. That difference, attributed to the high heat of hydration of Li^+, explains why lithium is a vigorous reductant in aqueous systems, but reacts slowly with H_2O, and not at all with dry oxygen except above 100°C.

The single $2s$ electron in the outer shell of lithium is easily removed to form the positive ion, and stability of the remaining $1s^2$ electron pair requires too high a potential (75.62 eV) for any further ionization (by chemical means) so that lithium is exclusively monovalent in its compounds.

Because of the reactivity of lithium with water to form its hydroxide, LiOH, and hydrogen, its properties when dissolved in other solvents have been studied extensively. It does not decompose liquid NH_3, but does form a blue solution, which decomposes to yield its amide, $LiNH_2$, and hydrogen, when catalyzed by metallic salts. With the elements of main groups 2 to 7, lithium in liquid NH_3 reacts to form binary compounds, which may vary from simple halides, as with the halogens, to intermetallic phases, as with cadmium and mercury. Lithium amide in liquid NH_3 is regarded in the same class as a hydroxide in aqueous solution.

Many other lithium compounds not obtainable in aqueous solution can be produced from the solution of lithium in liquid ammonia. Thus the acetylide is obtained by action of acetylene.

$$C_2H_2 + LiNH_2 \rightarrow LiC_2H + NH_3$$
$$2LiC_2H \rightarrow Li_2C_2 + C_2H_2$$

The amide, as stated above, is produced by catalyzed decomposition of the liquid NH_3 solution, and the nitride, Li_3N, by heating the amide or by direct combination of the elements.

Lithium salts exhibit general high solubility and a high degree of dissociation in other nonaqueous solvents than liquid ammonia, such as liquid sulfur dioxide and acetic acid.

Like the other alkali metals lithium forms compounds with virtually all of the anions, inorganic as well as organic. The lithium salts are in many instances different in their solubility properties from the corresponding salts of the other alkali metals. Thus lithium fluoride, phosphate, and carbonate are the least soluble alkali metal fluoride, phosphate, and carbonate, the solubilities for the other alkali metals increasing with increasing ionic radius. Lithium chlorate and dichromate are, on the other hand, the most soluble alkali chlorate and dichromate, the solubilities for the other alkali metals decreasing with increasing ionic radius. These differences are partly explained, as was that in the oxidation potential, by the considerable hydration of the lithium ion, which also explains the fact that lithium salts generally crystallize as hydrates. Lithium salts, probably because of the small size of the lithium ion, do not form mixed crystals with the other alkali salts, but they do form double salts, notably the two series of lithium-sodium and lithium-potassium sulfates.

Lithium forms several organic compounds. Most of them are lithium salts or lithium acid salts of organic acids or other oxygen-connected lithium compounds. The number of lithium-carbon bonded compounds that have been reported is very small including, in addition to the carbide, methyllithium, CH_3Li, ethyllithium, C_2H_5Li, *n*-propyllithium, C_3H_7Li, *n*-butyllithium, C_4H_9Li, benzyllithium, $C_6H_5{\cdot}CH_2Li$, and methylenedilithium $LiCH_2Li$.

The alkyllithium compounds are usually colorless, soluble in organic solvents, and capable of distillation or sublimation. They are nonelectrolytes and are widely used in synthetic organic chemistry, since, like other lithium compounds, they resemble in their properties the corresponding magnesium compounds.

Lithium carbonate. Li_2CO_3, mp 72.6°C, slightly soluble in H_2O. Used in glass, enamel, and ceramic formulations, in the electrowinning of aluminum, and in the manufacture of other lithium compounds. The compound also has been used in the treatment of manic-depressive psychoses.

Lithium hydride. LiH, mp 686.4°C, reacts vigorously with H_2O. With NH_3, it forms the amide. The compound is used to produce $LiAlH_4$ and other double hydrides. Lithium hydride is an excellent lightweight source of hydrogen. One pound yields 45 cubic feet of hydrogen (one kilogram yields 2.8 cubic meters of hydrogen) at standard conditions. The compound also can serve as a light-weight shield for thermal neutrons.

Lithium hydroxide monohydrate. $LiOH{\cdot}H_2O$ loses water at 101°C. LiOH melts at 450°C. The compound is soluble in water. The compound is used in the formulation of lithium soaps used in multipurpose greases; also in the manufacture of various lithium salts; and as an additive to the electrolyte of alkaline storage batteries. LiOH also is an efficient, lightweight absorbent for carbon dioxide.

Lithium bromide. LiBr, mp 550°C, soluble in H_2O or alcohols. The compound is very hygroscopic and forms four hydrates. Major use has been in absorption-refrigeration air-conditioning systems in which H_2O is the refrigerant—strong LiBr is used to absorb H_2O vapor.

Lithium chloride. LiCl, mp 608°C, soluble in H_2O or alcohols. Very hygroscopic and forms four hydrates like the bromide. The compound is a component of brazing fluxes for aluminum and magnesium; is used in dehumidification systems, as an additive to the electrolyte of dry cells for low-temperature applications, is used in low-freezing fire-extinguishing systems; as an ingredient of fused-salt baths to lower fusing temperature; and, as a coating, in humidity-sensing instruments.

Lithium fluoride. LiF, mp 848°C, soluble in H_2O (slight). Used in enamel and glass formulations; as a component of welding and brazing fluxes; in the electrowinning of aluminum; and as an ingredient of molten salts.

Lithium for Thermonuclear Fusion Reactors

The possible long range role of lithium in fusion reactor technology has caused concern over the future availability of lithium. As early as 1975, a lithium subpanel, formed under the auspices of the National Academies of Science and Engineering, evaluated the Li raw materials available in the western world. The study indicated that reserves (adjusted for mining losses) are 2.54×10^6 metric tons. This figure does not include potentially large sources contained in South American salares, geothermal brines, oil field brines, and lithium-rich clays. Tritium, required in the fusion reactor core, would be produced from lithium. Lithium will be used in the blanket surrounding the core. Assuming 100% efficiency, the Li consumption for a 1000-MW(e) fusion power plant would be approximately 200 kg/year. See also **Nuclear Power.**

Lithium in Biological Systems

Although much remains to be learned, there is considerable evidence that lithium can play an active role (positive and negative) in biological systems. Possibly most widely known is the use of lithium salts, notably lithium carbonate, in the therapy for mania (a condition where the patient is mentally and physically hyperactive, associated with an elevated mood and disorganized behavior). Frequently associated with mania is the broad swinging of the patient's mood (*bipolar disorder*). Studies have shown that persons with mania have a defect in the transmissions of impulses between and along nerve cells in the brain, which depends upon the regulated movement of ions across the membranes of those cells. Lithium antagonizes synaptic transmission of catecholamines in the brain by inhibiting norepinephrine and dopamine release. This is the result of weakly increasing their re-uptake by the presynaptic neuron and by decreasing storage. Lithium also interferes with the ability of several hormones to stimulate adenylate cyclase, a property that is believed to decrease the action of catecholamines at the postsynaptic receptor sites. See also **Nervous System and Brain**.

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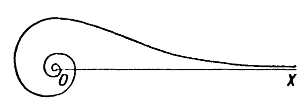
LITHOLOGY. Literally, the graphic study of rocks, hence a synonym for petrography, but not petrology. This term is usually restricted, however, to the purely descriptive macroscopic study of rocks, without the aid of the petrographic microscope.

LITHOPRISM. See **Prism (Optical).**

LITTORAL. Inhabiting the shore line of the ocean in shallow waters and in the tidal zone which is periodically exposed to the air.

LITUUS. A transcendental plane curve, a special kind of spiral. Its equation in polar coordinates is $r^2\theta = a^2$. It begins at infinity, constantly approaches the pole but never reaches it, and has the polar axis as an asymptote.



Lituus.

See also **Curve (Plane).**

LIVER. The largest and one of the most complex organs in the human body, consisting of four lobes and located in the upper abdominal cavity. The liver performs multiple functions, including: (1) secretion and excretion; (2) blood-related activities, including regulation of blood volume and the formation and disposal of various blood components; (3) storage for certain vitamins and minerals; (4) metabolic functions, including fat and protein processing; and (5) detoxification. The approximate size of the adult human liver is 8–9 inches (20–22.5 centimeters) side to side, 4–5 inches (10–12.5 centimeters) front to back, and 6–7 inches (15–17.5 centimeters) top to bottom along the thickest portion. The organ weighs between 42 and 56 ounces (1.2 and 1.6 kilograms). Five ligaments attach the liver to the anterior walls of the abdomen and undersurface of the diaphragm. The lobes are usually identified as the *right* (largest); the *left* (somewhat smaller and wedge shaped; the *quadrate* (roughly square-shaped); and the *caudate* (tail-like configuration). Principal diseases and disorders with liver involvement are cirrhosis, hepatitis, and jaundice. Primary hepatic carcinomas are also seen, but less frequently. (The word *hepatic* indicates a condition of, or affecting the liver.) The liver is subject to adverse effects caused by a number of substances, including certain antimicrobials, such as isoniazid, rifampin, and pyrazinamide.

The liver secretes bile, which is discharged into the intestine: absorbs from the blood the products of carbohydrate digestion and stores them as glycogen; acts on nitrogenous wastes and returns them to the blood in the form of urea and related compounds; and destroys "worn-out" red corpuscles. The liver also produces fibrinogen and prothrombin. The bile discharged through the intestine plays an important role in the digestion of fat and carries with it some of the more complex waste products of the body. See **Bile.** In structure, the liver is very complex. It develops as a hollow outgrowth of the embryonic gut just behind the stomach which forks to produce the gallbladder and the liver. The connection with the gut persists as the common bile duct. In the adult, the liver cells are arranged in cords, separated by blood channels with incomplete lining known as sinusoids. Within the cords, minute bile capillaries between the cells converge to larger and larger ducts, which ultimately form the main hepatic duct. The gland receives blood from an arterial supply and also from the portal vein. The latter drains blood from the capillaries of the intestine and breaks up into sinusoids in the liver. These small passages are drained by the hepatic vein. The formation of stones in the biliary tract is described in **Gallbladder and Biliary Tract Diseases.**

A more detailed description of the physiology and biochemistry of the liver is given toward the end of this entry.

Familial Hypercholesterolemia. This disease has been treated by way of liver transplantation. It is a genetic disease caused by mutations in the gene encoding the low-density lipoprotein (LDL) receptor. This cell-surface receptor normally removes cholesterol-carrying LDL from the circulation. Persons with two mutant LDL-receptor genes produce few or no LDL receptors and, therefore, remove LDL from plasma at a reduced rate. As a result, LDL accumulates in plasma to levels up to 8 times normal. Patients almost always have severe atherosclerosis in childhood, with death from myocardial infarction often occurring before age 20 years. A single mutant LDL-receptor gene is inherited and occurs at a frequency of 1 in 500 in the general population: affected persons accumulate twice the normal level of LDL and symptomatic atherosclerosis usually occurs in the fourth and fifth decades of life. Typically, each heterozygote for familial hypercholesterolemia will transmit the mutant gene to half of the children, who then become heterozygotes. When two heterozygotes marry (estimated to be 1 in 250,000 marriages), one-fourth of the offspring will inherit a copy of the mutant LDL-receptor gene from both parents, and these offspring will be homozygotes.

Until recently, traditional approaches to the disease have not been effective. Inasmuch as the liver manufactures large numbers of LDL receptors because the organ requires a large amount of cholesterol for secretion into the bile, for conversion to bile acids, and for the production of lipoproteins, some authorities reason that transplantation of a normal liver with its normal receptors should theoretically lower LDL levels profoundly in homozygotes. In early patients so treated, this has proved to be a correct assumption.

In mid-1990, Reinher and associated researchers (Karolinska Institute, Stockholm, Sweden) reported on the use of an inhibitor of cholesterol biosynthesis in the treatment of hepatic metabolism disorders. The cholesterol production rate-limiting enzyme (pravastatin) is 3-hydroxy-3-methyl glutaryl coenzyme A (HMG-CoA) reductase. In a study of ten patients over a period of three weeks, the group found that pravastatin therapy reduced total plasma cholesterol by 26 percent and LDL cholesterol by 39 percent. The report concludes: "Inhibition of hepatic HMG-CoA reductase by pravastatin results in an increased expression of hepatic LDL receptors, which explains the lowered plasma levels of LDL cholesterol."

Diseases of the Liver

Cirrhosis. In the Western Hemisphere, *chronic* diseases of the liver are comparatively infrequent—with exception of cirrhosis of the liver, which occurs frequently in Europe and the United States. In cirrhosis, the hepatic parenchyma (functioning tissue, as contrasted with connective tissue) is progressively destroyed and replaced by collagen (gelatinous substance found in connective tissue and bone). During surgery or autopsy, the organ will exhibit bands of collagen extending between the lobes and connecting portal areas. This process, if left untreated, ultimately grossly disorganizes the liver and leads to cessation of the organ's metabolic functions.

Alcoholic cirrhosis is the most common type of cirrhosis seen in the United States. Infrequently, the liver will shrink in this disease, but more frequently the organ will enlarge and may weigh two kilograms or more. Ingestion of large quantities of alcohol over a period of years[1] is the primary cause of alcoholic cirrhosis. It is no longer generally believed that poor nutrition, which often accompanies heavy alcohol consumption, is a primary cause of the disease, although it may be a secondary contributing factor to degeneration of the health of the individual.

In treatment, the logical first step for the patient is to stop drinking alcohol. Statistics indicate that the 5-year survival for patients who continue to drink is less than 50 percent. This may reach 80% in cases where the patient maintains abstinence. Treatment is directed toward maintaining good nutrition and preventing serious complications. Total fluid intake is supervised to effect an optimum fluid and electrolyte balance. Diuretics may be used, particularly where massive ascites (accumulation of serous fluid in abdominal cavity) are present. Vitamin K therapy is sometimes used. Protein intake may be restricted where hepatic encephalopathy may be suspected.

Cirrhosis increases the risk of gallstones as the result of elevated bilirubin. See **Bile.** The risk of peptic ulcer is also increased twofold in cirrhosis. Also, in well-established cirrhosis, the *hepatorenal syndrome* may be seen. Simply defined, this is functional renal (kidney) failure. Mortality can range from 60 to 90 percent. Treatment is directed toward eliminating exogenous sources of ammonia, usually accomplished by restricting dietary protein. Means are also taken to control gastrointestinal bleeding. Where alcoholic liver disease is well established, some 10% of patients may develop *spontaneous bacterial peritonitis*. The exact mechanism underlying this condition is not fully understood. *Bleeding esophageal varices* are a serious complication of alcoholic cirrhosis.

The veins of the portal venous system transport all blood from the abdominal gastrointestinal tract, spleen, pancreas, and gallbladder, returning it to the heart by way of the liver. Portal hypertension in patients with cirrhosis causes gastrointestinal bleeding and esophageal varices. T. Poynard and a team of researchers (Franco-Italian Multicenter Study Group) conducted a study to determine the effectiveness of beta-andrenergic-antagonist drugs in the prevention of gastrointestinal bleeding in patients with cirrhosis and esophageal varices. Prior studies had not been conclusive. Generally, it had been reported that the continuous administration of beta-adrenergic-antagonistic drugs had induced a sustained decrease in portal pressure in patients with cirrhosis. Conclusion of the Poynard team findings (March 1991): "Propranolol and nadolol are effective in preventing first bleeding and reducing the mortality rate associated with gastrointestinal bleeding in patients with cirrhosis regardless of severity."

As reported by Massimo Colombo, et al. "Hepatocellular carcinoma is a highly malignant tumor with an extremely poor prognosis and an estimated incidence of about 1 million cases per year worldwide. Patients with cirrhosis of the liver have been identified as being at risk for this carcinoma." The causation of this particular type of tumor in association with cirrhosis is poorly understood. A study group at the University of Milan has concluded: "In the West, as in Asia, patients with cirrhosis of the liver are at substantial risk for hepatocellular carcinoma, with a yearly incidence rate of 3 percent. Our screening program did not appreciably increase the rate of detection of potentially curable tumors."

Effective treatment is difficult and bleeding from this source may be fatal in 70% of cases.

Primary biliary cirrhosis, relatively uncommon, is typically seen in women during the fourth to sixth decade of life. Symptoms in the early phase include a generalized itching of the skin (pruritus) and minor, continuing fatigue. Often the condition persists for a long period before a physician is consulted. In this disease, there is a significant drop in biliary secretion, causing a marked rise in serum cholesterol level. Xanthomas (small, yellow neoplastic growths) may occur about the eyes and tuberous xanthomas may be seen over the extremities. Bone pain, resulting from chronic malabsorption of fat-soluble vitamins A, D, E, and K, may be apparent. Diagnosis can be difficult because of similarities of primary biliary cirrhosis with subclinical cholangitis or other biliary tract diseases. See **Gallbladder and Biliary Tract Diseases.**

Although the disease progresses slowly, the long-term prognosis is usually not good (10–20 years). Since there is no effective and specific therapy for the condition, treatment is directed toward alleviating symptoms.

Hemochromatosis. This is an infrequent liver disease in which inordinately large quantities of iron are deposited in the parenchymal cells

[1]Ten percent of heavy drinkers (1 pint of whiskey daily for a number of years) run a high risk of developing cirrhosis.

of the organ. Eventually, these deposits destroy and scar the liver as in cirrhosis. Males have this disease at ten times the rate of females. Although onset may be earlier, symptoms usually develop during the fourth and fifth decades of life. Because of malfunctions in processes which govern iron absorption, the iron deposits not only in the liver, but also in the skin, pancreas, and heart muscle. The symptoms include a brown cutaneous pigmentation (caused by melanin) and grayish appearance (due to iron). The liver may be enlarged. There may be weight loss, decrease in body hair, and weakness. Other symptoms parallel those of diabetes mellitus, congestive heart failure or arrhythmias, and stiffness in the joints. Therapy at one time was directed toward ameliorating the aforementioned symptoms (related disorders), but comparatively recently it has been found that removal of a point (0.47 liter) of blood at regular intervals (*phlebotomy*), depending upon the patient's specific condition, is effective in depleting the iron stores. General improvement occurs if arthropathy or pituitary insufficiency are not present.

Wilson's Disease. Related to excessive deposits of copper, this is a liver-related disease and discussed in the entry on **Wilson's Disease.**

Hepatitis. An inflammatory and necrotic disease of liver cells. Commonly, the disease will be virus-induced, although hundreds of drugs are also known to cause hepatitis. It is difficult to determine the source. Among the known viruses that produce acute hepatitis are: (a) Hepatitis type A, once called infectious or short-incubation hepatitis virus; (b) hepatitis type B, serum or long-incubation hepatitis virus; (3) the non-A, non-B hepatitis viruses; (4) Epstein-Barr virus, which also causes infectious mononucleosis; and (5) cytomegalovirus. See also **Virus.** Hepatitis continues to appear after blood transfusions—in about 30,000 cases per year in the United States even though research has centered on preventing these occurrences. It should be pointed out that, although excellent tests are available for hepatitis B, unfortunately most post-transfusion cases (up to 90%) develop as the result of the presence of non-A, non-B viruses.

There is some evidence that type A and type B hepatitis may be decreasing in the United States. The fatality rate of type A hepatitis is low, probably not exceeding 0.2 percent. The rate is higher in type B hepatitis, ranging from 0.3 to 15 percent. It is estimated that nearly 45% of the population has an immunity to type A infections; 5–10% of the population for type B infections. Immunity for non-A, non-B infections is unknown. Type A virus is transmitted by the fecal-oral route. Sewage-contaminated shellfish are sometimes implicated. The transmission of type B virus may be percutaneous (penetration through skin), oral-oral, or venereal. With non-A, non-B viruses, the route is percutaneous. The incubation period varies with each type of virus: Type A hepatitis, 20–37 days, but in extremes may range from 15 to 49 days; type B hepatitis, 60–110 days, but in extremes, from 25 to 160 days; non-A, non-B viruses, 37–70 days, but in extremes, from 21 to 84 days.

The course of hepatitis infections also varies with the causative agent. Type A hepatitis does not progress to chronic liver disease, whereas about 10% of cases of Type B infections will lead to chronic liver disease. The risk of chronic liver disease with non-A, non-B virus infections is higher, ranging from 10–40% of cases. In situations where exposure to the virus is known, but disease has not developed, the administration of pooled gamma globulin is effective in the cases involving type A and non-A, non-B viruses, but is not effective in type B cases. Where there has been exposure to type B virus, the use of specific hepatitis B immune globulin is effective.

It is not surprising that the onset and course of acute viral hepatitis vary considerably because of the several possible causative factors. Onset may be sudden or gradual. In general, all or some of the early symptoms will include combinations of fatigue, lassitude, drowsiness, loss of appetite, nausea and, most specific to the disease, dark urine. Headache and very mild fever, in the absence of chills, may be present. There is usually mild and generalized abdominal discomfort. Movement tends to aggravate this discomfort. Itching of the skin may occur. Mild arthritis may develop, although this symptom is usually limited to type B virus infections. As the disease progresses, jaundice will be evident. See **Jaundice.**

Particularly in older people with type B infections, recovery may be quite long—several months to a year—with recurring intermittent symptoms (relapse). A relapse is milder and of shorter duration than the initial attack. Rarely, the course of the disease will be rampant, leading to coma and even death. Fatal complications of hepatitis may include aplastic anemia, hemolytic anemia, hypoglycemia, and polyarteritis. Some people do not recover completely from type A and non-A, non-B viruses and develop *chronic hepatitis*.

Many authorities agree that treatment seldom alters the course of acute viral hepatitis. Sensible suggestions are made to the patient—bed rest when there is excessive fatigue and serious discomfort which may be present in the initial phase. Controlled studies have shown that vigorous physical exercise during the recovery phases does not increase the risk of relapse or chronic disease. It has been reported that a high-calorie diet (3000 + calories/day) for a few days at the appropriate time in the recovery stage may shorten the duration of the disease by a few days. Low-fat diets have not been shown to alter the course of the disease. During periods of nausea and vomiting, hospitalization may be required so that intravenous fluid and electrolyte replacement can be effected. Although alcohol has not been shown to aggravate the disease, abstinence is usually suggested in the interest of limiting any additional load on the hepatic and related systems.

In instances of severe acute viral hepatitis where the patient becomes encephalopathic, corticosteroids (not proven effective), hyperimmune gamma globulin, keto analogues of essential amino acids, and exchange transfusions, among other drugs and procedures, have been used. However, some authorities currently feel that acute encephalopathy responds little, if any, to treatment.

Chronic hepatitis may be *chronic active* or *chronic persistent*. Diagnosis is important because the therapy differs for the two conditions. Frequently, a percutaneous liver biopsy will be required. In chronic active hepatitis, the disease is variably progressive and eventually causes cirrhosis and hepatic failure. On the other hand, chronic persistent hepatitis does not progress. Persons with untreated chronic active hepatitis have a 5-year survival expectancy of less than 50%—possibly up to 90–95% where corticosteroid therapy is effective. This therapy in responsive patients brings about improvement of liver function within several months. In contrast, chronic persistent hepatitis is benign and usually patients lead a normal, active life, even though serum aminotransferase abnormalities may continue for many years.

Drug-induced hepatitis is often difficult to differentiate from the disease caused by a virus. Although uncommon, in one case in 9,000–10,000 patients the administration of halothane (see **Anesthesia**) will produce hepatic necrosis. This occurrence is most common in overweight females or after a second exposure to the drug, and is frequently fatal. Drugs which also are hepatitis related include isoniazid, methyldopa, phenytoin, and the sulfonamides. Although such side effects are uncommon, the physician will be on guard for signs of hepatitis and liver damage in deciding on starting or continuing therapy with these drugs. There is some evidence that cysteamine or acetylcysteine may reverse the actions of these drugs if noted promptly. Poison derived from the wild mushroom *Amanita phalloides* is capable of producing overwhelming hepatic necrosis. See **Foodborne Diseases.** The adverse effects of certain antimicrobial agents were mentioned earlier in this entry.

Reye's Syndrome. This is an often fatal systemic disorder that follows viral infection in children. A number of cases of what appears to be Reye's syndrome have also been described in recent years in young and middle-aged adults. Present in the syndrome are encephalopathy and fatty liver. The syndrome develops suddenly, with onset of intractable vomiting occuring a few days after the viral illness. Sensorial impairment appears and soon afterward may progress to coma. Seizures also may occur. The liver is usually found to be enlarged. Specific therapy is not available. Supportive measures include lactulose to control hyper-ammonemia; fresh frozen plasma to replenish clotting factors; mannitol or dexamethasone to lower increased intercranial pressure; and mechanical ventilation. A case fatality rate of 23% has been reported. Epidemiologic evidence strongly links Reye's syndrome with outbreaks of viral disease, especially influenza. Although the mechanism is unknown most physicians recommend that salicylates (aspirin and aspirin-containing medicants) not be given to children with chicken pox or influenza.

Physiology and Biochemistry of the Liver

The blood returning from the intestine to the heart is shunted through a capillary system, the hepatic sinusoids, which are surrounded by epithelial cells arranged in plate forms. These plates cross each other in space at different angles, to permit the greatest possible contact between the blood and these polygonal epithelial cells. The resulting spongelike organ located under the diaphragm and covered by the connective tissue capsule of Glisson is the largest organ of the body. Under normal circumstances, the major part of its blood, between 66 and 75%, comes from the portal vein which drains the splanchnic capillaries, particularly those of intestine, pancreas and spleen. Approximately one quarter to one-third of the hepatic blood comes from the hepatic artery originating from the aorta at the celiac axis. Both hepatic artery and portal vein enter at the hilus of the liver and divide in a dichotomic fashion into parallel running branches. They are surrounded by ramified extensions of Glisson's capsule. The hepatic artery sends branches to the capillary plexus of the portal tracts whereas the bulk of its blood is released into the sinusoids parenchyma as does the portal vein which forms by confluence of superior mesenteric, inferior mesenteric and splenic vein, and receives additional internal radicles from the portal capillary plexus. The sinusoids are blood spaces, normally without the basement membrane otherwise seen in capillaries; they are, therefore, characterized by great permeability for serum protein. Moreover, some of their lining endothelial cells, those which are star-shaped and called Kupffer cells, are part of the reticuloendothelial system. The lining cells form the sinusoidal wall and leave small stomata open through which macromolecular substances pass into a tissue space between liver cell plates and sinusoids and extending between neighboring hepatocytes almost to the bile canaliculus. Tissue fluid is drained toward the lymphatics in either the central canal or, in the human, mainly the portal tract. Arterial and venous blood, mixed to a varying degree, flows toward the tributaries of the hepatic veins which combine to larger veins into which frequently small branches enter at almost right angles. The largest branches enter into the vena cava inferior behind the liver. Vascular sphincter mechanisms in various locations regulate hepatic blood flow and thus function. The portal tracts and the central canals around the tributaries of the hepatic veins cross each other in space and are throughout the liver about 0.3 mm apart. The direction of the blood flow from the portal tracts to the central canals produces the concentric arrangement of the liver cell plates characterizing the liver lobule which conventionally is considered the structural unit of the liver.

The liver forms bile which is released into slits between the liver cells, the bile canaliculi, which are arranged in a chicken-wire-like fashion; the wall of the canaliculi is formed by part of the hepatocellular plasma membrane. The bile canaliculi are drained by small tubes with an independent cuboidal epithelial lining, the ductules or cholangioles. Under normal circumstances hardly any are found within the lobule, the majority being in the periportal zone or in the portal tract. Under abnormal circumstances, they increase in number and are then found deep within the lobule. The ductules continue into the bile ducts located in the portal tracts which unite in dichotomic fashion to finally form the common hepatic duct which leaves the liver where hepatic artery and portal vein enter it. It combines with the cystic duct draining the gallbladder, which concentrates bile by water reabsorption to form the common duct running toward the duodenum. This entrance is controlled by the choledochoduodenal sphincter of Oddi. Bile is produced at an almost constant rate but released from the biliary system in human beings and many animals only if food appears in the duodenum. As a result of this or other mechanisms, the sphincter of Oddi relaxes and the gallbladder contracts. This leads to proper utilization of bile which, while being partly an excretory product, is a secretion essential in intestinal digestion and absorption.

In the liver, several fluid currents exist. Blood and some tissue fluid flow toward the central canal, while bile and most of the tissue fluid (at least in the human) are flowing toward the portal canal. The normal liver consists of approximately 60% hexagonal epithelial cells (hepatocytes), 30% littoral endothelial or Kupffer cells, and about 2% each of bile duct cells, connective tissue and blood vessels. The hepatocytes have three types of borders. Where they are in contact with each other, the border is straight indicating limited, if any, exchange of substance between individual cells. The border toward the tissue space is elongated by narrow extensions of the space between neighboring hepatocytes and particularly by the formation of irregularly shaped finger-like projections in the form of microvilli. This tremendous elongation of the border of the hepatocytes and the preferential location of enzymes in this location reflects structurally the extensive exchange of substances between hepatocytes, tissue space and blood. Much shorter is the border toward the bile canaliculus which is also thrown into microvilli which are far more regular and disappear upon impairment of biliary secretion. Preferential accumulation of ATPase in the villi indicates the intensity of the metabolic processes in bile secretion.

The nucleus is normally vesicular and has conspicuous nucleoli. It varies considerably under normal and pathologic conditions, the majority being tetraploid in adult rodents. Binucleated cells increase in regeneration. The cytoplasm normally contains many and relatively large mitochondria in the matrix, of which the citric acid cycle enzymes and, in the double membrane, the electron transfer enzymes can be demonstrated. Ribosomes as ribonuclear protein are arranged around messenger RNA usually in helix form as polysomes. These polysomes as the site of protein biosynthesis may be either free in the cytoplasm or attached to the extensive endoplasmic reticulum which thus becomes granular and the site of secretion of protein such as serum proteins. The endoplasmic reticulum is also the site of steroid synthesis, and the smooth endoplasmic reticulum is the site of detoxification and of glucose-6-phosphatase. In addition, one notes the Golgi apparatus responsible for secretion and the perinuclear dark bodies, the lysosomes, which are the site of various hydrolytic enzymes mainly with peak activity in acid medium. They serve to segregate intracellular material after pinocytosis or for storage, secretion and separation of organelles undergoing destruction in the form of autophagic vacuoles. The soluble fraction of the cytoplasm, the hyaloplasm, corresponding to the supernatant fluid in cytochemical analysis, contains proteins and enzymes and cofactors related to carbohydrate metabolism and activation of amino acids and nucleic acids. In addition, in the normal liver, glycogen and few fat droplets are found as well as some ferritin crystals which, under abnormal circumstances, become hemosiderin deposits giving histochemical iron reaction.

The main functions of the liver cells are: (1) secretion of substances into the bloodstream of which the serum proteins particularly albumins, alpha-globulin, the proteins concerned with blood clotting, haptoglobin and transferrin, as well as some blood enzymes (*e.g.*, esterase), serum cholesterol, and blood glucose are the most important; in contrast to all other tissues which utilize but do not form blood glucose, the liver cells are the main source of the blood glucose because of a specific phosphatase system; (2) storage of various metabolites particularly glycogen, proteins, fat and vitamins; (3) transformation of various compounds into each other, *e.g.*, fats into carbohydrates and vice versa; (4) detoxification mainly by oxidation or conjugation, the latter mainly for better solubility and urinary excretion; (5) formation of the bile into which bile pigment is transmitted by conjugation and bile acids and cholesterol by transformation.

A variety of sinusoidal cells are seen. Some are flat endothelial cells, others are Kupffer cells with a cytoplasm of varying and irregular outlines and ameboid extensions. They contain few mitochondria but varying inclusions. They are engaged in phagocytosis of circulating exogenous and endogenous macromolecular or corpuscular elements, including bacteria, as well as of hepatocellular breakdown products, they are active in transformation of blood pigment to bile pigment. Other sinusoidal lining cells, rare under normal circumstances, form serum gamma-globulin and correspond to plasma cells. Also fibroblasts can be seen around the sinusoids.

The liver, as a whole, because of its strategic situation near the right heart and because of its sheer bulk, influences circulating blood volume, as well as electrolyte and water metabolism.

Liver Transplantations

It was just about a decade ago when a liver transplantation was covered during the prime time news. By the early 1990s, liver transplantation had become an established therapy, with such transplantations numbering in the thousands. Because of a very serious shortage of livers for donation, the waiting list numbers in the thousands. Liver transplantation ethics has become a major topic of discussion among medical and health professionals. A simple question typifies the current dilemma: "Who should receive a liver for transplantation? A young mother, whose prospects for surviving with a transplant is only 20 percent, or a 65-year-old alcoholic, whose chance of surviving is 80 percent if the patient stops drinking? Some medical professionals have observed that, if the criteria for selecting patients becomes too narrow, this could have a dampening effect on liver transplantation technology.

A few experts believe that alternate non-transplantation therapies ultimately will develop, thus obsoleting liver transplantation procedures in the long run. However, for the immediate decade, simple optimism for the future does not suffice.

When it was established that a human can regenerate a partly removed liver, some researchers proposed that pieces of liver tissue may be regenerated in the laboratory, thus maximizing the effectiveness of available whole organs from donors, who have been in short supply since the beginnings of liver transplant technology. Research along these lines commenced in the chemical and biochemical engineering laboratories at Massachusetts Institute of Technology. In the early phase of the project, liver cells were mounted on polymer mesh and treated with enzymes. The aim of the program is to grow the cells to about 10% of the mass of a whole liver. This tissue then would be transplanted into the patient and would, within several weeks, become fully grown and replace the original liver. Research is continuing.

In a large series of childhood liver transplantations, nearly one-third have been performed because of metabolic disorders. Liver transplantation now is a well-accepted treatment for inherited metabolic disorders. Glycogen storage (Type IV) disease can be reversed by successfully transplanting a liver with normal amounts of branching enzyme. Inherited abnormalities of glycogen metabolism have been recognized for many years, mainly as the result of pathologists identifying gross accumulations of glycogen in tissues during postmortem examination. As pointed out by R. R. Howell (University of Miami), "The specific patterns of tissue involvement permitted recognition of a number of clinical types of glycogen storage disease long before the biochemical pathways had been identified and the specific causes of the inborn errors of metabolism understood." Some researchers had postulated that, after liver transplantation, progressive and probably fatal myopathy, cardiomyopathy, or encephalopathy would develop, since the enzyme defect is present in all the affected tissues, and that the abnormal glycogen would continue to accumulate in them. Surprising and fortuitous are findings that the predicted and ultimately fatal conditions have not occurred. A satisfactory explanation remains to be developed.

R. W. Strong and associates (Royal Children's Hospital, Brisbane, Australia) have reported that orthotopic liver transplantation is an effective therapy for end-stage liver disease and has proved to be a major advance in treating liver disease in children. The foremost obstacle faced by transplantation units worldwide has been the relative paucity of infant and child donors. The principle of transplanting a portion of the liver from an adult into a child has been accepted in many centers. Ethical issues must be considered when contemplating liver transplantation from parent to child. These issues are similar to those associated with the transplantation of renal grafts from living, related donors. Experience in many centers has shown that the risk to the donor is considered minimal. The risk to the recipient is considered no greater than that with the transplantation of a reduced-size graft from a cadaver donor. In the opinion of many specialists, living-donor liver transplantations are not justified when there are sufficient numbers of cadaver donors. In special instances, however, the procedure can be justified, as, for example, with a patient with fulminant hepatic failure when no cadaver donor is available and when the recipient has a reasonable chance of a successful outcome.

Cyclosporine is the basis of the immunosuppressive regimen in most patients undergoing orthotopic liver transplantation. The drug is a cyclic polypeptide that is produced by two species of fungi and is essentially insoluble in water. Its oral form consists of a solution of 100 mg of cyclosporine per milliliter of olive oil containing 12.5% ethanol by volume. Cyclosporine is absorbed much as fat and other fat-soluble substances are. Children, particularly infants, require much higher doses of orally administered cyclosporine than adults after liver transplantation. Bowel length appears to be the main determinant of the large difference between the adult and infant populations. These factors are reported by Whitington and associates (University of Chicago).

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LIZARDS (*Reptilia, Sauria*). Animals closely related to the snakes but having eyelids and the ventral surface of the body as well as the upper covered with small scales. They are usually elongate, with short legs and a long tail. Most lizards are small, though some species attain a length of several feet.

The classification and nomenclature of these animals is confused. The lizards are sometimes grouped with the snakes but some authorities regard them as a separate order of *Reptilia*. See also **Fossil Reptiles.**

There are many species of lizards but with the exception of the poisonous Gila monster and the edible iguanas they are of no economic importance. A few of the smaller species are eaten to a limited extent.

A table, Classification of Lizards, is included at the end of this article. Containing over 3000 species, the order *Squamata* (scaly reptiles) embraces about one-half of all modern reptiles. As shown by the table, there are several infraorders, families, and subfamilies, some of which are described in separate articles in this encyclopedia. Thus, please refer to **Agamids** (the agamas and chameleons); **Geckos** (geckos, snake lizards, and diabinids); **Iguanids** (iguanas, basalisks, and anoles); and **Skinks.**

This immediate article describes the remaining principal lizards, including girdled lizards, whiptails, so-called true lizards, shovel-snouted legless lizards, beaded lizards, monitors, and earless and ringed lizards, among others.

In many families of lizards, there are special "break points" in the bodies of the tail vertebrae; the tail beyond these can be discarded in various emergencies. Although the tail may be regenerated later, the new tail is usually shorter than the old and is supported not by vertebrae, but by a central rod of cartilage. Further, according to K. Klemmer, the pattern of scales and coloration of the new tail frequently differ from the discarded tail.

Girdle-Tailed Lizards. Of the family *Cordylidae,* these lizards are essentially limited to sub-Saharan Africa. According to D. G. Broadley, these lizards are well armored. The legs are reduced. The body is covered with longitudinal and transverse rows of rectangular scales, each having a bony element beneath it. The tongue is simple, with only a slight notch and covered with papillae. Because these animals have adapted to a dry habitat, they are not sensitive to reasonable extremes of temperature. Thus, most of these species are well adapted to terrariums. They can also adapt to a wide variety of easily available foods. Principal species include the club-tailed lizard (*Cordylus*); the yellow-brown sungazer (*C. giganteus*), which is about 38 centimeters long and lives in South Africa (Fig. 1); the armadillo lizard (*C. cataphractus*), a rather slow, heavily armored animal, the nostrils of which are coinspicuously elongated; and the common cape girdled lizard, a rather flat animal about 20 centimeters long, with a somewhat spiny tail and black through dark-brown coloration, and that lives in the Cape Province of South Africa; and the blue-spotted girdled lizard (*C. caeruleopunctatus*), which is about 18 centimeters long and found in eastern Africa. The aforementioned lizards feed on insects and small animals.

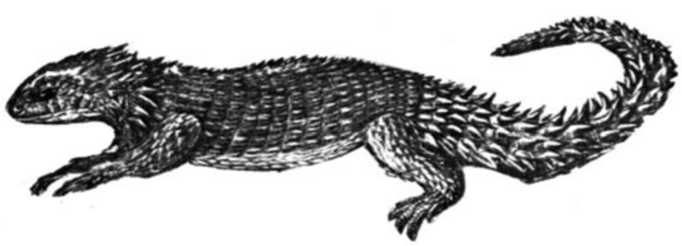


Fig. 1. Yellow-brown sungazer (*Cordylus giganteus*). Length is approximately 38 centimeters.

Of the genus *Pseudocordylus* (false club tails), the leathery crag lizard is representative. This animal (*P. microlepidotus*) is about 32 centimeters long, with a rather broad head. Dorsal scales are small and rounded. The upper side of the body is dark, usually with a dark upper side and frequently with pale bands or transverse patterns. The flanks are yellow-orange, and the belly is light. The animal inhabits the coastal mountains of Cape Province, South Africa. The habitat is characterized by shale, in deep crevices of which the lizard can hide. The diet is one of insects, other small animals, and lichens and other plants.

Of the genus *Platysaurus* (flat lizards), there are several species, including Wilhelm's red-tailed lizard and the Transvaal red-tailed lizard. The latter is the smallest of the genus and has a transparent window in the lower eyelid. In two subspecies, the males are a glowing green, with a red tail. Other species are red-brown. They live in the Soutpansberg and Drakensberg Mountains of South Africa. D. G. Broadley observes that the method of reproduction in the *Platysaurus* diverges from that of the *Cordylus* relatives. The females lay two elongated eggs in a crack in the rock during mid-winter. When the sun has warmed the rock surface in early morning, the *Platysauarus* emerges from an overnight shelter. Most of the daylight hours are spent under the sun. However, the mid-day heat drives the animals into shade. They eat primarily locusts and beetles. Males define their territories and display their colors to competitors by raising up. When sought, these lizards rush to narrow crevices and expand their bodies, thus making it virtually impossible to retrieve one alive.

Of the genus *Chamaesaura* (snake lizards), these animals are quite snakelike in appearance. They have pointed heads, slender bodies, and gradually tapering tails, which can be three times longer than the head and trunk together. The Cape snake lizard ranges up to 63 centimeters in length. Snake lizards prefer the grassy mountain slopes and high plateaus in southern and central Africa, where they feed mostly on grasshoppers. They move in grass with essentially the same ability as snakes. Their tiny limbs are seldom used. Generally, they are of a brown coloration.

The subfamily Gerrhosaurinae (plated or rock lizards) is distributed across Africa and is also found in Madagascar. They have well-developed bony plates under large horny scales. When well gorged with food, a fold makes it possible to expand the width of the body. This provision is also utilized when the female is carrying eggs. These animals are oviparous. The tail is long and can be autotomized. The better understood members are the African plated lizard (*G. major*), which is cylindrical in shape and grows to approximately 56 centimeters in length. The animals frequent southeastern Africa. Smith's plated lizard (*G. validus*) is the largest species in the genus. Its diet consists of insects, spiders, millipedes, or scorpions and smaller lizards. The female usually lays four eggs with leathery shells, most frequencly locating them in the cracks of rock.

Night Lizards (Xantusiidae). These lizards range in length from 12 to 15 centimeters. The animals have four normally developed limbs, each with five toes. The scales are small. The ventral shields are large and rectangular. The pupil is vertical. In place of movable eyelids, there is a transparent "spectacle" something like that found in the geckos. Also, the vertebrae, skull, tongue, and eyes resemble those of the geckos. These animals feed on insects and spiders. The common night lizard also consumes vegetation. Their range extends from the southwestern United States and the offshore islands as far as Panama. One species is found in Cuba. Their hiding places are cracks in rocks, under roots, and in the bark of trees. All are viviparous, bearing from one to nine young at a time. There are four genera, but only a total of twelve species.

The Cuban night lizard (*Cricosaura typica*) was not discovered until 1863 (by German zoologist Gundlach). The large-headed or common night lizard occurs on the rocky islands of San Clemente, San Nicholas, and Santa Barbara off the coast of California. They consume both seeds and flowers. This particular animal is not strictly nocturnal.

Of the three species of *Xantusia*, the granite night lizard (*X. henshawi*) is native to California. Total length ranges up to about 17 centimeters. It is found not only in California, but also in the southwestern United States. Herpetologists Schmidt and Inger mention: "One would not have thought a crowbar a suitable instrument for catching lizards this small, but in certain parts of California it is necessary." The animal is quite flat, allowing it to creep into very small crevices, often to avoid sunshine. The desert night lizard (*X. vigilis*) prefers soft yucca plants for its diet. This species has been studied in detail on the southwestern edge of the Mojave Desert in California. They are frequently found at the foot of yuccas. In 1950, it was discovered that the formation of a placenta, rare among lizards, occurred in the Xantusiidae. After a gestation period of 90 days, from one to three young are born.

Whiptails (Teiidae). These animals vary widely in length, from 7.5 to 140 centimeters. They are someties referred to as the New World counterpart of the true lizards (Lacertidae). Some species, such as *Ameiva*, have well-developed limbs and scales and are formed so as to appear very much like European lizards. In the Teiidae family, there are some 45 genera and about 200 species. With the exception of northern and northeastern states, the teiids range widely throughout the United States and in Central America, the West Indies, and parts of Argentina and Chile. Their wide distribution accounts for so many species that have evolved through adaptation to numerous living and survival conditions, including arboreal, aquatic, and ground forms in rain forests, plains, deserts, seacoasts and inland regions, lowlands, and high mountains, such as the Andes, to the tropical forests along the Amazon river.

Generalizing, the teiids have lizardlike tails, ranging from cylindrical in form to a laterally flattened form found in aquatic environments. Their scales range widely—rounded, elliptical, elongated, hexagonal, smooth, or keeled. Depending upon particular species, a variety of bands may run longitudinally, transversely, or diagonally. Often, the teiid body is covered with large scales. Often, the tongue can be extended considerably and possesses a deep notch at the end. Depending upon location, the teiid ranges from herbivorous to insectivorous. Mode of reproduction has not been observed in detail for many of the species.

Racerunner (genus *Cnemidophorus*) are whip-tail lizards, the most commonly encountered teiid in the United States, and are also found in southern regions extending to northern Argentina. The term *racerunner* stems from the fact that these animals run fast, stop suddenly to scan for their enemy, and continue immediately in a series of rapid dashes, frequently in a different direction. It appears difficult for these animals to resist almost constant motion. Members of *Cnemidophorus* include the six-lined racerunner (*C. sexlineatus*), which is usually close to 8 centimeters; the checkered whiptail with a pale blue body; the seven-lined racerunner (*C. deppei*), which can achieve a length of 24 centimeters, is found from Mexico to Costa Rica, and has a light blue throat, a turquoise belly, with blue spots on the sides and a brick red lateral stripe; the spotted whiptail (*C. sackii*), which is about 25 centimeters long. Those animals with striking colorations use these features when reacting to threats from pursuers.

A striking member of the teiid family is the common tegu (*Tupinambis teguixin*), which can achieve a length of 120 to 140 centimeters. The animal is essentially black, with numerous crossbands that incorporate round yellow spots. See Fig. 2. Other animals of this type include the northern tegu or jacura (*T. nigropunctatus*), with a length of up to 120 centimeters. With their rather squat bodies, these teiids live in wooded areas, with dense undergrowth and sunny clearings and an obviously abundant food supply. The meat is prized by local natives, who fish for them with meat-baited hooks. The yellow fat of the lizard's legs is particularly prized. Medicinal properties are ascribed by local people to several parts of the common tegu. Locally terms ascribed to this lizard testify to its reputation among the local populace —"chicken wolf" and "egg thief," among others.

Fig. 2. Common tegu (*Tupinambis teguixin*), which achieves a length of about 120 to 140 centimeters and which is sometimes referred to by native peoples in Central America and northern South America as a "chicken wolf" or "egg thief."

Another teiid of the genus *Crocodilurus*, the dragon lizardet (*C. lacertinus*), achieves a length of about 50 centimeters and is found in Central America and northern South America. The animal prefers a swampy habitat. The tail is somewhat flattened and features a double keel. The animal prefers fish and frogs in its diet. The animal seeks prey by hiding underwater in holes of a stream bank or under the roots of trees. Once seizing its prey, the animal returns to its hiding place to consume its catch. As observed by Gustav Lederer (Frankfurt Zoo), the dragon lizardet lies in its pool almost all day long, with only the head above the surface of the water. Occasionally, the animal will stretch up the front part of the body and utter squeaking sounds, which are assumed to be "attracting calls." See Fig. 3.

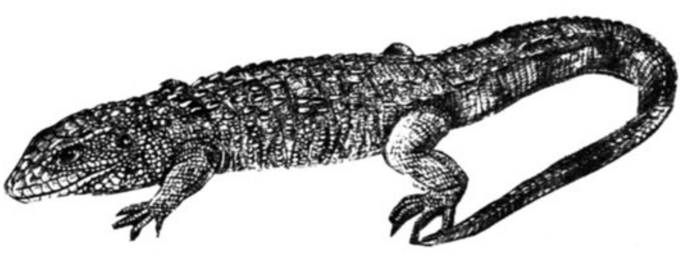


Fig. 3. Dragon lizardet (*Crocodilurus lacertinus*), achieves a length of about 50 centimeters. The animal prefers the swampy habitats of Central America and northern South America.

True Lizards (Lacertidae). These lizards are the Old World counterparts (analogues) of the Teiidae of the New World. Although these animals have not extended to Madagascar, New Guinea, and Australia, they are found elsewhere on the European and Asian continents and in most of South America as well. They range from the tiniest of Mediterranean islands to the far north of the Arctic Circle. All species of lacterids have well developed legs and a long tail. They range in length from about 12 to 90 centimeters. As observed by Klemmer, the lacertids lack many characteristics found in other lizard familes. They do not have dorsal crests or dewlaps. There are no other movable or expandable skin appendages. They have little or no ability to change their color. However, most feature an autotomous tail, part of which can be sacrificed to an enemy and later regenerated.

Many lacertids are found in dry regions. They adapt well to alternative habitats. Some species live in loose ground and elongated scales under the toes form combs at one or both sides, facilitating rapid locomotion on shifting sand. Some have a modified snout for digging. In some species there is a movable lower eyelid, often with a transparent window. Nearly all lacertids are egg layers. The species of the genus *Lacerta* are familiar to Europeans. Although subject to considerable research by noted herpetologists, the classification and relationships of species remains unclear. General characterizations include an unspecialized body structure, the band of enlarged scales around the neck, and the round or slightly compressed fingers and toes, which lack fringes or scales. There are numerous examples of the lacertids, including the sand lizard (*L. agilis*), an animal that is about 30 centimeters long; the emerald lizard (*L. viridis*), with a length up to 45 centimeters; the dwarf lizard (*L. parva*); and the largest lacertid, the jeweled or eyed lizard (*L. lepida*), which reaches a length up to 80 centimeters; among many others.

Lateral Fold Lizards (genus *Anguidae*). These animals historically are considered a rather young group that arose in the Cretaceous period. They are closely related to the monitors, which are described later. These animals generally lack limbs, but American species have four well-developed limbs. As with many other lizard families, limb degeneration in the lateral fold lizards usually can be correlated with adaption to a wide variety of native habitats. Lateral fold lizards may be shaped like a snake or like a lizard. The limbless species can achieve a length of from 50 to 100 centimeters, whereas the four-limbed species are smaller, some 20 to 40 centimeters in length. The most familiar lateral fold lizard is the slow-worm (*Anguis fragilis*), which ranges from 35 to 50 centimeters in length. This animal is not to be confused with the worm lizards (*Amphisbaenia*), to be described later.

The eastern glass lizard (genus *Ophisaurus ventralis*) is the longest lizard encountered in the United States. See Fig. 4. It can achieve a length of about 1 meter. Closely related species found in the Mississippi basin include the slender glass lizard (*O. attenautus*) and the island glass lizard (*O. compressus*), all generally confined to the southeastern United States, including Florida. Also closely related to the lateral fold lizards are the xenosaurids (Xenosauridae). There are two genera. One genus (*Shinisaurus*) embraces only one species, the Chinese crocodile lizard, which was first captured as recently as 1928 and first described in 1930 by the German herpetologist, Ernst Ahl. This lizard inhabits Kwangsi Province in southwestern China. The horny scales on the tail form a double row, as found in crocodiles. The teeth are fanglike. The animal inhabits areas near water. Chinese term the species the "lizard with great sleepiness." However, when threatened it will bite very quickly. When attacked, the lizard usually attempts to escape to water because of its diving and swimming agility.

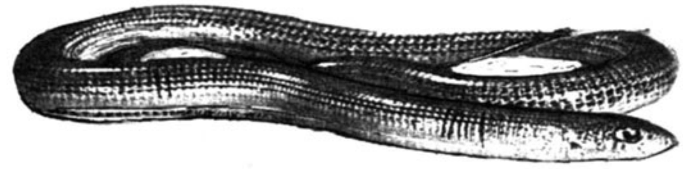


Fig. 4. Eastern glass lizard (*Ophisaurus ventralis*), the largest lizard occurring in the United States, achieves a length of 1 meter.

Beaded or Venemous Lizards (Helodermatidae). These animals are the only poisonous lizards. They are closely related to the ancestors of snakes. They are characterized by massive, un-snakelike bodies, and the presence of four fully developed limbs. The head is broad and somewhat flattened. It is joined with a short neck and an elongated, cylindrical body, ending with a thick, rounded tail. The back is covered with large, bony scales. The legs are short and powerful. Each foot has five clawed toes. The back is covered with large, bony scales. The belly bares flat, regularly arranged scales that are not fully ossified. The lower jaw teeth have single grooves on the front and back that permit the venom to flow into the wound. The two species of the beaded lizard are (1) the gila monster (*Heloderma suspectum*), which achieves a length up to 60 centimeters; and (2) the Mexican beaded lizard (*H. horridum*), which may be up to 80 centimeters long.

The gila monster (see Fig. 5) has a distinctive coloration of pink and black spots. Of two species, these animals are encountered from southern Nevada and southeastern Utah to Sonora in northwestern Mexico. Generally, these beaded snakes are active at night, but during cold weather they may appear by day. The diet consists of nestling rodents and birds and bird and reptile eggs. H. Wermuth notes that beaded lizards first move very slowly and clumsily, but as the night hours progress, they become quite agile. During long fast periods, demanded by their habitat, the animals survive on fat reserves stored up during more favorable times of the year. Fat is stored in the tail, which swells noticeably, but after a long fasting period, the tail becomes quite thin. Although seldom near water, they are good swimmers. As observed by Bogert and Martin, the gila monster can survive years of drought. In captivity, gila monsters can achieve an age of 20 years. They may be fed raw chicken eggs mixed with lean meat and supplemented with lime and vitamins.

Fig. 5. Gila monster. (*A. M. Winchester.*)

Herpetologists have reported that snake-bite antidote has no neutralizing action against gila monster venom. Typical localized consequences of a gila monster bite may include severe swelling of the victim's arm or part bitten and may be extremely painful for nearly 2 weeks. Evidence indicates that there is no damage to the victim's nervous system or vital organs, such as kidneys and liver. Experts insist that these animals belong only in the hands of very experienced keepers.

Monitors (Varanidae). As indicated by Fig. 6, the members of this genus have somewhat the appearance of legendary dragons. The Nile monitor has been known since antiquity. W. Neugebauer observes that Herodotus (who died around 424 B.C.) described the desert monitor as a "terrestrial crocodile," while the ancient Egyptians, who often depicted monitors on their monuments, knew the Nile monitor so well that they never confused it with crocodiles. Monitors occur in numerous sizes. However, they all are classified in one genus (*Varanus*). They range from 30 centimeters up to 3 meters in length. The body is usually quite massive, with four powerful legs, each bearing five clawed toes. The tail is thick and may be used as a prehensile organ or as a potent weapon. Monitors are diurnal and reach their full activity level when the sun is up and their habitat has warmed. They are good runners and climbers. Some species are arboreal; others prefer proximity to water. Such species dive and swim well. Their best weapons are their sharp teeth and dagger-sharp claws, which can inflict dangerous wounds. All monitors feed on other animals, such as insects, small lizards, and the nestlings of small mammals. The larger species seek prey, such as crustaceans, fishes, frogs, birds, rats, and snakes. It is known that the giant Komodo dragon can even take small deer and wild pigs, and it has been reported that two-banded monitors have fed on human corpses that have been interred in trees. Monitors especially like eggs in their diet and are known to eat eggs of their conspecifics. As observed by W. Neugabauer, most monitors in their native habitat suffer at the hands of humans. Their meat and eggs are eaten, and the animals are often used to produce various "medications" and amulets. The fat and oil from the paired fatty organis is used by Chinese druggists, whose buying agents travel as far as Australia. The skin of larger monitors is processed into leather.

Fig. 6. The Nile monitor (*Varanus niloticus*) is found south of the Sahara, both in deserts and on the plains and especially along rivers. The animal feeds heavily on' eggs.

Monitors are found in the tropical and subtropical parts of Africa, the Near East, southern Asia, the Indo-Australian islands, and Australia.

Earless Monitors (Lanthanotids). K. Klemmer has observed that, in 1878, the Viennese zoologist Franz Steindachner described the single animal before him as a *new* lizard from Borneo. Since it lacked external ears, Steindachner named it *Lanthonotus borneensis*, from the Greek (hidden ear). He realized that this new lizard species belonged to a unique family, which he classified as relatives of gila monsters. During the following 80 years, only six additional earless monitors were identified. Even experienced herpetologists failed to find like specimens of the animal in Borneo. It was not until 1961 that archeologists on the island of Borneo dug out and captured alive an unusual lizard. This was the first earless lizard to be captured within a 45-year period. The local Dayak tribesmen were promised a bounty for capturing live lanthanotids. By 1960, over 60 were found. These were went to zoo collections in Europe and the United States. Zoologists are much interested in earless monitors because they may be survivors of the animal group that gave rise to snakes. Many experts believe that snakes progressed from subterranean lizards in which the limbs and eyes regressed, the body became longer, and the number of vertebrae increased. the earless monitors possess these same characteristics.

Worm Lizards (Amphisbaenia). Questions persist concerning the the accurate classification of worm lizards, which combine several characteristics of snakes and only some of the characteristics that typify lizards. Some herpetologists have even questioned whether these animals are reptiles, let alone lizards. It is notable that the characteristics are not intermediate betwen snakes and lizards, but rather represent evolvement over some special path. Worm lizards have been osbserved and their full identity determined only since the early 1800s. Identification was difficult because they are subterranean and come out of their underground locations only after sunset and return just before sunrise. Their size and coloration are such that they can be easily mistaken for earthworms.

As described by Carl Gans (Department of Biology, State University of New York at Buffalo), the typical amphisbaenian body is cylindrical, bearing a loose skin with fairly defined rings. Their color may be reddish or brownish and many incorporate a pattern of dark-brown or black spots on a lighter background. Length ranges from 8 to 80 centimeters; diameter is between 1.5 and 30 millimeters. Their shovellike head structure is well adapted for digging. The eyes and ears are buried beneath the skin. Typically, worm lizards lay eggs, although a few species bear fully developed young. Worm lizards occur in several tropical areas of both the Old and New World. A white-bellied worm lizard (*Amphisbaena alba*) is shown in Fig. 7.

In lizard folklore, one finds references to unusual association between ants and amphisbaenians. For example, in Brazil, some species are referred to as "ant mothers" or "ant kings," where it is alleged that worm lizards are raised and fed by ants. On this point, Schmidt and Inger report: "Since worm lizards are often found in ant and termite

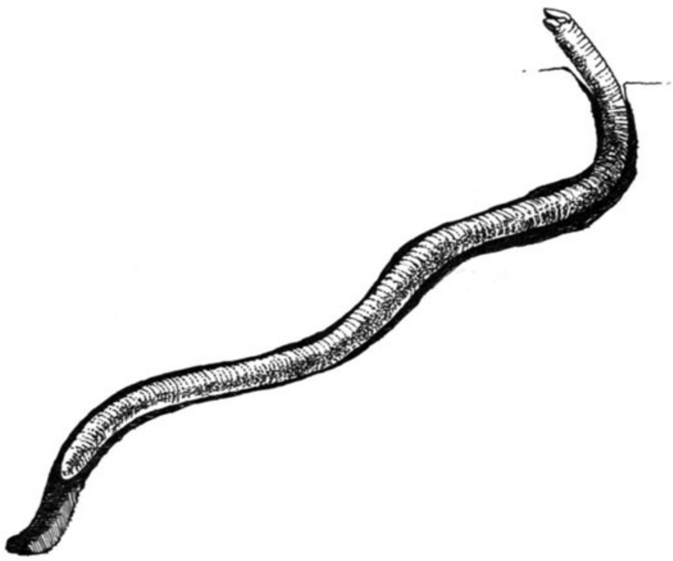


Fig. 7. A white-bellied worm lizard (*Amphisbaena alba*). Length is up to 52 centimeters.

colonies, the lizards probably have no trouble catching their prey. However, subterranean ant and termite colonies are not just food resources, but also serve as incubation chambers for egg-laying lizards. Most, if not all, worm lizard eggs discovered have been found in ant and termite colonies."

Some protection for the worm lizard is the enemy's difficulty in distinguishing the head from the tail. Both head and tail wag through the air in some species. Where Portuguese is spoken, the amphisbaenians may be called *cobras de dois cabeccas,* or two-headed snakes. Upon being grabbed by a carnivore, the outer tip of the tail may break off. The broken part twists about extensively and may momentarily divert the attention of the enemy. Unlike many other kinds of lizards, however, the broken tail does not regenerate. In the days of the early explorers, any worm lizard found was mistakingly regarded as poisonous.

By way of a series of underground tunnels, many worm lizards can hear the sounds of oncoming prey; their very long tongues can detect surrounding odors. In terrariums, amphisbaenians have been observed to crawl forward and backward with ease in their tunnels when they seek food. Also, their long tongue also assists in gathering insects for their diet. Their shovellike head allows them to enter tunnels with surprising speed. The shape of their head can be a disadvantage, however, because it requires short jaws and reduces the number of teeth. Thus, in the overall, it has been suggested that species with less-effective digging mechanisms enjoy wider distribution and numbers, which are definite advantages in the long term. See Fig. 8.

Fig. 8. Side views of heads of typical worm lizards: (a) Wiegmann's worm lizard; (b) *Pachycalamus brevis*.

The teeth of worm lizards interlock with an odd number of teeth in the upper jaw and an even number in the lower jaw. Thus, a biting action leaves a cut like that made by jagged scissors. The jaws tear or rip triangular pieces of flesh from the prey. The cheeks have strong muscles to assist the biting. Unlike the case of snakes, the amphisbaenians can rip out comparatively large pieces of flesh from small mammals.

Extensive information and excellent illustrations of the vast variety of lizards can be found in "Grzimek's Animal Life Encyclopedia," Vol. 6 (Reptiles), Van Nostrand Reinhold, New York.

CLASSIFICATION OF LIZARDS

CLASS: Reptilia (Reptiles)
ORDER: Squamata (Scaly Reptiles)
SUBORDER: Sauria (Lizards)
INFRAORDER: Geckos (Gekkota)

FAMILY: Geckos (Gekkonidae)
Examples: House Geckos (*Hemidactylus*); Common Gecko (*Tarentola mauritanica*), Asian Tokay (*Gekko gekko*); Banded Gecko (*Coleonyx variegatus*); Banded Leaf-toed Gecko (*Hemidactylus fasciatus*); Smooth Gecko (*Thecadactylus rapicauda*); Madagascar Geckos (genus *Phelsuma*); Japanese Gecko (*Gekko japonicus*); Least Geckos (genus *Sphaerodactylus*); African Tropical Gecko (*Hemidactylus mabouia*); Turkish Gecko (*H. turcicus*); European Leaf-fingered Gecko (*Phyllodactylus europaeus*); Sand Gecko (*Chondrodactylus angulifer*); Bibron's Gecko (*Pachydactylus bibronii*); Spotted Gecko (*Pachydactylus maculatus*); Web-footed Gecko (*Palmatogecko rangei*); Emerald Gecko (*Gekko smaragdinus*); Leaf-tailed Gecko (*Uroplatus fimbriatus*); Kuhl's Gecko (*Ptychozoon kuhli*); Panther Gecko (*Eublepharis macularius*); Padless Gecko (*Gonatodes albogularis*); Fat-tailed Gecko (*Oedura marmorata*).

FAMILY: Snake Lizards (Pygopodidae)
Examples: Western Scaly-foot (*Pygopus nigriceps*); Sharp-snouted Snake Lizard (*Lialis burtonis*); New Guinean Snake Lizard (*L.jicari*); Common Scaly-foot (*Pygopus lepidopodus*); Bailey's Scaly-foot (*Pygopus baileyi*).

FAMILY: Dibamids (Dibamidae)
Includes only one genus with three species. Some zoologists consider the dibamids as offshoots of the skinks.

FAMILY: Iguanids (Iguaninae)
SUBFAMILY: Sceloporinae
Examples: Southern Fence Lizard (*Sceloporus undulatus*); Western or Pacific Fence Lizard (*S. occidentalis*); Desert Spiny Lizard (*S. magister*); Tree Lizard (*Urosaurus ornatus*); Side-blotched Lizard or Ground Uta (*Uta stanburiana*); Banded RockLizard (*Petrosaurus mearnsi*); Fringe-toed Lizard (*Uma notata*); Greater Earless Lizard (*Holbrookia texana*); Zebra-tailed Lizard (*Callisaurus draconoides*); Short-horned Lizard or Pigmy Horned Lizard (*Phrynosomas douglasii*); Texas Horned Lizard (*P. cornutum*); Collared Lizard (*Crotaphytus collaris*); Leopard Lizard (*Gambelia wislizenii*).
SUBFAMILY: Tropidurinae
Examples: Spiny-tailed Iguanid (*Uracentron azureum*); Smooth-throated Lizards (genus *Liolaemus*); Crested Keeled Lizards (genus *Leiocephalus*); Narrow-tailed Lizards (genus *Stenocercus*); Madagascan Iguanid (*Oplurus sebae*); Weapon-tailed (*Hoplocercus spinosus*).
SUBFAMILY: Iguanidae
Examples: Common Iguana (*Iguana iguana*); West Indian Iguana (*I. delicatissima*); Rhinoceros Iguana (*Cyclura cornuta*); Marine Iguana (*Amblyrhynchus*); Spiny-tail Iguana (*Ctenosaura pectinata*); Desert Iguana (*Dipsosaurus dorsalis*); Chuckwalla (*Sauromalus ater*);
SUBFAMILY: Basiliscinae
Examples: Common Basilisk (*Basiliscus basiliscus*); Double-crested Basilisk (*B. plumifrons*); Banded Basilisk (*B. vittatus*); Helmeted Lizard (*Corytophanes*); Casque-headed Lizard (*Laemanctus serratus*).
SUBFAMILY: Anolinae
Examples: Long-legged Lizard (*Polychrus marmoratus*); Brazilian Tree Lizard (*Enyalius catenatus*); Patagonian Lizard (*Diplolaemus darwinii*); Sword-tailed Lizard (Genus *Xiphocercus*); Cuban Water Anoles (Genus *Deiroptyx*); False Chameleon (*Chamaeliolis chamaeleontides*); Carolina Anole (*Anolis carolinensis*); Knight Anole (*A. equestrus*).

FAMILY: Agamids (Agamidae)
Examples: Common Agama (*Agama agama*); Black-necked Agama (*A. atricollis*); Kirk's Rock Agama (*A. kirkii*); Bibron's Agama (*A. bibroni*); Desert Agama (*mutabilis*); Hardun (*A. stellio*); Caucasian Agama (*A. caucasica*); African Spiny-tailed Lizard (*Uromastyx acanthinurus*); Egyptian Spiny-tailed Lizard (*U. aegypticus*); Indian Spiny-tailed Lizard (*U. hardwickii*); Toad-headed Agamids (Genus *Phrynocephalus*); Bearded Lizard

(continued)

CLASSIFICATION OF LIZARDS *(continued)*

(*Amphibolurus barbatus*); Spotted Agama (*A. maculatus*); Australian Bloodsucker (*A. muricatus*); Lake Eyre Agama (*A. maculosus*); Frilled or King's Lizard (*Chlamydosaurus kingii*); Lesuer's Water Dragon (*Physignathus lesueurii*); Oriental Water Dragon (*P. cocincinus*); Soa-soa (*Hydrosaurus amboinensis*); Philippine Water Lizard (*H. pustulatus*); Weber's Sailing Lizard (*H. weberi*); Bornean Angle-headed Lizard (*Gonocephalus liogaster*); Lyre-headed Lizard (*Lyriocephalus scutatus*); Indian Bloodsucker (*Calotes versicolor*); Bornean Bloodsucker (*C. cristatellus*); Ceylon Deaf Agamid (*Cophotis ceylanica*); Butterfly Lizard (*Leiolepis belliana*);Sita's Lizard (*Sitana ponticeriana*); Flying Dragon (*Draco volans*); Black-bearded Dragon (*D. melanopogon*); Five-lined Dragon (*D. quinquefasciatus*); Indian Dragon (*D. dussumieri*).

FAMILY: Chameleons (Chamaeleontidae)
Examples: European Chameleon (*Chamaeleo chamaeleon*); African Chameleon (*C. africanus*); Common Chameleon (*C. dilepis*); Two-lined Chameleon (*C. bitaeniatus*); Dwarf Chameleon (*C. pumilus*); Oustalet's Chameleon (*C. oustaleti*); Panther Chameleon (*C. pardalis*); Short-horned Chameleon (*C. brevicornis*); Mountain Chameleon (*C. montium*); Armored Chameleon (*Leandria permeata*).

INFRAORDER: Skinks and Allies (Scincomorpha)

FAMILY: Skinks (Scincidae)
SUBFAMILY: Tiliquinae
Examples: Giant Skink (*Corucia zebrata*); Cape Verde Skink (*Macroscincus cocteaui*:); Stump-tailed Skink (*Tiliqua rugosa*); Blue-tongued Skink (*T. scincoides*); Spiny-tailed Skink (Genus *Egernia*).
SUBFAMILY: (Scincinae)
Examples: Common Skink (*Scincus scincus*); Eastern Skink (*S. mitranus*); Hemprich's Skink (*S. hemprichi*); Arabian Skink (*S. philbyi*); Persian Sand Skink (*Ophiomorus persicus); Speckled Sand Skink (O. punctatissimus*); Algerian Skink (*Eumeces algeriensis*); Schneider's Skink (*E. schneideri*); Five-lined Skink (*E. fasciatus*); Broad-headed Skink or Greater Five-lined Skink (*E. laticeps*); Great Plains Skink (*E. obsoletus*); Cylindrical Skinks (Genus *Chalcides*).
SUBFAMILY: (Lygosominae)
Examples: East Indian Brown-sided Skink (*Mabuya multifasciata*); Keeled Indian Mabuya (*M. carinata*); Müller's Tree Skink (*Sphenomorphus muelleri*); Emerald Skink (*Dasia smaragdina*); Spotted Skink (*D. vittata*) Schmidt's Helmeted Skink (*Tribolonotus schmidti*); Florida Sand Skink (*Neoseps reynoldsi*).

FAMILY: Feylinidae

FAMILY: Anelytropsidae
Example: Mexican Blind Lizard (*Anelytropsis papillosus*).

FAMILY: Girdle-tailed Lizards (Cordylidae)
SUBFAMILY: Cordylinae
Examples: Sungazer Giant Girdled Lizard (*Cordylus giganteus*); Armadillo Lizard or Armadillo Girdled Lizard (*C. cataphractus*); Common Cape Girdled Lizard (*C. cordylus*); Blue-spotted Girdled Lizard (*C. caeruleopunctatus*); Leathery Crag Lizard (*Pseudocordylus microlepidotus*); Transvaal Red-tailed Rock Lizard (*Platysaurus guttatus*); Transvaal Snake Lizard (*Chamaesaura aenea*); Cape Snake Lizard (*C. anguina*).
SUBFAMILY: Gerrhosaurinae
Examples: Smith's Plated Rock Lizard (*Gerrhosaurus validus*); Whip Lizards (genus *Tetradactylus*); Girdled Lizards (genus *Zonosaurus*).

FAMILY: Night Lizards (Xantusiidae)
Examples: Cuban Night Lizard (*Cricosaura typica*); Granite Night Lizard (*Xantusia henshawi*).

FAMILY: Whiptails (Teiidae)
Examples: Chilean Spotted Lizard (*Callopistes maculatus*); Six-lined Racerunner (*Cnemidophorus sexlineatus*); Spotted Whiptail or Blue-bellied Racerunner (*C. sackii*).

FAMILY: True Lizards (Lacertidae)
Examples: Sand Lizard (*Lacerta agilis*); Dwarf Lizard (*L. parva*); Jewelled Lizard (*Timon lepida*); Common Lizard (*Zootoca vivipara*); Plated Lacertids (genus *Psammodromus*); Snake-eyed Lacertids (genus *Ophisops*); Fringe-toed Lacertids (genus *Acanthodactylus*).

INFRAORDER: Anguimorpha

FAMILY: Lateral Foldl Lizards (Anguidae)
SUBFAMILY: Diploglossine Lizards (Diploglossinae)
SUBFAMILY: Alligator Lizards (Gerrhonotinae)
Examples: Glass Lizards (genus *Ophisaurus*); Sheltopusik (*O. apodus*); Eastern Glass Lizard (*O. ventralis*).
SUBFAMILY: Anguine Lizards (Anguinae)
Example: Slow-worm (*Anguis fragilis*).

FAMILY: Shovel-snouted Legless Lizards (genus *Anniella*).

FAMILY: Xenosaurids (Xenosauridae)

Example: Crocodile Lizards (genus *Shinisaurus*); Chinese Crocodile Lizard (*S. crocodilurus*).

INFRAORDER: Varanomorphs

FAMILY: Aigalosauridae

FAMILY: Dolichosauridae

FAMILY: Mosasauridae

FAMILY: Bearded Lizards (Helodermatidae)
Examples: Gila Monster (*Heloderma suspectum*); Mexican Bearded Lizard (*H. horridum*).

FAMILY: Monitors (Varanidae)
Examples: Desert Monitor (*Psammosaurus griseus*); Nile Monitor (*Polydaedalus niloticus*); Cape Monitor (*Empagusia albigularis*); Yellow Monitor (*E. flavescens*); Two-banded Monitor (*Varanus salvator*); Giant Monitor (*V. giganteus*); Dwarf Monitor (*Odatria storri*).

FAMILY: Earless Monitors (Lanthanotidae)
Example: Borneo Earless Monitor (*Lanthanotus borneensis*).

INFRAORDER: Worm Lizards (Amphisbaenia)

FAMILY: Bipedidae
Example: Common Two-Legged Worm Lizard (*Bipes biporus*).

FAMILY: Ringed Lizards (Amphisbaenidae)
Examples: White-Bellied Worm Lizard (*Amphisbaena alba*); Darwin's Ringed Lizard (*A. darwini*); King's Worm Lizard (*A. kingi*); Florida Worm Lizard (*Rhineura floridana*).

FAMILY: Trogonophids (Trogonophidae)
Example: Wiegmann's Worm Lizard (*Trogonophis wiegmanni*).

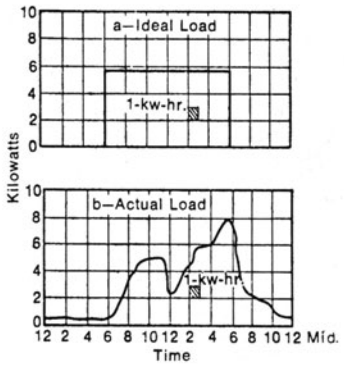
NOTE: Some of the better known as well as other species selected at random are included in the various examples given. The examples represent only some of the species of lizards. The process of classifying the lizards continues at a relatively slow pace toward refinements in classification as well as with the nomenclature used. A few classifications remain controversial among authorities.

LNG (Liquefied Natural Gas). See **Natural Gas.**

LOACHES (*Osteichthyes*). Of the order *Ostariophysi*, suborder *Cypriniformes*, and family *Cobitidae*, loaches are small bottom-feeding fishes of Europe and Asia. They have a long slender body and a group of barbels near the mouth. The European species are eaten. A few fishes of similar form, but not closely related are called African loaches.

Several species are quite popular with tropical-fish fanciers. A brilliantly colored orange species from Borneo and Sumatra, the clown loach (*Botia macracanthus*) is frequently found in small aquariums, despite a habit of uprooting vegetation in the tank. Some clown loaches may reach a length of about 12 inches (30 centimeters) and live up to a quarter-century. Under normal aquarium conditions, the length is limited to a maximum of about 4 inches (10 centimeters). Other small loaches include the coolie loach (*Acanthophthalmus kuhli*), which is somewhat snakelike and found in Thailand and environs; and the spotted weatherfish or spined loach (*Cobitis taenia*) which is noted for its tolerance to wide temperature variations. A spotted species (*Cobitis biwae*) is found in Japan and is sometimes considered a food item. *Misgurnus fossilis* is the European weatherfish, known for its sensitivity to barometric pressure—hence the name, weatherfish. It is quite a large loach, sometimes attaining a length of about 20 inches (51 centimeters), but usually not exceeding 10 inches (25 centimeters).

LOAD FACTOR (Electric). The ideal electric load, from the standpoint of equipment needed and operating routine, is one of constant magnitude and steady duration. A load of this type is shown in the upper part of the accompanying diagram (a). The cost to produce an elementary area of this load curve could be from one-half to three-fourths of that to produce the same unit under the more frequently realized condition shown in the lower curve (b). The problem of variable load is of vital importance to utilities whose chief concern is to put each kilowatt-hour on the transmission line at the lowest possible production cost. Interconnections between utilities and regional power generating facilities help in matching load with generation. See also **Electric Power Production and Distribution.**



Comparison of ideal and actual electric power loads.

LOADING COIL. 1. An inductance inserted at regular intervals along a long transmission line or cable to increase the line's characteristic impedance and reduce its attenuation constant. 2. An inductance inserted in series with an antenna to increase its electrical length.

LOAD MATCHING. 1. Maximum power is delivered to a load when the impedance of the generator is the image impedance of the load. The adjustment of a circuit to provide this condition is called load matching. 2. In induction heating and dielectric heating usage, the process of adjustment of the load circuit impedance to produce the desired energy transfer from the power source to the load.

LOBECTOMY. Surgical removal of a lobe of a gland or organ, such as the lung.

LOCAL AREA NETWORKS. As automation in the factory increases, the need for communication between computers, controllers, and other "intelligent" machines has become critical. In the past, when factory automation was limited to the use of programmable controllers, numerical control of machines, and similar traditional approaches, communication was not a major limiting factor. Each tool was essentially self-contained and the communication requirement was mainly a user interface for controlling and updating machine operation. With the accelerated growth of automated tools and processes, communication between these entities is required to control not only their operation, but also their interrelationships. To this is added the desire to overlay environmental control, energy management, and materials requirement planning (MRP) to the factory operation. The end result is that inter-computer/controller communication has become the largest single problem to be addressed for factory automation. These interrelationships are shown in a generalized fashion in Fig. 1.

Early communication needs were served with point-to-point data links, as simply indicated in Fig. 2. Communication was relatively simple. The *star topology* for the communication system (Fig. 3) was developed so that multiple computers could communicate. The central or "master" node uses a communications port with multiple drops as shown in Fig. 4. The master is required to handle traffic from all the nodes attached, poll the other nodes for status, and, if necessary, accept data from one node to be routed to another. This heavy software burden on the master is also shared to a lesser degree among all the attached nodes. In addition, star topology requires routing a separate wire for every piece of equipment attached. This makes it difficult to wire and even more difficult to change. Also, the star topologies are inflexible regarding the number of nodes that can be attached. Either the user must invest in unused connections for further expansion, or have a system that cannot grow.

To overcome some of these disadvantages, multidrop protocols were established and standardized. Data loops, such as SDLC (Synchronous Data Link Control), were developed as well as other topologies, including buses and rings. Some of the early standards are shown in Fig. 5. The topology of these standards makes it easy to add (or subtract) nodes

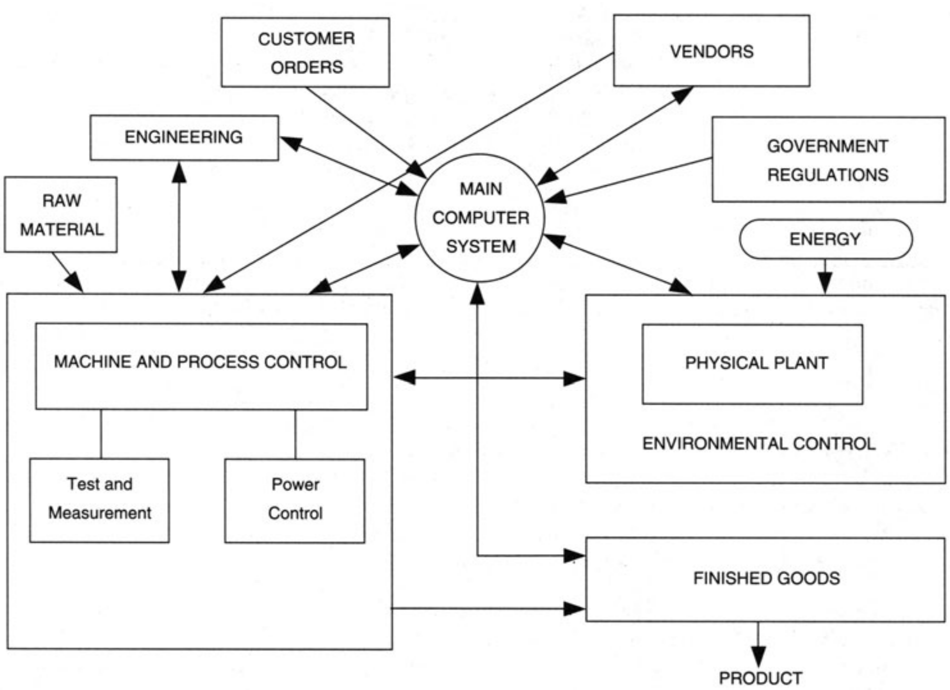


Fig. 1. Intercommunications in a generalized automated factory situation.

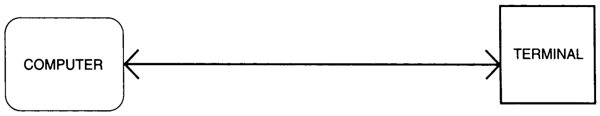


Fig. 2. Point-to-point communication.

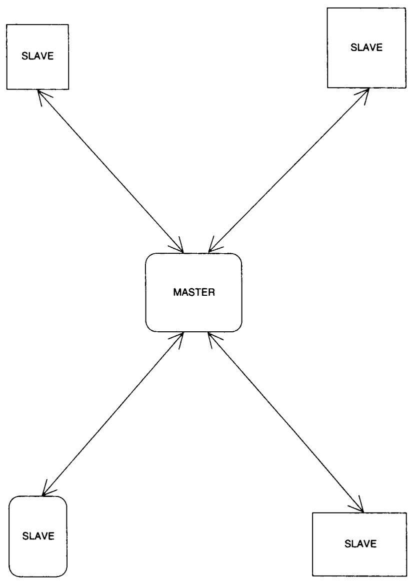


Fig. 3. Star topology.

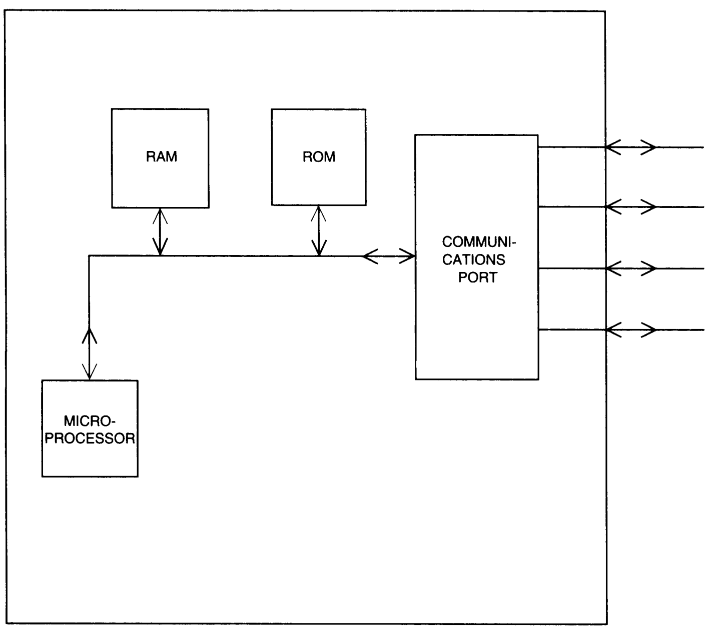


Fig. 4. Master node for star topology. RAM = randon access memory; ROM = read only memory.

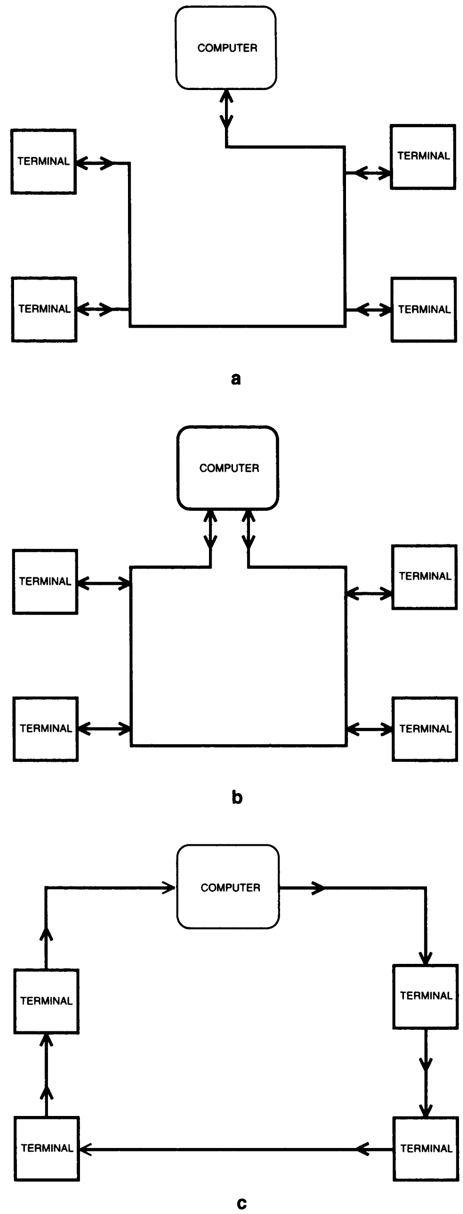


Fig. 5. Communication protocols and standards: (a) Data loop topology; (b) bus topology; (c) ring topology.

on the network. The wiring is also easier because a single wire is routed to all nodes. In the case of the ring and loop, the wire is also returned to the master. Inasmuch as wiring and maintenance are major costs of data communication, these topologies have virtually replaced star networks. These systems do have a common weakness, i.e., one mode is the "master," with the task of determining which station may transmit at any given time. As the number of nodes increases, throughput becomes a problem because: (1) a great deal of "overhead" activity may be required to determine who may transmit, and (2) entire messages may have to be repeated because some protocols only allow master-slave communications, i.e., a slave-to-slave message must be sent first to the master and then repeated by the master to the intended slave receiver. Reliability is another problem. If the master fails, communication comes to a halt. The need for multinode networks without these kinds of problems and restraints led to the development of *Local Area Networks* (LANs), which use peer-to-peer communication. In this system, no one node is in charge and all nodes have an equal opportunity to transmit.

A LAN is a distributed communication network with the following characteristics: (1) peer-to-peer oriented (no master); (2) from 2 to 200 data devices (nodes) may be incorporated; (3) distance is limited to less than 2 km (1.2 mi); and (4) from 1 to 20 M bits per second data rates.

In a local area network, each node is an independent computer system. Since there is no master to control traffic, each node must determine when it has the right to transmit. In a typical system, the host computer of the node is free to perform its job while the LAN protocol unit is moving data on and off the network. See Fig. 6.

As previously shown in Fig. 1, the need for LANs is driven by the proliferation of computer functions. The real motivating factor is that computers are now so cost-effective and relatively inexpensive that they are used throughout the factory floor, processing plant, and office. These computers not only must communicate with the large mainframe computers, but also with each other. The demands of materials require-

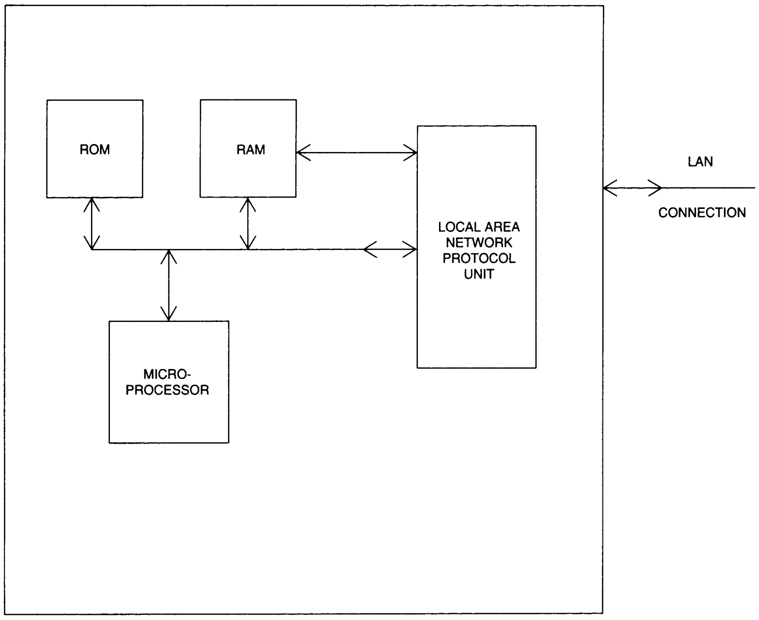


Fig. 6. A local area network node. RAM = random access memory; ROM = read only memory.

ment planning, such as scheduling, inventory control, management, etc. require constant monitoring and data acquisition. Further, factory floor management requires coordination of machine operation; environmental control requires constant monitoring, among numerous other factors that go well beyond the traditional tasks of regulatory or sequencing controllers. The need to communicate status between devices (for a total integrated environment) and the sharing of costly resources, such as large-capacity disk storage, line printers, etc., have driven the requirement for LAN communication networks.

Requirements of the Industrial Environment. Although the need for LANs exists both in the office area and on the factory floor, the latter environment imposes some special restraints. The needs of the industrial environment require:

1. *Noise Tolerance.* Since a LAN will have long cables running throughout the factory, the amount of noise picked up can be large. The LAN must be capable of performing reliably in an electrically noisy area. The physical interface should be defined to provide a significant degree of noise rejection. The protocol must allow for easy recovery from data errors.

2. *Fast Response.* The LAN in an industrial situation should have a guaranteed maximum response time, i.e., the network must be able to transmit an urgent message within a specified time limit. The real-time characteristic of industrial control demands communication within a known time frame.

3. *Ability to Handle Priority Messages.* On the factory floor, both control and status data will be carried over the same network. A control message should have a higher priority and be transmitted before other messages.

Common Standards. A local area network standard that serves the harsh factory environment well can also be used for the less demanding office and administrative areas. Unless the requirements for the factory add too much cost, the factory floor standard can be the choice for the entire network. A common standard is advantageous inasmuch as system and information handling elements located in both office and factory environments must communicate with each other.

To meet current communication needs, different types of LANs are possible. Many of these already have been developed. All LANs provide the same basic service—to allow computers to pass data. However, since the major function of a network is to connect many different computers, standards are needed. The standard not only should describe how nodes are connected, but the protocol followed in transferring data as well. Ideally, the standard should be sufficiently comprehensive to permit any computer following it to pass data to any other computer that follows the same standard.

Several organizations have been working on the standards problem for a number of years, including the International Standards Organization (ISO), the Institute of Electrical and Electronics Engineers (IEEE), as well as some major users of LANs, such as General Motors Corporation, which has developed MAP (Manufacturing Automation Protocol).

Examples of Protocols

Carrier Sense Multiple Access with Collision (CSMA/CD). This is a baseband system with a bus architecture. See Fig. 5(b). Baseband is a term used to describe a system where the information being sent over the wire is not modulated, i.e., the "ones" are represented by one voltage level and the "zeros" by another voltage level. Normally, only one station transmits at any one time. All other stations hear and record the message. The receiving stations then compare the designated address of the message with their address. The one station which matches will pass the message to its upper layers, while the others will throw it away. Obviously, if the message is affected by noise (detected by the frame check sequence), all stations will throw the message away.

The CSMA/CD protocol requires that a station listen before it can transmit data. If the station hears another station already transmitting (*carrier sense*), the stations wanting to transmit must wait. When a station does not hear any other station transmitting (*no carrier sense*), it can start to transmit. Since more than one station can be waiting, it is possible for multiple stations to start transmitting at the same time. This causes the messages from both stations to become garbled (called a "collision"). A collision is not a freak accident, but is a normal way of operation for networks using CSMA/CD. The chances of collision are increased by the fact that signals take a finite amount of time to travel from one end of the cable to the other. If a station on one end of the cable starts transmitting, a station on the other end will "think" that no other station is transmitting during this travel-time interval and that transmission can be resumed. After a station has started transmitting, it must detect when another station is also transmitting. If this happens (*collision detection*), the station must stop transmitting. Before quitting the transmission, however, the station must make sure that every other station is aware that the last frame is in error and must be ignored. To do this, the station sends out a "jam" which is simply an invalid signal. This jam guarantees that the other colliding station also detects the collision and quits transmitting. Each station that was transmitting must then wait before trying again. To make certain that the two (or more) stations that just collided do not collide again, each station selects a random time to wait. The first station to time out will then look for silence on the cable and retransmit its message.

Token Bus. This standard has been selected for use in the previously mentioned MAP protocol. Token bus is also a bus topology, but differs in two ways from CSMA/CD: (1) The right to talk is controlled by passing a "token," and (2) the data on the bus are always carrier modulated. In the token bus system, one station is said to have an imaginary token. This station is the only one on the network that is allowed to transmit data. When this station has no more data to transmit (or it has held the token beyond the specific maximum time limit), it passes the token to another station. This token pass is accomplished by sending a special message to the next station. After this second station has used the token, it passes it to the next station, and so on. After all other stations have used the token, the original station is passed the token again. A station (example, station A) will normally receive the token from one station (B) and pass the token to the third station (C). The token ends up being passed around in a logical token ring (A to C to B to A to C to B ...). The exception to this is when a station wakes up or dies. For example, if a fourth station (D) gets in the logical token ring between A and C, station A would then pass the token to D so that the token would go: (A to D to C to B to A to D ...). Only the station with the token can transmit so that every station gets its turn to talk without interfering with anyone else. Obviously, the protocol also has provisions which allow stations to enter and to leave the logical token ring.

The second difference between token bus and CSMA/CD is that with the token bus, data are always modulated before being sent. The data are not sent out as a level, but as a frequency. There are three different modulation schemes allowed—two single-channel and one broadband. Single-channel modulation permits only the token bus data on the cable. The broadband method is similar to CATV (community antenna television) and allows many different signals to exist on the same cable, including video and voice, in addition to the token bus data. The single-channel techniques are simpler, less costly, and easier to implement than broadband. Broadband, however, is of higher performance, permitting much longer distances and, very important, satisfies both present and future communication needs by allowing as many channels as needed (within the bandwidth of the cable).

Token Ring. Originally, token ring and token bus used the same protocol with different topologies. The two systems still remain rather similar. As shown by Fig. 5, any one node will only receive data from the "upstream" node and will only send data to the "downstream" node. All communication is done on a baseband point-to-point basis. This would seem to imply that one node can talk only to its downstream node. This is not the case, inasmuch as each station repeats what it hears from the upstream station to the downstream station. Since the last station is connected to the first (forming a ring), any station can send data to any other station. Precaution must be taken to prevent a short message from being retransmitted around the ring forever. This is prevented by having the transmitting station remove its messages from the ring once they have gone around the ring one time.

The "right to talk" for the token ring scheme is also an imaginary token. The simplicity of the token ring system is that the station with the token simply sends it to the next downstream station, which either uses the token or passes it on to the next station. Space here does not permit the inclusion of numerous other pros and cons pertaining to these systems.

Additional Reading

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LOCOMOTION. The process of moving from place to place, a characteristic power of most animals and a lesser distinction between them and the majority of plants.

Locomotion is necessary to animals because their food is organic and in most environments does not reach the animal through external forces. Even the sessile animals, which may or may not be capable of some locomotion, often accomplish the same end by bringing food within reach through their own activities.

In the water the weight of the surrounding medium is so great that the animal may float, and the resistance offered to its body is sufficient to be utilized for propulsion. The body is so shaped that resistance is little in the direction of locomotion but great where propulsive effort is expended. Projections from the surface which beat against the water like oars or push or pull by undulating are common organs of locomotion here. They include cilia and flagella in one-celled and small multicellular forms, specialized jointed appendages of arthropods, and fins and flippers of vertebrates. Undulation of the body itself is a sufficient means of propulsion in some animals.

Some aquatic forms rest on the bottom and the terrestrial animals are forced to rest on some solid support at least intermittently because the air is too light to float them. In many of these forms the friction of contact with a solid is utilized by the development of movable supporting appendages which are shifted alternately to change the animal's position. This means of locomotion is known as walking. Other animals, notably the worms, creep through the action of muscles in the body wall. The body is progressively elongated and shortened, parts being thrust ahead and then drawn up to the maxiumum point of advance. In this type of locomotion they are aided by suckers or setae in some cases to grip the supporting surface.

Running may involve no other difference from walking than more rapid movement or it may also involve a change in the order of movement of the appendages and in their position when used, as in the various gaits of a horse. Jumping always differs in that the appendages set farthest back must be powerful enough to project the entire animal through the air. It is highly developed in such insects as the flea beetles and the grasshoppers and in the frogs and kangaroos among the vertebrates. In this class a gallop is no more than a series of leaps.

Locomotion in the terrestrial vertebrates also shows progressive change in the manner of using the appendages. The entire sole of the foot rests on the ground in man and the apes, and they are said to be plantigrade. This posture is well adapted to walking but not to running. Animals that need speed are digitigrade, resting on the toes. This position adds the length of the feet to that of the legs and permits a longer stride. It also adds the springiness incidental to the greater freedom of the ankle joint. The final expression of this position of the leg appears in the unguligrade (Ungulata) hoofed animals where only the hoof comes into contact with the ground. Man is plantigrade in walking and at rest but rises to his toes when he runs.

The locomotion of snakes is a highly specialized creeping process in which the ribs serve as the movable appendages and the grip of the body on the ground is provided by the broad scales of the ventral surface which project backward.

Climbing animals may merely run along branches, aided by sharp claws to provide a secure grip. The sloths, however, have the claws developed as great hooks which suspend them in an inverted position. They walk as well as hang upside down. The primates show the most extreme specialization for a form of locomotion in the trees called brachiation. Their pectoral appendages are arms, adapted for grasping and suspension, and the pelvic appendages are supporting legs. They move by swinging from branch to branch or by shifting from one hold to another, as human beings climb.

Locomotion in the air is the highly specialized process of flight.

LOCUST (Grasshopper; *Insecta, Orthoptera*). The term "locust" is more properly applied to the "short-horned" grasshoppers which include the migratory locusts appearing in Africa, Asia and the plains of North America as serious crop pests. These animals have always been an important food of aboriginal peoples, and are still consumed in many parts of the world. The term grasshopper is now generally restricted to the common non-migratory forms. The name locust is also commonly but wrongly applied to the Cicada or tree cricket.

The adults measure about 2 inches (5 centimeters) in length; some are much larger. The body is thick, strong, tapering to the end of its folded wings. Large eyes are high on the head and located just above the two short antennae. A stiff, thick jacket joining the head covers and protects the back of the neck. The wings, shaped much like some aircraft wings, are held close to the body when at rest. The wings are strong, large, heavily veined and twice as wide at the center as at the ends, and about $\frac{1}{2}$ inch (12 millimeters) longer than the body.

The six legs of the locust are comprised of several segments, connected to the body; the two shorter front legs are located under the shoulders. The two legs located near the center of the body are long and powerful and used for making long hops. The thigh of the leg is oar-shaped and heavily veined. It is connected to a strong joint at the body which aids greatly in the thrust of the kick when jumping.

The "song" of the locust is created by spurs or spines on the inner area of the hind legs which are rubbed against the wings which have a raised, spurred area on the outside surface. This gives a rasping sound.

The desert locust (*Schistocera gregaria*) of India swarms during the summer monsoons. In autumn, these insects migrate to Iran, Arabia, Soviet Asia, Syria, and Egypt. In early winter, they return to India and East Africa to breed.

Possibly the most impressive of all insect flights will be that of a swarm of billions of locusts. In 1889, a flight crossed the Red Sea estimated to be 2000 square miles (5180 square kilometers) in extent. Desert locusts have appeared in England, apparently flying from southern algeria, possibly assisted by a tail wind. Swarms of locusts have been reported since ancient times. The Book of Joel describes the army that blackened out the sun—behind them a desolate wilderness and nothing escaping them.

There are about 2,000 species of locusts, of which about 20 important species are capable of causing crop desolation and accompanying famine in some areas of the world. The huge migrations are caused by hunting for food, and where locusts find food they will eat their weight daily if it is plentiful. Fortunately, the mortality rate of the insect is high when a swarm encounters stormy weather. See also **Cicada.**

LOCUST (Seventeen year; *Insecta, Homoptera, Cicadidae*). The 17-year locust (*Magicicada septendecim*, Linne), not a true locust, is named for its life cycle of 17 years. There are, however, many broods which overlap, with adults appearing in different years. There is also the 13-year locust (*M. septendecim tridecim*, Riley). These insects are sometimes confused with the *dog-day cicada* (*Tibicen linnei*) (Smith and Grossbeck), also commonly termed a locust, which has a 9-year cycle. The latter insect is not nearly so damaging as the other two species just mentioned. See also **Cicada.**

The 17-year and 13-year locusts (properly known as *periodical cicadas*) make rough punctures in twigs and small branches of apple and numerous other fruit trees. Damage does not result from the feeding of the insects, but from the puncturing of the twigs when the female deposits her eggs. In the United States, the 13-year locust ranges in a line from Virginia to southern Iowa and thence southward to the Gulf of Mexico and eastward to the Atlantic shore. The 17-year locust is found from the New England states westward to Wisconsin, southwestward into Kansas and Missouri, and south as far as Alabama and northern Georgia. It is most abundant east of the Mississippi River and greatly infests much of Wisconsin, Iowa, Tennessee, and South Carolina, and nearly all of Illinois, Indiana, Ohio, Virginia, Kentucky, West Virginia, Pennsylvania, Connecticut, and New Jersey.

The 17-year locust has the longest period of development of any insect known. As previously described, the eggs are laid in twig punctures in late spring and mid-summer. Each female deposits 400–600 eggs, with about 20 eggs per puncture. Within 6–7 weeks the eggs hatch. The young are antlike in appearance and, when hatched, drop to the ground where they enter into cracks in the soil. They feed on sap from tree rootlets. Feeding is very slow and their presence cannot be detected by any apparent deterioration of the tree, even though there may be many thousands of these creatures at the base of an affected tree. Depending upon the species, these nymphs require from 13 to 17 years to achieve maturity. At that time, the insects are about an inch (2.5 centimeters) long and appear something like a crayfish. The insects burrow to the surface, sometimes forming mud cones or chimneys that may protrude 2–3 inches (5–8 centimeters) above groundlevel. Massive numbers of these nymphs emerge within a very short period, usually climbing the tree after sunset. They temporarily take hold of the bark until the adult insect removes itself from the nymphal shell. Leaving the empty skins on the tree, they take flight and are ready to mate during a period of 30–40 days. It is estimated that over 40,000 adults may emerge from a single large tree within a period of a few days.

LOCUST TREES. Several leguminous trees are commonly called locust trees. Most notable is the black locust, *Robinia pseudoacacia*. It is a medium-sized tree native to the Appalachian and Ozark mountains. The twigs have a pair of spines about $\frac{1}{2}$ inch (1.3 centimeters) long at the base of each leaf, although spineless varieties have been developed. The leaves are compound, 8–14 inches (20.3–35.5 centimeters) long. In the spring the tree produces its flowers, which are white or pink and very fragrant, hanging in clusters 4–8 inches (10–20.3 centimeters) long. The wood is hard and tough, and is used for fence posts, mine timbers, and rough construction. The tree is commonly planted for ornament and shade, or for erosion control.

The honey locust, *Gleditsia*, is also frequently known simply as locust. A variety of this genus, the Moraine locust, has become popular as a shade tree.

Record locust trees, as reported by The American Forestry Association, are listed in the accompanying table.

LOESS. Loess is a buff-colored, wind-blown deposit of fine silt or marl, usually unstratified, which is often exposed in bluffs with steep to vertical faces. Loess is found in the United States in the Mississippi valley from Louisiana to Iowa, and along the course of the Missouri. The average thickness of the loess here is about 20 feet (6 meters), but may range to 50–100 feet (15–30 meters). Loess also occurs in central Europe, Mongolia and China where it is said to attain a thickness as great as 300 feet (90 meters). The loess of the United States and Europe is believed to be the finer materials first transported and deposited by the waters of the melting ice sheets of the glacial period, and later blown to considerable distances and sometimes deposited in lakes. The Asiatic loess seems to be wholly wind transported, the source of the dust being, perhaps, the great deserts of central Asia. In the latter case the accumulation of such thick deposits is attributed to the binding power of successive generations of grasses whose former existence is suggested by a network of narrow tubes.

See also **Erosion.**

LOGARITHM. If B is an arbitrarily chosen number greater than unity, then the logarithm L of any other number N is defined by $N = B^L$; $L = \log_B N$. The chosen number B is the base of the system of logarithms. For any base, $\log_B 1 = 0$; $\log_B B = 1$. The fundamental properties of logarithms are $\log ab = \log a + \log b$; $\log a/b = \log a - \log b$; $a^n = n \log a$, where a, b are positive numbers and n may be greater than unity or less than unity (thus a power or a root of the number, rational or irrational). Two systems of logarithms are generally used: common and natural.

The former, also called *Briggs logarithms*, uses the base 10 and is particularly useful for numerical calculations. The common logarithm

RECORD LOCUST TREES IN THE UNITED STATES[1]

Specimen	Circumference[2] Inches	Circumference[2] Centimeters	Height Feet	Height Meters	Spread Feet	Spread Meters	Location
Black locust (1974) (*Robinia pseudoacacia*)	280	711	96	29.3	92	28.0	New York
New Mexico locust (1981) (*Robinia neomexicana*)	89	226	23	7.0	19	5.8	New Mexico
Thornless honey locust (1991)[3] (*Gleditsia trianconthos var. inermis*)	231	587	90	27.4	85	25.9	Pennsylvania
Thornless honey locust (1991)[3] (*Gleditsia trianconthos var*)	216	549	115	35.1	124	37.8	Michigan
Water locust (1991)[3] (*Gleditsia aquatica*)	109	277	74	22.6	73	22.3	Pennsylvania
Water locust (1991)[3] (*Gleditsia aquatica*)	89	226	94	28.7	49	14.9	Illinois

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

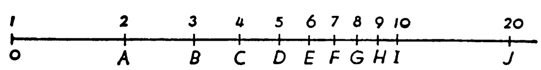
of a number N could be indicated by $\log_{10} N$ but the notation $\log N$ is more usual. Since any number may be written in the form $N = 10^n \times M$, where n is an integer, positive or negative, its common logarithm is $\log N = n + \log M$. The first part of this sum is the *characteristic* of the logarithm and it may be obtained by inspection of the given number. The second part, called the *mantissa*, is an irrational number, less than unity and usually given in decimal form. Tables of the mantissas of common logarithms are available with four, five, six, etc., significant figures so that any required accuracy can be obtained in numerical calculations.

If a logarithm is given, the number which corresponds to it is the *antilogarithm*. The *cologarithm* is the logarithm of the reciprocal of a number. Thus $\operatorname{colog} N = \log(1/N) = -\log N$. In principle, it could simplify calculations involving quotients, for negative mantissas would not occur. This is not a serious handicap, however, and although cologarithm tables are sometimes given in handbooks, they are seldom used.

The *natural logarithm* has the irrational base $e = 2.71828\ldots$ and is also called the Napierian or hyperbolic system. Such logarithms occur as the result of differentiation, integration, etc., and they often appear in equations representing physical phenomena. Instead of the more exact symbol $\log_e N$, the abbreviated notation $\ln N$ is customary, especially in scientific work. The modulus of the common system relative to the natural system of logarithms is $\log e = 0.434294\ldots$ and, inversely, $\ln 10 = 2.302585\ldots$ is the modulus of the natural system. Conversion of one system of logarithms to another is made from the resulting relations: $\log N = 0.434294\ldots \ln N$; $\ln N = 2.302585\ldots \log N$. Although tables of natural logarithms are available, it is normally simpler to compute with common logarithms and then convert to the natural base.

A logarithmic scale is one in which the distance from the origin to any scale mark is proportional to the logarithm of the number attached to that mark.

Thus, the accompanying figure shows a (common) logarithmic scale with the numbers 1, 2, 3, . . . , 10, . . . attached to the division marks; if we take OI as the unit of length, then the distances OA, OB, OC, etc., are represented by log 2, log 3, log 4, etc., so that $OI = \log 10 = 1$. In going from left to right, the scale marks will become closer and closer together.

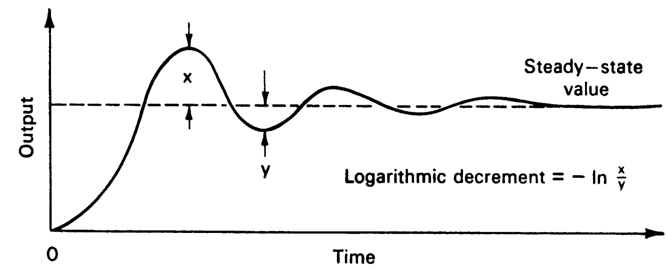


Logarithmic scale.

The logarithmic scale is applied in the slide rule and in logarithmic paper. In the latter case, if both abscissa and ordinate are marked logarithmically, the paper is called log-log paper; if the abscissa is equally spaced and only the ordinate is logarithmic, it is semilog paper. See also **Curve Fitting;** and **Logarithmic Function.**

LOGARITHMIC CHART. A graph where one or both axis are scaled in terms of logarithms of the variables. The chart may be called a semi- or double-logarithmic chart according to whether only the ordinate or both the ordinate and abscissa are on a logarithmic scale. In general, the logarithmic method of plotting is used when relative changes are important, since equal linear displacement on a logarithmic scale indicates equal proportional changes in the variable itself.

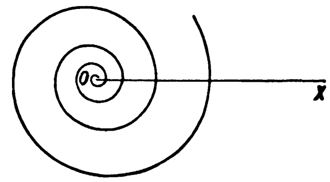
LOGARITHMIC DECREMENT. For a system having an oscillatory response which decays with increasing time, the logarithmic decrement is defined as the negative natural logarithm of the ratio of two consecutive excursions of the response about the bias or steady-state value. The ratio is taken as shown in the figure. This quantity is used as a measure of the internal friction in an oscillating body. The logarithmic decrement equals one half the specific damping capacity.



Logarithmic decrement of oscillatory response.

LOGARITHMIC FUNCTION. Its equation in rectangular coordinates is $y = a \ln x$ and the exponential function, $x = e^{y/a}$ is its inverse. When plotted on ordinary graph paper, the ordinate increases in arithmetic progression and the abscissa in geometric progression. Semilogarithmic paper is more convenient for such a graph (see **Logarithm**). A plot of $x + \log y =$ constant or $y + \log x =$ constant is then a straight line. See also **Curve Fitting.**

LOGARITHMIC SPIRAL. A transcendental plane curve, with equation in polar coordinates, $\ln r = a\theta$. It is also known as an equiangular spiral since it cuts the radius vectors at a constant angle. As θ increases, the curve winds around the pole at ever-increasing distances; when θ is negative, the curve continually approaches, but never reaches, the pole. Its evolute is a congruent logarithmic spiral. It is said that the tombstone of James Bernoulli (1654ü1705), the celebrated Swiss mathematician, is inscribed with this curve and the words "Eaden mutato resurgo." See also **Bernoulli Number and Polynomial.**

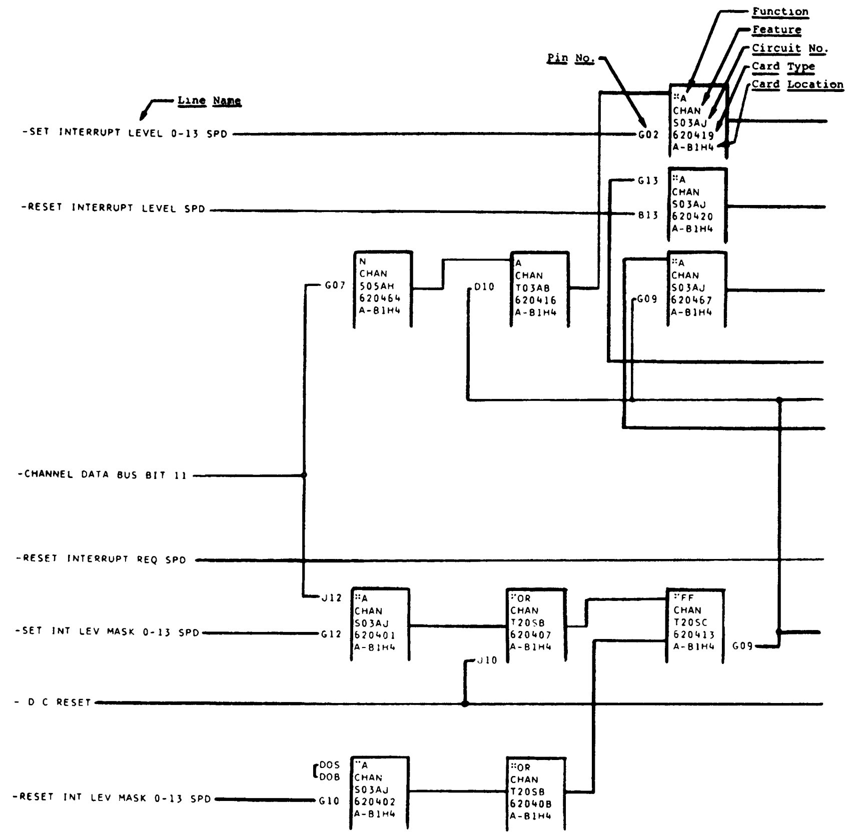


Logarithmic spiral.

LOGICAL OPERATION (Computer System). 1. A logical or Boolean operation on N-state variables which yields a single N-state variable; e.g., a comparison on the 3-state variables A and B, each represented by $-$, 0, or $+$, which yields: $-$ when A is less than B, 0 when A equals B, and $+$ when A is greater than B. Specifically, operations such as AND, OR, and NOT on two-state variables which occur in the algebra of logic; i.e., Boolean algebra. 2. The operations of logical shifting, masking, and other non-arithmetical operations of a computer.

See also **AND (Circuit); NAND (Circuit); NOR (Circuit); NOT (Circuit);** and **OR (Circuit).**

LOGIC (Computer System). In hardware, a term referring to the circuits which perform the arithmetic and control operations in a computer. In designing digital computers, the principles of Boolean algebra are employed. The logical elements of AND, OR, INVERTER, EXCLUSIVE OR, NOR, NAND, NOT, etc. are combined to perform a specified function. Each of the logical elements is implemented as an electronic circuit which, in turn, is connected to other circuits to achieve the desired result. The word logic is also used in computer programming to refer to the procedure or algorithm necessary to achieve a result.



Computer logic diagram.

LOGIC DIAGRAM (Computer System). A drawing which indicates the interconnection of the individual logic elements in a computer. A logic diagram incorporates all of the information needed for wiring a computer. The logic diagram given in the accompanying figure shows the logic blocks and electrical interconnections. The logic block in the diagram contains the name of the function performed by it, such as AND, OR, and FLIP-FLOP; and the physical location of the circuit within the computer. Input and output signal connections for the function are given by the logic block. The logic diagram also shows the connection of a given logic block to other logic blocks. The computer supplier often provides wiring lists for the machine and a printed logic diagram. The logic diagram is used as an aid in troubleshooting the system. When utilizing large-scale integrated (LSI) circuits, each block may contain a complex logic function, such as an adder or register, rather than single elemental logic functions. Although not needed for physical wiring or troubleshooting, these complex logic functions can, in turn, be expressed in terms of elemental logic similar to that shown in the accompanying figure.

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LOGISTIC CURVE. The logistic curve is a growth curve used to describe functions which continually increase, gradually at first, more rapidly in the middle growth period, and slowly again, reaching a maximum at the end of the growth. We write its equation

$$y = \frac{k'}{1 + e^{a+bt}}, \quad b < 0$$

the symmetrical logistic, where t represents time and y is the population size. A more general form is

$$y = k_1 + \frac{k_2}{1 + e^{a+bt-ct2}}, \quad c < 0$$

the asymmetrical or skew logistic. This usage of the word "logistic" has nothing to do with the military connotation or with the European meaning of "formal logic."

LOGIT. A transformation of a variable used particularly in the analysis of dose-response relationships. If p is the probability of a certain response on dose x, the logit is defined as

$$y = \log_e\{p/(\mathrm{a} - p)\}$$

and analysis proceeds by considering the relation between y and x.

LOG MEAN TEMPERATURE DIFFERENCE (LMTD). An average temperature difference between the hot and cold side of a piece of equipment, e.g., a heat exchanger, for use where the temperature difference may vary along the equipment. Its form is

$$\frac{\Delta T_1 - \Delta T_2}{2.3 \log \dfrac{\Delta T_1}{\Delta T_2}}$$

where ΔT_1 is the temperature difference at one end of the equipment and ΔT_2 is the temperature difference at the other.

LOG (Navigation). A term used with two different meanings: a speed-measuring device, and a record book. Prior to the middle of the nineteenth century, the speed of a ship relative to the surface of the water was measured by the log chip and line. "Heaving the log" was a duty performed every time the ship's bell was struck, i.e., every half-hour, and the speeds determined were entered in a book, which came to be known as the log book. Since this log book was always available to the watch officer, it became customary to enter all important incidents relating to the operations of the ship, behavior of members of the crew, conditions of the weather and sea, and, in fact, anything the watch officer might think worthy of recording. This "deck log" was turned over to the captain, who abstracted all important material and made up the ship's log. At present, practically every department of the ship, deck, engine room, ordnance, steward, etc., keeps its individual rough log, and the ship's log is made up from these under the supervision of the captain.

The oldest and perhaps even now, the most accurate and reliable method for determining the speed of a ship relative to the water is by use of the log chip and line. The log chip is a wooden quadrant, loaded along its circular edge so that it will float upright in the water. It is attached to the log line by a 3-legged bridle, the upper leg of which is attached to the apex of the log chip in such a manner that a sharp jerk will free it. Then the log chip may be easily hauled back to the ship. When the log chip is thrown overboard, it floats, nearly submerged, with the flat surface perpendicular to the motion of the ship, and the line runs out over the stern with the speed at which the ship is moving through the water. For measuring the speed at which the line runs out, a sandglass was originally used. For speeds under 4 knots, a 28-second-glass was employed, and for higher speeds, a 14-second-glass was available. The line was divided into lengths by threading pieces of fish line through it at specified distances. In the pieces of fish line, 1, 2, 3, etc., knots were tied to indicate the number of lengths that had run out in the given time. The distance between markers was determined by the ratio between 28 seconds and the number of seconds in an hour:

$$\frac{\text{distance}}{6{,}080 \text{ ft.}} = \frac{28 \text{ sec.}}{3{,}600 \text{ sec}}\text{; hence, distance} = 47'\ 3.5''$$

In order that the log chip may be well clear of the turbulence in the wake, a certain amount of stray line is provided, with a red marker indicating the beginning of measurement.

To "heave the log," two men are necessary, one to tend line and the other to operate the sandglass. The first operator throws the chip overboard and gives the word "tip" when the red marker passes through his fingers. When the last grain of sand runs through the glass, the timekeeper calls "check," and the line-tender grabs the line. This gives the sudden jerk necessary to free the bridle. The line-tender then notes the number of "knots" and approximate fractions thereof that have passed through his fingers and reports that number to the officer of the deck, the number of knots being equal to the speed of the ship in nautical miles per hour.

The use of the log chip and line gives the instantaneous speed of the ship, and the values obtained over a day must be averaged to obtain the total distance run. Many different types of patent logs have been devised to give both speed and distance run. They are all subject to unexpected errors and should be checked frequently by comparison with the log chip and line. On the high-speed ships of the modern navy, the old-fashioned line is impracticable. To check patent logs, the number of seconds required for a given length of line to run out is determined by means of a stop watch, and the speed is calculated.

See also **Course; Navigation;** and **Patent Log.**

LOGWOOD TREE. Of the family *Caesalpiniaceae* (senna family), the *Haematoxylon campechianum* is a small tree native to central America and has been extensively planted there and in the West Indies and South America. Rarely exceeding 25 feet (7.5 meters) in height, the tree has pinnately compounded leaves with smooth obovate leaflets, and fragrant yellow flowers in terminal racemes. The fruit is a dry two-seeded pod. The wood is very hard and yellow. Upon exposure to air, it turns red. It has a rather pleasant scent. Dye is obtained from the heartwood which is cut into chips. In earlier years, logwood dye was extensively used. To make the dye, mordants must be added, in this case the salt of some metal, usually iron. Haematoxylon has been used in the manufacture of inks and as a histological stain in the preparation of organic tissues for microscopic examination.

LONDON DIPOLE THEORY. A theory that accounts for the attractive forces between molecules by considering the interactions between the instantaneous dipole moments of the molecules. By considering the first order perturbation, it is shown that the interaction energy varies inversely as the sixth power of the distance between the molecules.

Weak forces of attraction (also called dispersion forces) are exerted on each other by inert atoms. London, on the basis of quantum mechanics, has shown that they are due to the perturbation of the repulsive ground state by the higher electronic states of the system consisting of the two atoms. This perturbation at large internuclear distances r, gives a potential energy decreasing as $-(1/r^6)$ toward smaller r values. At smaller distances r the strong repulsion of the zero-valent atoms sets in, so that only a very shallow minimum at a relatively large internuclear distance results. Analogous forces also add to the mutual attraction of atoms with free valences and of molecules with or without permanent dipole moments. See also **van der Waals Forces.**

LONGITUDE. The longitude of a point on the surface of the earth is the angular distance measured along the earth's equator from the meridian, through Greenwich, England, to the point where the local meridian of the point cuts the equator. Longitude may be expressed either in units of time (hours, minutes, and seconds) or of angle (degrees, minutes, and seconds). It is measured east or west from Greenwich, through 12 hours or 180°. For convenience in navigation, west longitude is marked plus (+) and east longitude minus (−).

LONG RANGE ORDER. A system may be said to possess long range order if it is possible to assign letters A, B, C, etc., to the sites of the lattice in such a way that there is a greater probability of finding an atom of type A on an A-site, of type B on a B-site, and so on, than any other arrangement. Such order is characteristic of order-disorder transitions in binary alloys. It is measured by the parameter

$$S = \frac{r - w}{r + w}$$

where r and w are the numbers of atoms on right and wrong sites respectively.

LOON (*Aves, Gaviiformes, Gaviidae*). Large birds sometimes called divers which range in length from 58 to 90 centimeters (23 to 35 inches) and in weight from 1 to 6.4 kilograms (2.2 to 14.1 pounds). See also **Gaviiformes.** There are extensive webs between the three front toes, and they have a short tail with 16 to 20 tail feathers and 11 primaries. In flight the head and neck are somewhat lowered.

There is only 1 genus, *Gavia*, with four species found in the northern tundra and forest zones of the Old and New Worlds. Three of the species have a black and white checkered pattern on the wings of the breeding plumage: (1) Arctic Loon (Black-throated diver, *Gavia arctica*; see accompanying illustration; the length is 70 centimeters (28 inches) and the weight is 2–3.5 kilograms (4–8 pounds); the nape is gray. (2) The Common Loon or Great Northern Diver (*Gavia immer*). The length is 75 centimeters (30 inches) and the weight is 4 kilograms (9 pounds); the nape and beak are black. (3) The Yellow-Billed Loon or White-Billed Northern Diver (*Gavia adamsii*). The length is 87 centimeters (34 inches) and the weight is 4.5–6.4 kilograms (10–14 pounds). The beak is ivory-colored in adults. (4) The fourth species, with small white stripes on the non-breeding plumage, is the Red-Throated Loon or Diver (*Gavia stellata*). The length is 58 centimeters (23 inches) and the weight is 1–2.4 kilograms (2–5 pounds).

Arctic loon. (*Sketch by Glenn D. Considine.*)

The loon or diver is found in Canada, the northern parts of the United States, particularly the mountain lakes of New York and Pennsylvania and the lakes of Michigan as well as in California and south to Mexico. Some species of loon are found in Europe and essentially in most parts of the world.

LOOP (Computer System). A sequence of instructions that may be executed repeatedly while a certain condition prevails. The productive instructions in the loop generally manipulate the operands, while bookkeeping instructions may modify the productive instructions, and keep count of the number of repetitions. A loop may contain any number of conditions for termination, such as the number of repetitions or the requirement that an operand be non-negative. The equivalent of a loop could be achieved by the technique of straight line coding, whereby the repetition of productive and bookkeeping operations is accomplished by explicitly writing the instructions for each repetition.

See also **Program (Computer).**

LOOP DIURETICS. See **Diuretics.**

LOOP GAIN. In feedback terminology, the gain around the feedback loop, numerically equal to the product of the forward gain by the gain of the feedback network when the circuit configuration permits meaningful identification of these two separate transmissions. The feedback network is also called the *beta-network*.

See also **Gain Magnitude Ratio.**

LOOP (Mathematics). A closed path. An immediate consequence of this definition is that a finite graph contains only a finite number of loops, a conclusion which is critical to the practical application of Kirchoff's law for voltage.

LORENTZ FRAME. Any of the set of coordinate systems in Minkowski space for which the square of the interval between two events is $c^2\,dt^2 - (d\mathbf{r})^2$. Any such coordinate system may be obtained from another by means of a Lorentz transformation (together perhaps, with an orthogonal transformation of the space axes). With each Lorentz frame may be associated a point observer, each of whom moves with constant velocity relative to the others.

LORENTZ INVARIANCE. The equivalence principle of special relativity, which states that physical principles must be invariant under a transformation from one coordinate system to another. Since this is a Lorentz transformation, the invariance itself is often called by the name of Lorentz.

LORENTZ TRANSFORMATION. Relations connecting the space and time coordinates of an event as observed from two Lorentz frames. If S' moves relative to S with velocity v in the x-direction, x, y, z, t denote position and time coordinates of two events as measured by S, and x', y', z', t' the corresponding quantities for S', then

$$x' = \gamma(x - vt)$$
$$y' = y,\ z' = z$$
$$t' = \gamma\left(t - \frac{x}{c^2}\right)$$

where $\gamma = (1 - v^2/c^2)^{-1/2}$. These relations were shown by Einstein (1905) to be a consequence of the special relativity theory. See also **Relativity;** and **Vector.**

LORISOIDS (*Mammalia*). Primitive animals of the order *Primates* and similar to lemurs. See also **Lemur.** Lorises are found in the warmer parts of southeastern Asia and the East Indian islands. They have large staring eyes and from their very slow movements have sometimes been referred to as slow lemurs. The lorises are forest animals of nocturnal habits. The slender loris is known as *Loris gracilis*; the slow loris as *Nycticebus tardigradus*.

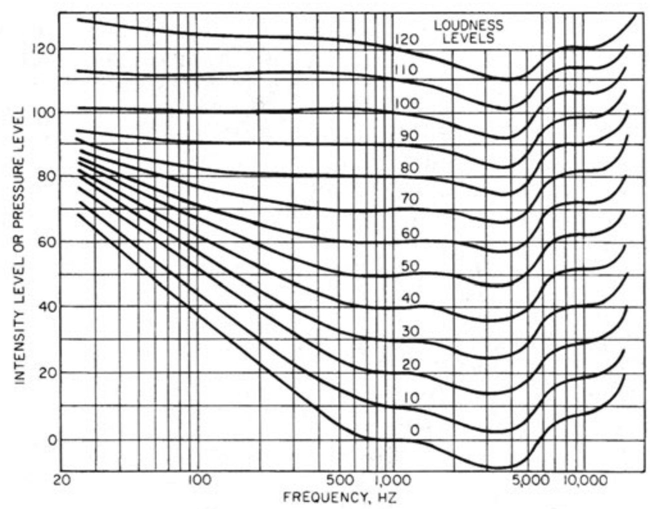
LOSSER. A dielectric material which dissipates energy. A dissipative element placed in a circuit to prevent oscillation.

LOSS FACTOR. The rate at which heat is generated in a dielectric is proportional to its loss factor, which is equal to the product of its dielectric constant by its power factor. Both the dielectric constant and power factor are usually functions of frequency; therefore, the loss factor changes with changing frequency.

LOSS FUNCTION. In decision theory, a function of the decision and the true underlying distributions which expresses the loss incurred in taking that decision. If there are a number of possible situations, the array of losses according to situation and decision is called the *loss matrix*. It is analogous to the payoff matrix of games theory.

LOSS (Transmission). A general term used to denote a decrease in signal power in transmission from one point to another. Usually expressed in decibels.

LOUDNESS LEVEL. The loudness level of a sound is the sound-pressure level of a standard tone (usually 1,000 Hz) which sounds equally as loud as the sound under measurement. In 1933, Fletcher and Munson published their *loudness level contours* for pure tones. See accompanying figure. The curves commonly are referred to as Fletcher-Munson curves or equal-loudness contours. The numbers on the contours are the loudness levels of the sound in phons (the sound-pressure level of a 1,000 Hz tone that is equally loud). Thus, if a certain complex sound wave sounds equally as loud as a 1,000 Hz tone having a sound-pressure level of 60 db re 0.0002 dyne/cm^2, the complex wave is said to have a loudness level of 60 phons regardless of its sound-pressure level. Other sets of equal-loudness contours which deviate in some respects from the Fletcher-Munson curves have been developed since by other investigators.

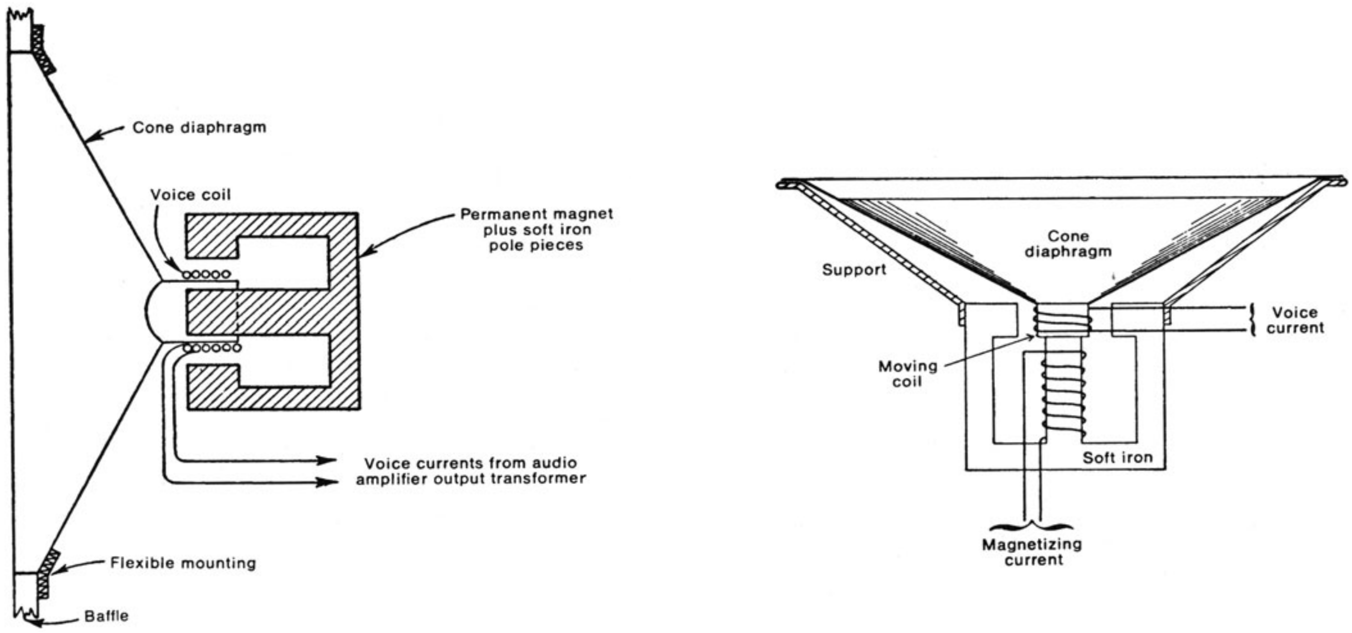


Fletcher-Munson curves indicating equal-loudness contours.

See also **Acoustics.**

LOUDSPEAKER. A transduction device, usually based on the dynamic (moving-coil) principle, that converts electrical energy into mechanical energy or sound. A coil of wire located in the magnetic field of a permanent magnet is attached to a paper cone. See diagram. The cone, at its outer edges, is flexibly attached to a support ring. When an electric current is passed through the coil, a force is created that acts upon the cone. Cone movement generates sound waves that are proportional to the frequency of the exciting current. Loudspeakers usually are low in efficiency, about 5%. By using horns of a gradually increasing cross sectional area, efficiencies of 30 to 50% can be achieved.

Liquid-Phase Projectors. Piezoelectric, magnetostrictive, and electromagnetic are the principal types of transduction used in transmitters

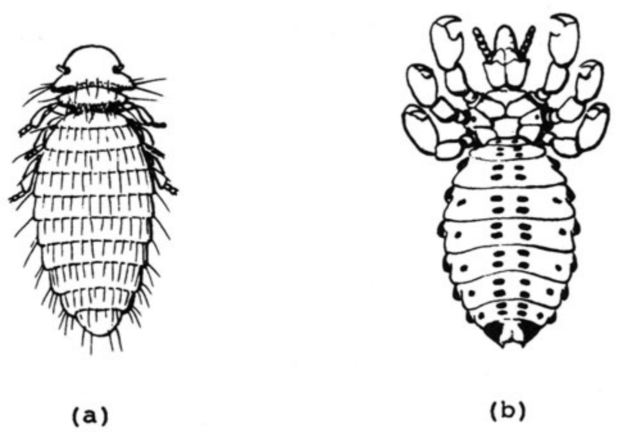


Two common types of loudspeakers: (a) permanent magnet type; (b) electromagnetic type.

to excite acoustical waves in liquids. Designed to be resonant, these devices usually operate at their fundamental frequency. Piezoelectric liquid-phase projectors are effective from 20 kHz to above 100 MHz; while magnetostrictive transducers will handle the range from 10 to 100 kHz. With bandwidth ratios of 5 to 20, both types have efficiencies on the order of 0.5 to 4 watts per square centimeter. Increasing the pressure of the liquid mass and providing special cooling will raise radiation intensities up to 50 watts per square centimeter. Many liquid-phase projectors are made up of arrays of individual transducers to control directivity and to increase power-handling capacity.

A large number of projectors cannot be used as receivers. The latter are designed to excite acoustical vibrations in air and are not reversible. Nonreversible projectors include modulated air-flow speakers, whistles, and sirens. In a modulated air-flow speaker, a valve controlled by an electrical signal modulates the flow of the airstream. Used for public address systems, these speakers have high efficiency and power output, but also have high distortion and poor frequency response. Whistles have high efficiency, but suffer from frequency and amplitude instability unless driven by an auxiliary device, such as the resonant cavity of an organ pipe. A siren, in which a stream of compressed air is chopped by a series of rotating blades, combines high efficiency and high intensities, with stable, easily controlled frequencies.

LOUSE (*Insecta*). An external parasitic insect found on warm-blooded animals, both mammals and birds. Two kinds of lice occur: (1) The true or sucking lice (order *Anoplura*); and (2) the bird or biting lice (order *Mallophaga*). As the titles suggest, the sucking lice have piercing and sucking mouth parts; the bird lice have biting mouth parts. See accompanying illustration.



(a) Biting louse (*Mallophaga*); (b) sucking louse (*Anoplura*.) (*USDA*.)

Livestock Lice

Hog louse (Haematopinus suis or *adventicius*, Linne). Of the order *Anoplura*. This is a wingless, rather large, flat, gray-colored louse, about $\frac{1}{4}$-inch (6 millimeters) long that infests hogs. It does not affect other livestock. It is the largest of the blood-sucking lice. The head and legs are comparatively long. The insect is equipped with a hooklike member for clasping hairs of the host. A favorite habitat is in between folds of the hog's skin. The entire life cycle of the insect occurs on the host. The hog has only two commonly encountered insect parasites—the hog louse and the mange mite. Control is by dipping or spraying with Co-Ral, lindane, malathion, methoxychlor, ronnel, or toxaphene, where available and permitted. Medicated hog wallows, which incorporate a surface film of petroleum or pine oil, can be effective. The louse dies within a few days if not on a host.

Sheep and Goat Lice. There are several species:

Bloodsucking body louse (*Haematopinus orvillus*, Neumann). Of the order *Anoplura*.

Bloodsucking foot louse (*Linognathus pedalis*, Osborn). Of the order *Anoplura*. A pale, rather slender louse, about $\frac{1}{12}$ inch (2 millimeters) long.

Sheep-biting louse (*Boricola or Trichodectes oris*) (Linne). Of the order *Mallophaga*. A pale-brown insect with a red head, about $\frac{1}{20}$ inch (1+ millimeter) in length. This louse eats the wool fibers, causing fibers to become soiled and tangled. Its full life cycle is spent on the host. Treatment for this and other sheep and goat lice mentioned may include dipping or spraying lindane, methoxychlor, rotenone, or toxaphene. All of these lice can be quite damaging to the wool of the animals. The clip of mohair from an Angora goat, for example, may be reduced severely if there is an infestation.

Cattle and Horse Lice. These also affect mules and donkeys.

Long-nosed cattle louse (*Linognathus* or *Haemotopinus vituli*, Linne). Of the order *Anoplura*. This is a red louse, about $\frac{1}{8}$ inch (3 millimeters) long. The insect pierces the skin and sucks blood. In heavy infestations over long periods, animals become emaciated. Patches of skin become bare and sore. The insect seems to prefer unhealthy, poorly fed animals. Control is by spraying or dipping with lindane, malathion, methoxychlor, toxaphene, or ronnel, when available and approved. Special directions must be followed for spraying dairy cows to avoid any contamination of milk.

Horse-sucking louse (*Haemotopinus asini*, Linne or *H. macrocephalus*, Burmeister). Of the order *Anoplura*. This is a medium-size louse, about $\frac{1}{8}$ inch (3 millimeters) long. The bite is very painful. There are several generations per year. Treatment is about the same as for cattle louse.

Poultry Lice. A number of species attack poultry and wild fowl and birds, including pigeons. These lice are of the order *Mallophaga* and they do not suck blood, but rather they feed on bits of skin, scabs, feathers, and other organic debris found on the bird's body. Irritation is caused by the sharp mouth parts and sharp claws. Nibbling of the lice prevents rest and sleep, and causes loss of appetite, diarrhea, droopy wings, leading to progressive emaciation, reduction in egg production, and death of the birds when left unattended. Poultry lice are distributed worldwide. They are wingless, with six legs, flat bodies, and round heads. Treatment is by spraying or dusting with malathion or rotenone, where approved. Nests and litter also should be sprayed. Painting the roosts with 40% nicotine sulfate or 3% malathion can be effective. Some of the species include:

Chicken-head louse (*Cuclotogaster heterographus*, Nitzsch). An insect about $\frac{1}{10}$ inch (2.5 millimeters) long that severely irritates the birds around the neck and head area. They are particularly irritating to young turkeys and chicks.

Chicken-body louse (*Menacanethus stramineus*). An insect about $\frac{1}{8}$ inch (3 millimeters) long that irritates areas under the wings and about the vent. It attacks both young and old birds. Records indicate that over 35,000 of these lice were found on one chicken.

Common body louse (*Menopan gallinae*, Linne). Also called *shaft louse* or *small body louse*, about $\frac{1}{16}$ inch (1.5 millimeters) long. Lives mostly on the feathers and one of the most commonly encountered on fowl.

Fluff louse (*Goniocotes gallinae*, De Geer). Prefers operating in the fluff and under the vent; about $\frac{1}{16}$ inch (1.5 millimeters) long.

Brown chicken louse (*Goniodes dissimilis*, Denny). A medium-size louse, about $\frac{1}{10}$ inch (2.5 millimeters) long.

Large chicken louse (*Goniodes gigas*, Taschenberg). A comparatively large louse, about $\frac{5}{32}$ inch (2.5 millimeters) long.

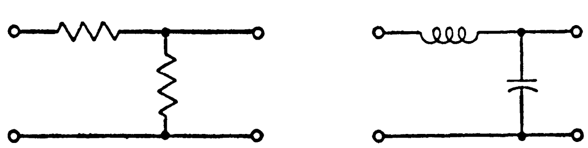
Wing louse (*Liperus caponis*, Linne). Prefers the barbules of the wing feathers; about $\frac{1}{10}$ inch (2.5 mullimeters) long.

Small pigeon louse (*Campanulotes* or *Goniocotes bidentalus*, Burmeister). Inhabits feathers of both young and old pigeons.

Slender pigeon louse (*Columbicola columbae*, Linne). Much like the small pigeon louse.

> NOTE: Plant lice are not true lice, but members of the order *Homoptera* and commonly called aphids.

L-SECTION. This refers to an elementary section of a network such as a filter where the components are connected in the form of an L, i.e., one component in series with one side and the other in shunt across the two sides of the circuit.

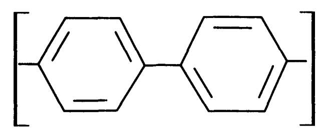


L-section.

LUBRICANT. A material used to diminish friction between the moving surfaces of machine parts; also to decrease friction between a cutting tool and the material being cut. A wide variety of materials is used for manufacturing lubricants. Animal lubricants are obtained from the fat of common animals and can be classified as hard fats (stearin) and soft fats (lard) or naturally occurring combinations. Vegetable lubricants include rape seed oil, cottonseed oil, soybean oil, castor oil, and linseed oil. They range in properties from solid to liquid. Petroleum and mineral oil lubricants, because of their greater stability, are usually preferred for machine applications. Lubricants range from light oils to very heavy solid greases. Graphite, a solid, is also used as a lubricant.

Because of increased requirements for lubricants, including higher temperature and pressure applications, greater durability, and tolerance to wide changes in ambient temperature conditions, numerous synthetic lubricants have been developed. These include synthetic hydrocarbons, carboxylic acid esters, silicones, polyethers (polyalkylene glycols), phosphate esters, silicate esters, highly fluorinated compounds, and polyaromatics (polyphenyls and polyphenyl ethers). In selecting a lubricant, the following characteristics are considered: (1) lubricity and antiwear properties; (2) fluid range; (3) viscosity index; (4) additive response of base oil; (5) oxidation stability; (6) thermal stability; (7) hydrolytic stability; (8) fire resistance; (9) compatibility with petroleum products; (10) compatibility with paints, plastics, and elastomers; and (11) cost.

Over a number of years, the early polyol esters, the formulas of some of which are shown below, appeared to be adequate for coping with the increasing rigorous properties required for increasingly difficult lubrication problems. They continue to be used, but some professionals in the field have developed a number of proprietary formulations that are claimed to be superior to the polyol esters.



Polyaromatics (polyphenyls)

$$R{-}O{-}\overset{O}{\overset{\|}{C}}{-}(CH_2)_n{-}\overset{O}{\overset{\|}{C}}{-}O{-}R$$

Diester

$$\left[-CH_2-\overset{R}{\overset{|}{C}}-O-\right]_n$$

Polyether (polyalkylene gloycols)

$$Si{-}(O{-}R)_4$$

Silicate ester

$$[-CH_2-CH_2-CH_2-]_2$$

Synthetic hydrocarbon

$$C{-}(CH_2{-}O{-}C{-}R)_4$$

Neopentyl polyol ester

More attention has been given to tribology, the scientific discipline of friction, wear, and lubrication. The underlying principles of tribology have been investigated intensively by a number of research groups, including the U.S. Naval Research Laboratory and physicists at the Georgia Institute of Technology, Atlanta, Georgia. Researchers are attempting to develop a better theoretical basis for understanding the processes that occur when two solid bodies move past each other at close to overlapping distances. Most of this new knowledge has stemmed from working models as well as from computer models. Researchers at the Georgia Institute of Technology have used a large Cray computer to predict what occurs when the tip of a thin nickel needle, for example, is pressed repeatedly onto a flat gold surface. The investigators initially employed quantum mechanics for answering such questions as adhesion, cohesion, and the making and breaking of chemical bonds. Researchers at the Naval Research Laboratory have used an atomic force microscope as a tool to check their experiments. The present goal is to model more complex systems. As pointed out by one researcher, "To make progress in the molecular engineering of lubricants, you need to know the molecular details of the process."

LUMINESCENCE. A characteristic nonthermal emission of electromagnetic radiation by a material upon some form of excitation. Some luminescent materials are called phosphors. E. Wiedemann defined the term in 1888 as "all those phenomena of light not solely conditioned by the rise in temperature."

Whereas the output from blackbody radiators consists of broad-band emissions which follow the Stefan-Boltzmann temperature relationships, luminescence emission from phosphors consists of relatively narrow bands, which do not follow the blackbody laws. Thus, light emission due solely to the temperature of a source is referred to as *incandescence*, while *luminescence*, unlike incandescence, is a function of the specific material involved. Although *fluorescence* and *phosphorescence* are sometimes used synonymously with luminescence, a more rigid definition of *fluorescence* would be luminescence having a persistence (afterglow) shorter than about 10^{-8} second, with *phosphorescence* being longer than 10^{-8} second.

The luminescence process itself involves (1) absorption of energy; (2) excitation; and (3) emission of energy, usually in the form of radiation in the visible portion of the spectum. The *type* of luminescence is usually defined by the excitation means, i.e., *cathodo*luminescence where excitation is by cathode rays, as in a television kinescope. The most commonly encountered types of luminescence are listed in the accompanying table.

TYPES OF LUMINESCENCE

Luminescence Type	Excitation Source	Example
Photoluminescence	Photons	$ZnS \cdot Ag$
Cathodoluminescence	Cathode Rays	$Zn_2SiO_4 \cdot Mn$
Electroluminescence	Electric Fields	$Zn \cdot (S \cdot Se) \cdot Cu$
Chemiluminescence	Chemical Reactions	Oxidation of Luminol
Bioluminescence	Biochemical Reactions	Luciferin
Triboluminescence	Mechanical Disruption	$ZnS \cdot Mn$

The luminescent material may be considered as a transformer of energy, i.e., from ultraviolet photons to photons of lower energy; from cathode rays to photons; from electric fields to photons, etc. An inorganic luminescent material, or phosphor, usually consists of a crystalline host material to which is added a trace of an impurity (activator and coactivator).

Chemiluminescence

Numerous chemical reactions produce heat, but relatively few release their energy as light. This latter phenomenon is termed *chemiluminescence*. Representative reactions that release light are shown in Fig. 1. As pointed out by A. K. Campbell (University of Wales College of Medicine, Cardiff), "Absorption of energy by an atom or molecule raises an electron to a higher energy level. This is known as an 'excited state' and is inherently unstable. When the electron drops back to its ground state the energy must either be transferred to another atom or molecule, be released as heat or be emitted as light. The decay of the electron to ground state is very fast, occurring within 1–10 nanoseconds ($10^{-9} - 10^{-8}$ second)."

In chemiluminescence, the chemical reaction raises an electron to a higher level, which then decays back to the ground state, releasing a photon of light, the energy of which is predictable by Einstein's equation. When the energy drop is large, the light is blue; if small, the light is red. Inasmuch as the electronic excitation-decay process is extremely fast, the intensity of light in chemiluminescence is determined by the kinetics of the chemical reaction.

The distinction of chemiluminescence from fluorescence results from two factors:

1. When an atom or molecule fluoresces, it remains chemically unchanged and can be immediately be re-excited once light emission has occurred.
2. In chemiluminescence, each molecule only reacts once to form an excited state, while the excited product (actual emitter) has a different chemical structure from the initial substrate.

See also **Bioluminescence.**

Light-Emitting Diode (LED)

Recombination of injection electroluminescence was first observed in 1923 by Lossew, who found that when point electrodes were placed on certain silicon carbide crystals and current passed through them, light was often emitted. Explanation of this emission has been possible only with the development of semiconductor theory. If minority charge carriers are injected into a semiconductor, i.e., electrons are injected into *p*-type material or "positive holes" into *n*-type material, they recombine spontaneously with the majority carriers existing in the material. If some of these recombinations result in the emission of radiation, electroluminescence results. Minority-carrier injection may occur not only at point contacts, but also at broad area rectifying junctions; in this case, the junction must be biased in the forward or "easy flow" direction, and the electric field in the junction is lower when the voltage is applied than in its absence. This type of emission has been observed in several materials, including SiC, diamond, Si, Ge, CdS, ZnS, ZnSe, ZnO, and some of the so-called III-V compounds, such as AlN, GaSb, GaAs, GaP, InP, and InSb. The emission of many of these materials lies in the infrared region of the spectrum. For radiation in the visible region (instead of the infrared), the energy difference between the holes and electrons (band gap of the semiconductor) must be more than 1.8 eV. Numerous materials satisfy this requirement, notably those used for cathode-ray tube phosphors, but the materials present difficulties in fabricating *p-n* junctions and thus are not candidates for light-emitting diodes.

The list of materials for LEDs includes GaP, GaAsP, GaAlAs, GaN, and SiC. The two materials of choice to date have been GaP and GaAsP. Early commercial LEDs were made from $GaAs_{0.6}P_{0.4}$ deposited epitaxially as a thin layer on a GaAs crystal substrate. With these, *p-n* junctions were made, using diffusion techniques similar to those used in making silicon diodes. The band gap is 1.92 eV. There is an emission band of red light with a peak at about 650 nm, resulting from direct recombination of electrons and holes.

GaAsP has a high index of refraction, and consequently only light emitted toward the surface (4%) is usable—the remainder is reflected back. A diode can be encapsulated in epoxy material to take on the shape and form of a lens. These diodes are particularly effective where a number are fabricated in close proximity on a single-crystal chip.

Diodes that emit light in shorter wavelengths (green, yellow, etc.) can be made by increasing the phosphorus content, but only up to about 40% because of rapid decrease in efficiency. Efficiency can be increased by incorporating nitrogen atoms into crystals. The N atoms act as isoelectronic centers, trapping electron-hole pairs in an excited state. Three types of nitrogen-doped diodes have gained some importance: $GaAs_{0.65}P_{0.35}$ (orange light); $GaAs_{0.85}P_{0.15}$ (yellow light); and GaP (green light). Zinc and oxygen doping are also used. Diodes operate more efficiently if driven with periodic pulses of high current rather than with constant current. The short response time of junction diodes to current pulses (a fraction of a microsecond) and their rectifying property (they block current flow in the reverse, nonemitting direction) combine to make the diodes a good choice for *X-Y* addressing arrangements.

LUMINOSITY FUNCTION. Because of the variable sensitivity of the human eye to radiation of different wavelengths, a standard function has been established. For the standard conditions chosen in establishing this standard luminosity function (photopic vision), the luminously effective radiant intensity in lumens of radiation of spectral energy distri-

$$680 \int_{\lambda=0}^{\lambda=\infty} y_\lambda J_\lambda \, d\lambda$$

bution J_λ watts/unit wavelength is given by
where y_λ is the standard luminosity function normalized to a value of unity at 555 nanometers. The numerical values for y_λ are commonly given as a luminosity curve.

For very low levels of intensity (scotopic vision) the peak of the luminosity function curve shifts toward the violet for young eyes (507 nanometers) with an absolute value of 1,746 lumens/watt.

LUMINOUS COEFFICIENT. A coefficient which measures the integrated fraction of the radiant power that contributes to its luminous properties as evaluated by means of the standard luminosity function.

Luminous coefficient

$$= \int_{\lambda=0}^{\lambda=\infty} y_\lambda J_\lambda \, d\lambda \Big/ \int_{\lambda=0}^{\lambda=\infty} J_\lambda \, d\lambda$$

where y_λ is the standard luminosity function and J_λ is the spectral energy distribution of the radiant intensity. The luminous coefficient is unity for a narrow band of wavelengths at 555 nanometers.

LUNAR ECLIPSE. See **Eclipse.**

LUNG CANCER. See **Cancer and Oncology.**

LUNGFISHES (*Osteichthyes*). Of the order *Dipneusti*, the lungfish has an air bladder opening from the pharynx which can be filled with air gulped through the mouth and serving as a lung. It is the only known species of fish that can live out of water for a period as much as four years. The dipnoids live in transient streams and swamps of Australia, South America, and South Africa. Some varieties pass the dry season in cells which they form in the muddy bottom as the water dries up. They resemble amphibians in some details of structure and are reminiscent of creatures that existed during Devonian times. See also **Fossil Fishes.** Probably the most primitive species is the lungfish of Australia (*Neoceratodus forsteri*). Originally, the fish was found only in the Burnett and Mary rivers (northeastern Australia). However, it has been transported quite successfully to several lakes in Queensland where it thrives.

The average lungfish measures up to $3\frac{1}{2}$ feet (1 meter) in length when grown, although specimens up to 6 feet in length and weight of 100 pounds (45.4 kilograms) have been recorded. Possibly the largest recorded specimen was an African lungfish (*Protopterus aethiopicus*), measuring some 7 feet (2.1 meters) and found in Lake Victoria. See accompanying illustration.

African lungfish.

It is interesting to note that similarities between the African and South American lungfishes have contributed to the hypothesis that there may have been a land connection between South America and Africa at an earlier period. See also **Earthquakes, Seismology, and Plate Tectonics.**

Lungfishes form mud-ball cocoons within which they encase themselves. This is a mucous cocoon which becomes quite leathery after hardening. During estivation in their cocoons, lungfishes lose both weight and length, factors which are quickly recovered in about a month after coming out of the cocoon.

LUPUS. See **Systemic Lupus Erythematosus.**

LUPUS (the wolf). A southern constellation located near Libra.

LUSTER. This term is used by mineralogists to describe the appearance of the surface of a mineral, usually a crystal face, in reflected light. The principal types of luster are: metallic, adamantine, vitreous, resinous, greasy, pearly. The degrees of luster may be defined as: splendid, shining, glistening, or dull. Schillerization is a peculiar form of submetallic luster observed in different directions in certain minerals such as schillerspar, diallage, hypersthene, etc.

LUTETIUM. Chemical element symbol Lu, at. no. 71, at. wt. 174.97, fourteenth in the Lanthanide Series in the periodic table, mp. 1,663°C, bp 3402°C, density 9.842 g/cm^3 (20°C). Elemental lutetium has a close-packed hexagonal crystal structure at 25°C. The pure metallic lutetium is silver-gray in color and retains its luster at room temperature indefinitely. Although experimental observations of the element remain limited (workability, alloying behavior, etc.), extrapolations of known data do not forecast any anomalies in the chemical, mechanical, or physical properties of the element as compared with the other elements in the Lanthanide Series. There are two natural isotopes ^{175}Lu and ^{176}Lu. The latter isotope is radioactive with a half-life of 2.2×10^{10} years. Fourteen artificial isotopes are known. The element was first identified by G. Urban in 1907 and independently by C. A. von Welsbach in 1908. Although not investigated fully, lutetium is classified with a low acute-toxicity rating. Lutetium is the least abundant of the Lanthanide elements, estimated as present on the average of 0.5 ppm in the earth's crust. Potentially, however, it is more plentiful than mercury, cadmium, or any of the precious metals. Electronic configuration

$$1s^2 2s^2 2p^6 3s^2 2p^6 3d^{10} 4s^2 4p^6 4d^{10} 4f^{14} 5s^2 5p^6 5d^1 6s^2.$$

Ionic radius Lu^{3+} 0.848Å. Metallic radius 1.735Å. Other important physical properties of lutetium are given under **Rare-Earth Elements and Metals.**

The source of lutetium to date has been the processing of the other heavy rare-earth metals. Because of very limited availability, little research was conducted on lutetium until the mid-1960s. Most of these studies now are concentrating on prospective uses in phosphors, semiconductor, and other electronic circuitry components. A lutetium dithalocyanine complex has received much consideration recently for application in large, thin screens for television projection.

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.

LUTH (*Reptilia, Chelonia*). Also called the leathery turtle. A large marine turtle, *Dermochelys coriacea*, which reaches a length of 6 feet (1.8 meters). Its carapace is formed of bony plates connected together but not joined to the spinal column or ribs. Its flesh is not palatable.

LYASES. See **Enzyme.**

LYCOPSIDA. A group of plants (Club Mosses) that contains about 500 species, most of which are included in two genera, *Lycopodium* and *Selaginella*. The species of *Lycopodium* are trailing plants often called ground pines, or ground hemlock, as well as club mosses. Many of them are common plants of dry open places in the temperate zone. The plants have long creeping stems growing on the surface of the ground or several inches beneath it. From this prostrate stem short dichotomously branched roots extend down into the ground, and erect branches grow upward. The stems are covered with many small, pointed, dark green leaves. In the more primitive species the reproductive structures or sporangia are found in the axils of ordinary leaves. In other species the sporangia are borne in the axils of modified leaves which are aggregated at the tip of an erect branch, forming a slender cone or strobilus. The many spores borne within the sporangia are all alike and for this reason Lycopodium species are said to be homosporous. The spores are disseminated by the wind, and in time develop into gametophytes. The gametophytes of *Lycopodium* species are extremely small tuberous bodies which grow slowly and reach maturity only if they are invaded by an endophytic fungus. Generally the gametophyte or prothallus develops underground. In the upper surface of the prothallus both antheridia and archegonia are found. Each antheridium contains many straight, biciliate sperms. These swim to the egg, with which one unites, forming a zygote. From this the new sporophyte develops. At first the sporophyte depends on the gametophyte for its food substances. Thus it obtains nutriment by means of a special absorbing structure called a suspensor which grows into the tissue of the gametophyte.

Lycopodium plants are widely used as material from which to make Christmas wreaths. For this purpose the entire plant is often ripped from the ground. The spores of *Lycopodium* are also gathered and sold under the name of Lycopodium powder. Formerly these spores were used in making explosive mixtures and for flashlights.

The genus *Selaginella*, containing some 400 species, is most abundant in the tropics. A few species of small plants are found in temperate regions. In the tropics there are both terrestrial and epiphytic species. The general habit of the plant is much like that of *Lycopodium*. The sporangia are formed in the axils of leaves at the tips of the branches, forming terminal strobili or cones. At the base of the sporophylls there is also a small scale, called a ligule, of unknown function. The sporangia are of two kinds, one, a megasporangium, containing four large spores, called megaspores; the other a microsporangium containing many small spores or microspores. Species of *Selaginella* are therefore heterosporous, a character which distinguishes them from *Lycopodium* species.

LYME DISEASE. A clustering of arthritis cases with distinctive skin lesions, observed in Lyme in southeastern Connecticut in 1975, led to the recognition of a new tick-borne disease. Since its first discovery, more than 500 cases have been recounted in the northeastern United States, as well as Wisconsin, Minnesota, Oregon, and California. Cases also have been reported in Europe and Australia.

Spirochetes have been isolated from patients suffering with the syndrome and have been shown to be transmitted to humans by bites of the ticks *Ixodes dammini* and *Amblyomma americanus*, both of which are known to infest the whitefooted mouse (*Peromysus leucopus*) and the white-tailed deer (*Odocoileus virginianus*).

Lyme disease typically occurs in adults in the summer and early fall; 3 to 20 days following a tick bite, a red macule or papule appears at the bite site on the thigh, buttocks or trunk. It then expands to a large, annular erythematous lesion (*erythema chronicum migrans*) about 6 to 16 centimeters or larger in diameter. The lesion sometimes shows central clearing and secondary concentric rings may develop within the original ring. The lesions are not pruritic and may be multiple.

Associated symptoms include malaise, fatigue, chills, fever (38–39.5°C; 100.4–103.1°F), headache, and nuchal stiffness. In most patients, the skin lesion and symptoms fade in 3 to 4 weeks and recovery is complete. In some patients (about 25%) further immunologically medicated responses develop—arthritis and neurological abnormalities.

Tetracycline therapy is the treatment of choice, followed by penicillin or erythromicin with nonsteroidal antiinflammatory agents directed at Lyme arthritis.

R. C. V.

LYMPH. A clear fluid which circulates in the tissue spaces of vertebrates and passes into the venous system by way of a tubular lymphatic system. It is derived from the liquid plasma of the blood but is more watery and contains no red corpuscles. It serves as an intermediary between the blood itself and the tissues of the body.

Lymph is found lying free in the serous sac cavities of the body, i.e., the peritoneum, pleura, and the spaces in the brain filled with cerebrospinal fluid, which may also be classified as lymph, although it differs in composition from the fluid found in the lymph vessels.

Lymph is derived from the plasma of the blood either by filtration, diffusion, or osmosis through the capillary walls or by active secretion of endothelial cells making up capillary walls. Its composition is very similar to that of the blood plasma.

The function of lymph is to bring nourishment to the tissue cells and to return waste matter and other toxic material to the blood stream by way of the lymphatic vessels, or directly into the blood stream through the capillary walls. Lymph has been compared with the body fluids of invertebrates, commonly designated as hydrolymph and hemolymph.

The small tubules of the lymphatic system resemble capillaries and the larger ducts, called lymphatics, are similar to veins, although of a more delicate structure in relation to their size. Like veins, they have valves which aid in promoting flow through the movements of the surrounding muscles. They are irregular in diameter, forming reservoirs at some points and dilating in the amphibians, reptiles, and birds to form lymph hearts whose pulsations propel the lymph toward the heart. The smaller vessels converge like blood vessels to form larger trunks. In humans, the chief vessels are the thoracic duct, and the right lymphatic duct, which empty into the large veins at the sides of the neck. Along the course of the lymph vessels are groups of lymph nodes. They serve as filters which localize and retard the spread of toxic and infectious elements that are being returned to the blood stream. The lymph nodes also serve as centers for formation of lymphocytes which form one of the main divisions of blood cells.

LYMPHOGRANULOMA VENEREUM. Caused by strains of *Chlamydia trachomatis*, this is a venereal disease which features the appearance of a transitory primary genital lesion, a vesicle or a papule, followed by other stages. In the male, the primary genital lesion, which occurs within 5 to 21 days after the implicating sexual exposure, usually takes the form of a painless vesicle or papule, or as a chancriform lesion (*chancroid*). In the male, it is commonly located on the coronal sulcus of the penis; in females, on the labia or posterior vaginal wall. The situation is usually self-limiting and healing occurs within a few days. Extragenital infections occur in persons who deviate from normal sex profiles. The infection may continue in a second and a third stage, where the disease is extended to include regional lymph nodes. These nodes enlarge, produce pus, and ultimately form buboes. In a third stage, lymphatic obstruction may occur. Lymphatic obstruction may infrequently produce dilation of lymphatic channels and hypertrophy of subcutaneous connective tissue and of skin (*genital elephantiasis*).

Treatment may include oral tetracycline or sulfonamides, such as sulfisoxazole. It is a good practice to test all patients with lymphogranuloma venereum for syphilis because the two infections are frequent companions.

LYMPHOKINES. These are soluble substances (factors) produced by lymphocytes which aid in regulating a variety of immune responses. See also **Immune System and Immunology.** They are not immunoglobulins, but are synthesized by lymphocytes of undetermined structure and are classified according to the target cells they affect.

Chemotactic factors are lymphokines which attract certain cells, e.g., monocytes, neutrophils, etc., to a particular site. They are synthesized and released within 24 hours of lymphocyte stimulation.

Migration inhibition factor (MIF), the first lymphokine discovered, inhibits macrophage migration and is produced by certain sensitized lymphocytes stimulated by an exquisite antigen. Mitogens and antigen-antibody complexes can, however, nonspecifically trigger lymphocytes to produce MIF and antigen need not be present for MIF to inhibit macrophage migration.

Macrophage-activating factor (MAF) induces morphologic, metabolic, and functional changes in macrophages which enhance the cell's ability to kill microorganisms and tumor cells.

Leucocyte-inhibitory factor (LIF) inhibits neutrophil (but not monocyte) migration in vitro.

Interleukin. Macrophages produce a *monokine* termed Interleukin-1 (lymphocyte activating factor) which combines with an antigen or mitogen to stimulate T lymphcytes to produce Interleukin-2 (T-cell growth factor) which in turn causes the proliferation of other T cells, such as helper, suppressor, or cytotoxic cells.

Cytotoxic and Cystostatic Factors. Lymphocytes from several species produce lymphokines which kill or inhibit the growth of susceptible target cells. Among these lymphokines are several proliferation-inhibiting or colony-inhibiting factors and inhibitors of DNA synthesis.

Tissue Factor. Lymphocytes stimulated in vitro by an antigen or mitogen produce a procoagulant material which is biologically similar to tissue thromboplastin, but is antigenically distinct from it. The pathophysiologic importance of tissue factor is undetermined.

Lymphokines which enhance antibody production are termed helper factors and are produced by helper T lymphocytes. Lymphokines that suppress antibody production are termed suppressor factors and are produced by suppressor T lymphocytes.

Colony-stimulating activity is a lymphokine which induces differentiation of bone marrow cells in vitro into granulocytes and agranular mononuclear cells.

Osteoclast activating factor (OAF) is a lymphokine which causes bone resorption by activating osteoclasts.

Lymph node lymphocytes from rats immunized with myelin basic protein release immunoglobulin-binding factor which combines with antigen-antibody complexes causing IgG-sensitized sheep erythrocytes to agglutinate.

Interferons. These are a family of glycoproteins which act against a wide range of viruses. These types are known and can be distinguished on the basis of their physiocochemical characteristics, the cell types producing them, and the types of stimuli causing their production. They do not however act directly upon viruses, but render cells resistant to viral infection.

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LYMPHOMA. A lymphoma is defined as a tumor of lymphoid tissue. Some diseases are also referred to as lymphomas and comprise a group of malignant disorders that usually arise in the lymph nodes. These diseases present a broad spectrum of clinical features. Sometimes the term non-Hodgkin's lymphomas is used to distinguish other lymphomas from Hodgkin's disease because they do not have the characteristic giant cells of Hodgkin's disease. The causes of the lymphomas remains poorly understood. One suspected cause is viral infection. The epidemiology of the lymphomas may provide evidence as to the etiology of the diseases. Some lymphomas occur with much higher frequency in some geographical regions than in others. Diagnosis of a lymphoma is by biopsy. Treatment of the non-Hodgkin's disease lymphomas tends to parallel the therapy used in treating Hodgkin's disease. See **Hodgkin's Disease;** and **Lymphosarcoma.**

LYMPHOSARCOMA. A disease, similar to leukemia and Hodgkin's disease, in its symptoms. Although still generally used, the term lymphosarcoma is pathologically obsolete; it basically refers to what are now known as lymphocytic lymphomas and includes what were formerly known as lymphocytic lymphosarcomas, the so-called lymphoblastic lymphosarcoma, follicular lymphoma, and Burkitt's tumor.

It is important to realize that the distinction between chronic lymphocytic leukemia and those cases of lymphocytic lymphoma in which some of the tumor cells enter the circulating blood stream is not made easily and indeed may be impossible.

The lymphocytic lymphomas may arise in any group of lymph nodes in a tonsil, the spleen, or thymus, but lymph nodes are by far the most common site of their origin. The invasive growth of the sarcoma leads to fusion and coalescence of affected nodes and their adherence to adjacent tissues—these latter may be deeply penetrated by the tumor. Microscopically the normal structure of the affected tissue is obliterated by masses of closely packed tumor cells and in many cases the cells cannot be distinguished from normal lymphocytes. The cells of any given tumor are, however, usually uniform in type—either small or large. Occasionally a well differentiated tumor that consists predominantly of small lymphocytes may be found to include areas in which the cells are larger lymphocytes; more rarely, pleomorphic areas are seen. Whatever the type of cell, the lymphomas with mixed cytological constitution trend to be more rapidly progressive. An early indication of developing malignancy by cell differentiation is enlargement of the nucleolus. There is seldom any tendency for necrosis to occur in the lymphomas unless they have become anaplastic.

Acute lymphatic leukemia of childhood may be lymphosarcoma with the abnormal cells overflowing into the circulating blood. Whereas usually this disease is fatal if untreated, when patients receive early and proper treatment, both blood and bone marrow may revert to normal condition with full restoration of health for variable periods. In lymphosarcoma, x-ray therapy is the major and most universally effective form of treatment. Chlorambucil and nitrogen mustard have been effective chemotherapeutic agents in earlier stages of the disease; cyclophosphamide and vinblastine sulfate, in later stages. Prednisone may prove beneficial in patients with fever and hematologic disturbances no longer suitable for treatment with x-ray or other drugs.

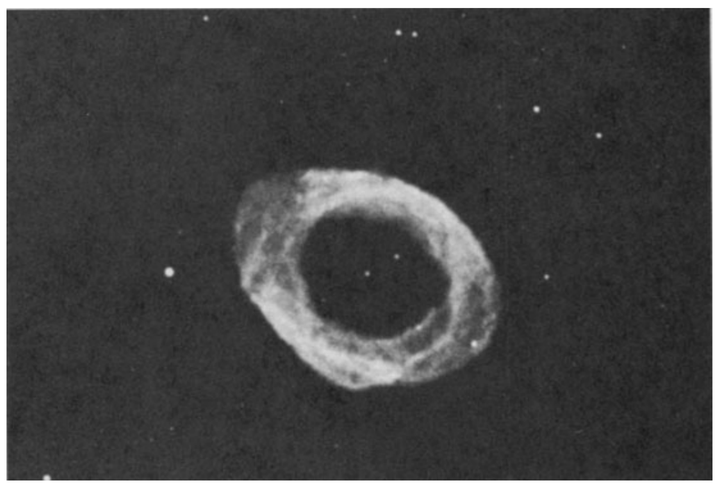
R. C. V.

LYNX. See **Cats.**

LYOPHILIC SOL. See **Colloid System.**

LYOPHOBIC SOL. See **Colloid System.**

LYRA (the harp). One of the small constellations, which contains a number of most interesting objects for viewing through a small or large telescope. Lyra is most easily distinguished by an equilateral triangle having the star Vega at one of its apexes. This star is the brightest in the northern celestial hemisphere. It lies almost in the direction in which the sun and all the planets are moving, due to solar motion, and although it is a long way from the pole of rotation of the celestial sphere at present, it will be the pole star about 12,000 years hence, due to precession. The star Epsilon Lyrae is one of the most famous multiple stars in the entire sky. It can be resolved into two components through a field glass and, on a clear night, into four components through a 6-inch ($\approx$15-centimeter) telescope. Also to be found in this constellation are several other double stars, and the famous ring nebula, an interesting object when viewed through a 6-inch telescope. (See accompanying figure, and map accompanying entry on **Constellations.**)



Ring nebula in Lyra. This nebula is gradually expanding, and its edges are quite red. (*Lick Observatory.*)

LYRE BIRD (*Aves, Passeriformes.*) This bird ranges from Queensland to Victoria and is highly regarded in Australia, sometimes pictured on stamps and official seals. The male lyre bird is well known for its display of plumage and performance to attract females of the species. The "lyre frame" feathers are something like those of a bird of paradise or peacock when displayed. Of several species, there is the superb lyre bird (*Menura novaehollandiae*) and Albert's lyre bird (*M. alberti*). The tail feathers range from $1\frac{1}{2}$ to $2\frac{1}{2}$ feet (0.5 to 0.8 meter) in length and usually are white and brown. The head is small, the legs are long, the claws are strong. The male is known for its incredible mimicking of other birds. There is one pale-purple, thick-shelled egg of about the size of a chicken egg. The incubation period is 6 weeks. The young do not leave the nest for 12 weeks after hatching.

LYRIDS. A name given to certain meteor showers that are observed about April 20 of each year. The orbit of the radiant point was definitely associated with the orbit of comet 1861 I. by Weiss. Records of showers from this radiant are found as far back as 687 B.C. A report written by the Chinese in 15 B.C. indicates that during the Lyrid shower of that year, "after the middle of the night, stars fell like rain." Several other accounts of striking showers during April are on record; in particular, we find many newspaper accounts of the Lyrid shower of 1803, which was observed over the United States from North Carolina to New Hampshire. The Richmond, Va., *Gazette* of April 23, 1803, gives a long and vivid account of the shower occurring on the morning of April 20, stating that "from one until three those starry meteors seemed to fall from every point of the heavens, in such numbers as to resemble a shower of sky rockets."

A few scattered members of this shower are observed coming from the radiant point in the constellation of Lyra every year, but there has not been any very striking display since 1803. Because there have been striking showers in the past, the assumption is that a large swarm of meteors exists at some undetermined point along the orbit, and that we may be treated to another brilliant display during some April in the future.

LYSIS (Bacteriology). The dissolution of cells, e.g., bacteria, or red blood cells; their breakdown from structural form to a structureless fluid.

LYSIS (Physiology). The gradual decline of the symptoms of disease, referring especially to the gradual abatement of fever.

LYSOSOMES. Subcellular organelles which are believed to contain digestive enzymes capable of breaking down many of the cellular constituents. Disruption of the lysosomes and liberation of these enzymes may occur under certain conditions and can lead to lysis of the cell. See also **Cell (Biology).**