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Editorial Address:
448, Pocket V, Mayur Vihar, Phase - I
Delhi - 110091
Phone: 49050844, 22053916, 47541851, 64683387
e-mail: editorialindia@gmail.com

Editorial

Science and technology have profoundly influenced the course of human civilization. Science has provided us remarkable insights into the world we live in. The scientific revolutions of the 20th century have led to many technologies, which promise to herald wholly new eras in many fields. As we stand today at the beginning of a new century, we have to ensure fullest use of these developments for the well being of our people. Science and technology have been an integral part of Indian civilization and culture over the past several millennia. Few are aware that India was the fountainhead of important foundational scientific developments and approaches. These cover many great scientific discoveries and technological achievements in mathematics, astronomy, architecture, chemistry, metallurgy, medicine, natural philosophy and other areas. A great deal of this travelled outwards from India. Equally, India also assimilated scientific ideas and techniques from elsewhere, with open-mindedness and a rational attitude characteristic of a scientific ethos. India's traditions have been founded on the principles of universal harmony, respect for all creation and an integrated holistic approach. This background is likely to provide valuable insights for future scientific advances. During the century prior to Independence, there was an awakening of modern science in India through the efforts of a number of outstanding scientists. They were responsible for great scientific advances of the highest international caliber.

Science in India is on the move in a big way. The new government should initiate multibillion dollar investments to kick start research, education, and innovation over the next five years. Though several challenging issues remain for the country, India's best and brightest expats living in the United States and Europe are being enticed back to 'Mother India' with the promise of world-class research infrastructure and solid funding. Both challenges and hope lie ahead for India. Other often neglected but important science-related issues to address include establishing university curricula to improve the ability of young students to communicate in English, especially technical writing; the introduction of coordinated proactive strategies by research institutes to improve the 'visibility' of their scientists; incentives and financial support for entrepreneurial scientists to set up companies to commercialize ideas; and changes in labour laws to enable universities to hire qualified scientists irrespective of nationality.

—Editors

Contents

On Generalized Existence Theorem for Projective Limit

Amit Prakash

Research Scholar, Department of Mathematics, J.P.U., Chapra

Abstract

In this article we introduce the notion of generalized existence theorem for projective limit of sequence space. Also we define normal sequence space, section of space. We also give some theorem related to it.

Key words : Co-ordinate convergence, projective convergence, projective limit, Normal sequence space, section of space.

AMS classification No……[46A45]

1. Introduction

The notion of generalized existence theorem for projective limit is pretty old. It is found in cooke [1].

Let E be a sequence space, and let $\left(x^{(n)}\right)$ be a sequence in E, where $x^{(n)} = \left(x_k^n\right)_{k=1}^{\infty}$ for each $n \in N$. The sequence of points $x^{(n)}$ is said to be coordinate convergent (*c–cgt*) if $\lim_{n\to\infty} x_k^n = x_k$ exists for every $k \in N$. The point $x = (x_k)$ is called the co-ordinate limit (*c*-limit) of $\left(x^{(n)}\right)$ and we write $c - \lim x^{(n)} = x$.

Throughout $w, c, c_0, l_p, l_\infty, v, \phi$ denote the space of all, convergent, null, p-absolutely summable, bounded, convergent series, and finite sequences respectively. The notion of generalized kothe-Toeplitz dual i.e. η-dual of sequence spaces was introduced by Chandra and Tripathy [2] as follows :

Let E be a sequence space and $1 \le r < \infty$, then the -dual of E is defined as.

$$E^{\eta} = \left\{ (x_k) \in \omega : \sum_{k=1}^{\infty} | x_k y_k |^r < \infty \text{ for every } (y_k) \in E \right\}$$

For $r = 1$, it reduces to the usual kothe-Toeplitz dual or the α-dual of E.

Let $\phi \leq E \leq F^{\alpha}$. If the projections of the sequence $\left(x^{(n)}\right)$ in F on every fixed direction in E are convergent i.e. if the sequence $u'_n = \sum_{k=1}^{\infty} x_k^{(n)} y_k$ converges for every y in E, we say that $\left(x^{(n)}\right)$ is projective convergent (p – cgt) relative to E or FE – cgt. When $E = F^{\alpha}$, we say that $\left(x^{(n)}\right)$ is p–cgt in F or F–cgt.

2. Definitions and Preliminaries

A sequence (x_k) in F or outside F, is called the projective limit (p-limit) of $\left(x^{(n)}\right)$ in F relative to E or $F\ E - \lim x^{(n)}$, when

(i) $\sum_{k=1}^{\infty} u_k x_k$ is absolutely convergent for every u in E and

(ii) $\lim_{n\to\infty} \sum_{k=1}^{\infty} x_k^{(n)} u_k = \sum_{k=1}^{\infty} x_k u_k$ for every u in E.

By v_r we denote the space of all sequences (x_k) such that the series $\sum_{k=1}^{\infty} x_k^r$ converges, where $1 \leq r < \infty$. Now we introduce some definitions generalizing some previous definitions. Throughout we avoid cases when r is an even integer. So it includes the case when $r = 1$.

Let $\phi \leq E \leq F^{\eta}$. If the projections of the sequence $\left(x^{(n)}\right)$ in F in every fixed direction converges in the sense of v_r i.e. the sequence $u'_n = \sum_{k=1}^{\infty} \left[x_k^{(n)} u_k \right]^r$ converges for every u in E, we say that $\left(x^{(n)}\right)$ is r-projective convergent (r–p–cgt) relative to E. when $E = F^{\eta}$, we say that $\left(x^{(n)}\right)$ is r–p–cgt in F.

A sequence (x_k) in F or outside F is called the r-projective limit (r-p-limit) of $\left(x^{(n)}\right)$ in F relative to E, when

(i) $(x_k u_k) \in l_r$ for every $(u_k) \in E$ i.e. $x \in E^{\eta}$

(ii) $\lim_{n\to\infty} \sum_{k=1}^{\infty} \left[x_k^{(n)} u_k \right]^r = \sum_{k=1}^{\infty} \left(x_k u_k \right)^r$ for (u_k) in E.

When $E = F^{\eta}, x$ is called the $r-p-$limit of $\left(x^{(n)}\right)$ in F, or $r - F - \lim x^{(n)}$

Definition

A sequence space E is said to be normal, if whenever x is in E and $|y_k| \le |x_k|$ for every k, then y is in E.

Examples

σ and c are not normal while ϕ and c_0 are normal As given in cooke (10.2.I) [1] it can be shown that

(i) A necessary and sufficient condition for the r-FE- convergence of $x^{(n)}$ in F is that to every u in E and to every $\in > 0$, there corresponds a positive number $N(\in, u)$ such that for every $p, q \ge N$,

$$\left| \sum_{k=1}^{\infty} u_k^r \left[\left\{ x_k^{(p)} \right\}^r - \left\{ x_k^{(q)} \right\}^r \right] \right| \le \in$$

(ii) When E is normal, the necessary and sufficient condition that $x^{(n)}$ in F should be r-FE-cgt is that to every u in E, and to every $\in > 0$, there corresponds a positive number $N(\in, u)$ such that for every $p, q \ge N$,

$$\sum_{k=1}^{\infty} \left| u_k^r \left[\left\{ x_k^{(p)} \right\}^r - \left\{ x_k^{(q)} \right\} \right] \right| \le \in$$

We have seen before that $x^{(n)}$ may be r-FE-cgt and its coordinate limit x may not satisfy the conditions necessary for r-FE-limit. But in the examples given before σ and c are not normal. When E is normal we obtain the following existence theorem for r-FE-limits.

Theorem-1

When E is normal, the c-limit of every r-FE-cgt sequence is the r-FE-limit of that sequence.

Proof

In analogy to cooke (10.2.I) [1] we have in case $x^{(n)}$ in F is r-FE-cgt sequence is that given any u in E and any $\in > 0$, there corresponds a positive number $N(\in, u)$ such that for $p, q \geq N$,

$$\sum_{k=1}^{\infty} \left| u_k^r \left[\left\{ x_k^{(p)} \right\}^r - \left\{ x_k^{(q)} \right\}^r \right] \right| \leq \in$$

Thus for every m and for $p, q \geq N$,

$$\sum_{k=1}^{m} \left| u_k^r \left[\left\{ x_k^{(p)} \right\}^r - \left\{ x_k^{(q)} \right\}^r \right] \right| \leq \in$$

If q is fixed and p increased, since $\underset{p \to \infty}{lt}\, x_k^{(p)} = x_k$, due to c-convergence, we have

$$\sum_{k=1}^{m} \left| u_k^r \left[x_k^r - \left\{ x_k^{(q)} \right\}^r \right] \right| \leq \in \qquad \ldots(1)$$

for $q \geq N$ and every m. Let $m \to \infty$, then for $q \geq N$,

$$\sum_{k=1}^{\infty} \left| u_k^r \left[x_k^r - \left\{ x_k^{(q)} \right\}^r \right] \right| \leq \in \qquad \ldots(2)$$

From (1), since $|a-b| \geq |a| - |b|$, we obtain

$$\in \geq \sum_{k=1}^{m} \left| u_k^r x_k^r - \left\{ u_k x_k^{(q)} \right\}^r \right| \geq \sum_{k=1}^{m} \left| u_k x_k \right|^r - \sum_{k=1}^{m} \left| u_k x_k^{(q)} \right|^r$$

Hence $\sum_{k=1}^{m} \left| u_k x_k \right|^r \leq \in + \sum_{k=1}^{m} \left| u_k x_k^{(q)} \right|^r$

But since $x^{(q)}$ is in F and u is in $E(\leq F^n)$, $\sum_{k=1}^{\infty} |u_k x_k^{(q)}|^r$ converges,

and thus $\sum_{k=1}^{\infty} |u_k x_k|^r$ converges.

Hence x is in E^{η} and condition (i) of projective limit is satisfied.

Also by (2),

$$\left| \sum_{k=1}^{\infty} u_k^r \left[\left\{ x_k^r \right\} - \left\{ x_k^{(q)} \right\}^r \right] \right| \leq \sum_{k=1}^{\infty} \left| u_k^r \left[x_k^r - \left\{ x_k^{(q)} \right\}^r \right] \right| \leq \in$$

Therefore $\lim_{q\to\infty} \sum_{k=1}^{\infty} u_k^r \left\{x_k^{(q)}\right\}^r = \sum_{k=1}^{\infty} u_k^r x_k^r$

and condition (ii) is satisfied

Hence r-FE-lim $x^{(n)} = x$

This proves the theorem.

From the condition (i) of r-FE-limit, it is in E^η. Next result enables us to determine the set of r-FE-limits when $\phi \le F$.

If $x = \{x_k\}$ and $x^{(n)} = \{x_1, x_2, ..., x_n, o, o, ...\}$, then $x^{(n)}$ is called a section of x.

Theorem-2

If $\phi \le F$, every sequence in E^η is the r-FE-limit of its sections.

Proof

Let $x = \{x_k\}$ and $x_k^{(n)} = x_k (1 \le k \le n)$

$= 0 (k > n)$

Then $x^{(n)}$ is a section of x and is in ϕ and hence in F. If $x \in E^\eta$, then $\sum_{k=1}^{\infty} |u_k x_k|^r$ converges for every u in E, and

$$\lim_n \sum_{k=1}^{\infty} \left[u_k x_k^{(n)}\right]^r = \lim_n \sum_{k=1}^{n} u_k^r x_k^r = \sum_{k=1}^{\infty} u_k^r x_k^r$$

Hence r-FE-lim $x^{(n)} = x$

which proves the result.

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Working Principle of DC Machine

Rajesh Kumar

Surveyor & Loss Assessor, Muzaffarpur

A *DC motor* is a mechanically commutated electric motor powered from direct current (DC). The stator is stationary in space by definition and therefore its current. The current in the rotor is switched by the commutator to also be stationary in space. This is how the relative angle between the stator and rotor magnetic flux is maintained near 90 degrees, which generates the maximum torque.

DC motors have a rotating armature winding (winding in which a voltage is induced) but non-rotating armature magnetic field and a static field winding (winding that produce the main magnetic flux) or permanent magnet. Different connections of the field and armature winding provide different inherent speed/torque regulation characteristics. The speed of a DC motor can be controlled by changing the voltage applied to the armature or by changing the field current. The introduction of variable resistance in the armature circuit or field circuit allowed speed control. Modern DC motors are often controlled by power electronics systems called DC drives.

The introduction of DC motors to run machinery eliminated the need for local steam or internal combustion engines, and line shaft drive systems. DC motors can operate directly from rechargeable batteries, providing the motive power for the first electric vehicles. Today DC motors are still found in applications as small as toys and disk drives, or in large sizes to operate steel rolling mills and paper machines.

Brushed DC Electric Motor

A brushed DC motor is an internally commutated electric motor designed to be run from a direct current power source. Brushed motors were the first commercially important application of electric power to driving mechanical loads, and DC distribution systems were used for more than 100 years to operate motors in commercial and industrial buildings. Brushed DC motors can be varied in speed by changing the operating voltage or the strength of the magnetic field. Depending on the connections of the field to the power supply, the speed and torque characteristics of a brushed motor can be altered to provide steady speed or speed inversely proportional to the mechanical load. Brushed motors continue to be used for electrical propulsion, cranes, paper machines and steel rolling mills.

Since the brushes wear down and require replacement, brushless motors using power electronic devices have displaced brushed motors from many applications.

When a current passes through the coil wound around a soft iron core, the side of the positive pole is acted upon by an upwards force, while the other side is acted upon by a downward force. According to Fleming's left hand rule, the forces cause a turning effect on the coil, making it rotate. To make the motor rotate in a constant direction, "direct current" commutators make the current reverse in direction every half a cycle (in a two-pole motor) thus causing the motor to continue to rotate in the same direction.

A problem with the motor shown above is that when the plane of the coil is parallel to the magnetic field—i.e. when the rotor poles are 90 degrees from the stator poles—the torque is zero. In the pictures above, this occurs when the core of the coil is horizontal—the position it is just about to reach in the last picture on the right. The motor would not be able to start in this position. However, once it was started, it would continue to rotate through this position by momentum.

There is a second problem with this simple pole design. At the zero-torque position, both commutator brushes are touching (bridging) both commutator plates, resulting in a short-circuit. The power leads are shorted together through the commutator plates, and the coil is also short-circuited through both brushes (the coil is shorted twice, once through each brush independently). Note that this problem is independent of the non-starting problem above; even if there were a high current in the coil at this position, there would still be zero torque. The problem here is that this short uselessly consumes power without producing any motion (nor even any coil current.) In a low-current battery-powered demonstration this short-circuiting is generally not considered harmful. However, if a two-pole motor were designed to do actual work with several hundred watts of power output, this shorting could result in severe commutator overheating, brush damage, and potential welding of the brushes—if they were metallic—to the commutator. Carbon brushes, which are often used, would not weld. In any case, a short like this is very wasteful, drains batteries rapidly and, at a minimum, requires power supply components to be designed to much higher standards than would be needed just to run the motor without the shorting.

One simple solution is to put a gap between the commutator plates which is wider than the ends of the brushes. This increases the zero-torque range of angular positions but eliminates the shorting problem; if the motor

is started spinning by an outside force it will continue spinning. With this modification, it can also be effectively turned off simply by stalling (stopping) it in a position in the zero-torque (i.e. commutator non-contacting) angle range. This design is sometimes seen in homebuilt hobby motors, e.g. for science fairs and such designs can be found in some published science project books. A clear downside of this simple solution is that the motor now coasts through a substantial arc of rotation twice per revolution and the torque is pulsed. This may work for electric fans or to keep a flywheel spinning but there are many applications, even where starting and stopping are not necessary, for which it is completely inadequate, such as driving the capstan of a tape transport, or any instance where to speed up and slow down often and quickly is a requirement. Another disadvantage is that, since the coils have a measure of self inductance, current flowing in them cannot suddenly stop. The current attempts to jump the opening gap between the commutator segment and the brush, causing arcing.

Even for fans and flywheels, the clear weaknesses remaining in this design—especially that it is not self-starting from all positions—make it impractical for working use, especially considering the better alternatives that exist. Unlike the demonstration motor above, DC motors are commonly designed with more than two poles, are able to start from any position, and do not have any position where current can flow without producing electromotive power by passing through some coil. Many common small brushed DC motors used in toys and small consumer appliances, the simplest mass-produced DC motors to be found, have three-pole armatures. The brushes can now bridge two adjacent commutator segments without causing a short circuit. These three-pole armatures also have the advantage that current from the brushes either flows through two coils in series or through just one coil. Starting with the current in an individual coil at half its nominal value (as a result of flowing through two coils in series), it rises to its nominal value and then falls to half this value. The sequence then continues with current in the reverse direction. This results in a closer step-wise approximation to the ideal sinusoidal coil current, producing a more even torque than the two-pole motor where the current in each coil is closer to a square wave. Since current changes are half those of a comparable two-pole motor, arcing at the brushes is consequently less.

If the shaft of a DC motor is turned by an external force, the motor will act like a generator and produce an Electromotive force (EMF). During normal operation, the spinning of the motor produces a voltage, known

as the counter-EMF (CEMF) or back EMF, because it opposes the applied voltage on the motor. The back EMF is the reason that the motor when free-running does not appear to have the same low electrical resistance as the wire contained in its winding.

This is the same EMF that is produced when the motor is used as a generator (for example when an electrical load, such as a light bulb, is placed across the terminals of the motor and the motor shaft is driven with an external torque).

Therefore, the total voltage drop across a motor consists of the CEMF voltage drop, and the parasitic voltage drop resulting from the internal resistance of the armature's windings. The current through a motor is given by the following equation:

$$I = \frac{V_{applied} - V_{cemf}}{R_{armature}}$$

The mechanical power produced by the motor is given by:

$$P = I \cdot V_{cemf}$$

As an unloaded DC motor spins, it generates a backwards-flowing electromotive force that resists the current being applied to the motor. The current through the motor drops as the rotational speed increases, and a free-spinning motor has very little current. It is only when a load is applied to the motor that slows the rotor that the current draw through the motor increases.

"In an experiment of this kind made on a motor with separately excited magnets, the following figures were obtained:

Revolutions per minute	0	50	100	160	180	195
Amperes	20	16.2	12.2	7.8	6.1	5.1

Apparently, if the motor had been helped on to run at 261.5 revolutions per minute, the current would have been reduced to zero. In the last result obtained, the current of 5.1 amperes was absorbed in driving the armature against its own friction at the speed of 195 revolutions per minute."

The Commutating Plane

In a dynamo, a plane through the centres of the contact areas where a pair of brushes touch the commutator and parallel to the axis of rotation of the armature is referred to as the *commutating plane*. In this diagram the commutating plane is shown for just one of the brushes, assuming the other brush made contact on the other side of the commutator with radial symmetry, 180 degrees from the brush shown.

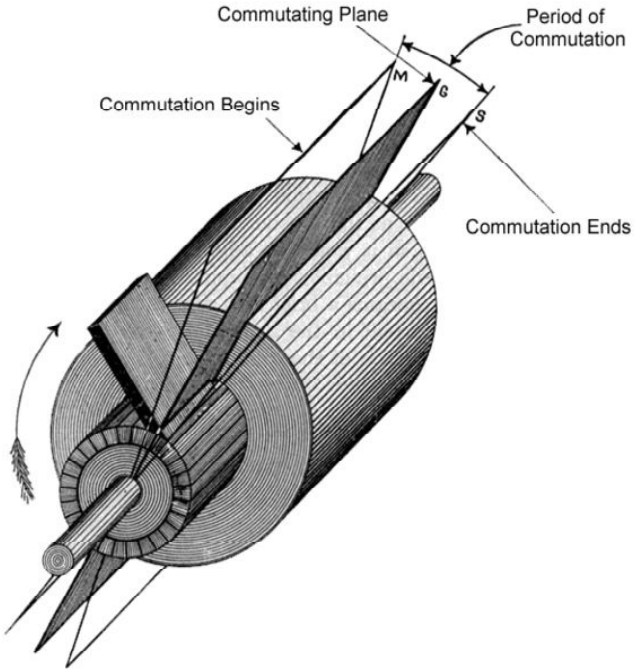


Figure: *Commutating plane definitions*

Compensation for Stator Field Distortion

In a real dynamo, the field is never perfectly uniform. Instead, as the rotor spins it induces field effects which drag and distort the magnetic lines of the outer non-rotating stator.

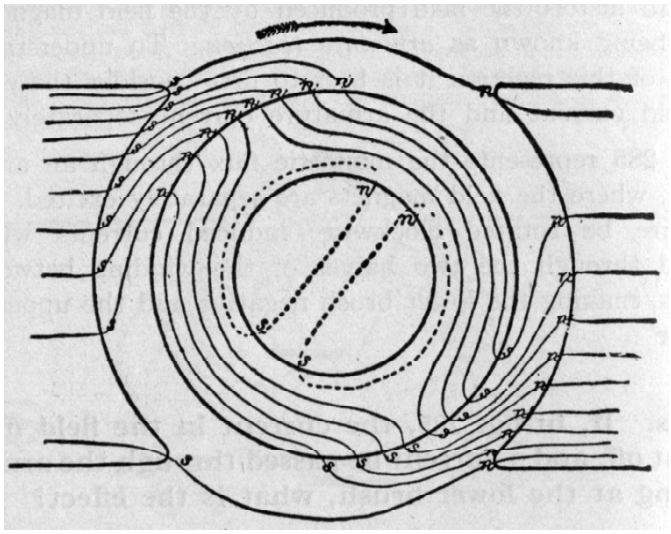


Figure: *Exaggerated example of how the field is distorted by the rotor*

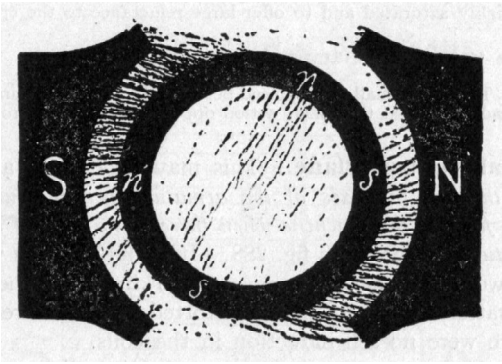


Figure: *Iron filings show the distorted field across the rotor.*

The faster the rotor spins, the further the degree of field distortion. Because the dynamo operates most efficiently with the rotor field at right angles to the stator field, it is necessary to either retard or advance the brush position to put the rotor's field into the correct position to be at a right angle to the distorted field.

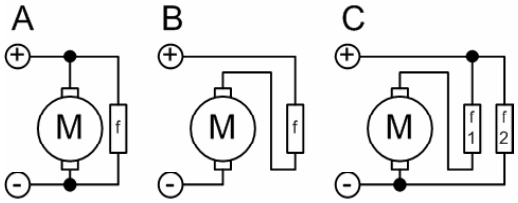
These field effects are reversed when the direction of spin is reversed. It is therefore difficult to build an efficient reversible commutated dynamo, since for highest field strength it is necessary to move the brushes to the opposite side of the normal neutral plane. The effect can be considered to be somewhat similar to timing advance in an internal combustion engine. Generally a dynamo that has been designed to run at a certain fixed speed will have its brushes permanently fixed to align the field for highest efficiency at that speed.

Motor Design Variations

DC Motors

DC motors are commonly constructed with wound rotors and either wound or permanent-magnet stators.

Wound Stators



A: shunt

B: series

C: compound

f = field coil

The field coils have traditionally existed in four basic formats: separately excited (sepex), series-wound, shunt-wound, and a combination of the latter two; compound-wound.

In a series wound motor, the field coils are connected electrically in series with the armature coils (via the brushes). In a shunt wound motor, the field coils are connected in parallel, or "shunted" to the armature coils. In a separately excited (sepex) motor the field coils are supplied from an independent source, such as a motor-generator and the field current is unaffected by changes in the armature current. The sepex system was sometimes used in DC traction motors to facilitate control of wheelslip.

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Concept of Rotating Magnetic Field

Arjun Sharma
Senior Lecturer, Advanced Technical Training Centre,
Bardang, Singtam, Sikkim

A rotating magnetic field is a magnetic field that has a moving polarity in which its opposite poles rotate about a central point or axis. Ideally the rotation changes direction at a constant angular rate. This is a key principle in the operation of the alternating-current motor.

Rotating magnetic fields are often utilized for electromechanical applications such as induction motors and electric generators, however they are also used in purely electrical applications such as induction regulators.

Description

A symmetric rotating magnetic field can be produced with as few as two polar wound coils driven at 90 degrees phasing. However, 3 sets of coils are nearly always used because it is compatible with a symmetric 3-phase AC sine current system. The three coils are driven with each set driven 120 degrees in phase from the others. For the purpose of this example, the magnetic field is taken to be the linear function of the coil's current.

The result of adding three 120-degrees phased sine waves on the axis of the motor is a single rotating vector. The rotor has a constant magnetic field. The N pole of the rotor will move toward the S pole of the magnetic field of the stator, and vice versa. This magneto-mechanical attraction creates a force which will drive the rotor to follow the rotating magnetic field in a synchronous manner.

A permanent magnet in such a field will rotate so as to maintain its alignment with the external field. This effect was utilized in early alternating current electric motors. A rotating magnetic field can be constructed using two orthogonal coils with a 90 degree phase difference in their AC currents. However, in practice such a system would be supplied through a three-wire arrangement with unequal currents. This inequality would cause serious problems in the standardization of the conductor size. In order to overcome this, three-phase systems are used where the three currents are equal in magnitude and have a 120 degree phase difference. Three similar coils having mutual geometrical angles of 120

degrees will create the rotating magnetic field in this case. The ability of the three phase system to create the rotating field utilized in electric motors is one of the main reasons why three phase systems dominate in the world electric power supply systems.

Rotating magnetic fields are also used in induction motors. Because magnets degrade with time, induction motors use short-circuited rotors (instead of a magnet) which follow the rotating magnetic field of a multicoiled stator. In these motors, the short circuited turns of the rotor develop eddy currents in the rotating field of the stator which in turn move the rotor by Lorentz force. These types of motors are not usually synchronous, but instead necessarily involve a degree of 'slip' in order that the current may be produced due to the relative movement of the field and the rotor.

History

The discovery of the rotating magnetic field is generally attributed to two inventors, the Italian physicist and electrical engineer Galileo Ferraris, and the Austrian/Serbian inventor and electrical engineer Nikola Tesla. Tesla claimed in his autobiography that he identified the concept in 1882 while Ferraris wrote about researching the concept and built a working model in 1885, although there is no independent verification for either claim. In 1888 Tesla obtained a United States patent (U.S. Patent 0,381,968) for his design and Ferraris published his research in a paper to the *Royal Academy of Sciences* in Turin.

Slip and its Significance

An AC (Alternating Current) induction motor consists of two assemblies - a stator and a rotor. The interaction of currents flowing in the rotor bars and the stators' rotating magnetic field generates a torque. In an actual operation, the rotor speed always lags the magnetic field's speed, allowing the rotor bars to cut magnetic lines of force and produce useful torque.

The difference between *the synchronous speed of the magnetic field*, and *the shaft rotating speed* is slip - and would be some number of RPM or frequency.

The slip increases with an increasing load, thus providing a greater torque. It is common to express the slip as a ratio between shaft rotation speed and synchronous magnetic field speed. The Slip is often expressed as

$$S = (n_s - n_a)\ 100\% / n_s \qquad (1)$$

where

S = slip

n_s = *synchronous speed of magnetic field (rev/min, rpm)*

n_a = *shaft rotating speed (rev/min, rpm)*

When the rotor is not turning the slip is *100 %*.

Full-load slip varies from less than *1 %* in high hp motors to more than *5-6 %* in minor hp motors.

Motor Size (hp)	0.5	5	15	50	250
Typical Slip (%)	5	3	2.5	1.7	0.8

Synchronous speed at different number of poles and frequency

No. of poles	*50 Hz*	*60 Hz*
2	3000	3600
4	1500	1800
6	1000	1200
8	750	900
10	600	720
12	500	600
16	375	450
20	300	360

Slip and Voltage

When the motor starts rotating the slip is *100 %* and the motor current is at maximum. The slip and motor current are reduced when the rotor starts to turn.

Slip Frequency

Frequency decrease when slip decrease.

Slip and Inductive Reactance

Inductive reactance depends on the frequency and the slip. When the rotor is not turning, the slip frequency is at maximum and so is the inductive reactance. A motor has a resistance and inductance and when the rotor is turning, the inductive reactance is low and the power factor approaches to *one*.

Slip and Rotor Impedance

The inductive reactance will change with the slip since the rotor impedance is the phase sum of the constant resistance and the variable inductive reactance. When the motor starts rotating the inductive reactance is high and impedance is mostly inductive. The rotor has a low lagging power factor. When the speed increases the inductive reactance goes down equaling the resistance.

Applications of Squirrel Cage

A squirrel-cage rotor is the rotating part (rotor) used in the most common form of ACinduction motor. It consists of a cylinder of steel with aluminum or copper conductors embedded in its surface. An electric motor with a squirrel-cage rotor is termed a squirrel-cage motor.

Structure

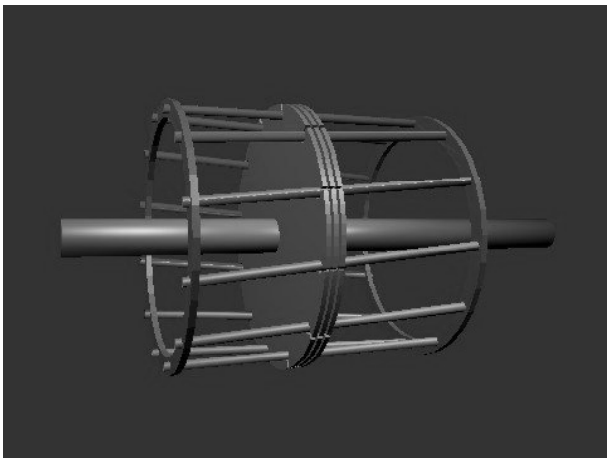

Figure: *Diagram of the squirrel-cage (showing only three laminations)*

In overall shape, it is a cylinder mounted on a shaft. Internally it contains longitudinal conductive bars (usually made of aluminum or copper) set into grooves and connected at both ends by shorting rings forming a cage-like shape. The name is derived from the similarity between this rings-and-bars winding and a squirrel cage.

The solid core of the rotor is built with stacks of electrical steel laminations. Figure shows one of many laminations used. The rotor has a smaller number of slots than the stator and must be a non-integral multiple of stator slots so as to prevent magnetic interlocking of rotor and stator teeth at the starting instant.

Theory

The field windings in the stator of an induction motor set up a rotating magnetic field through the rotor. The relative motion between this field and the rotation of the rotor induces electric current in the conductive bars.

In turn these currents lengthwise in the conductors react with the magnetic field of the motor to produce force acting at a tangentorthogonal to the rotor, resulting in torque to turn the shaft. In effect the rotor is carried around with the magnetic field but at a slightly slower rate of rotation. The difference in speed is called *slip* and increases with load.

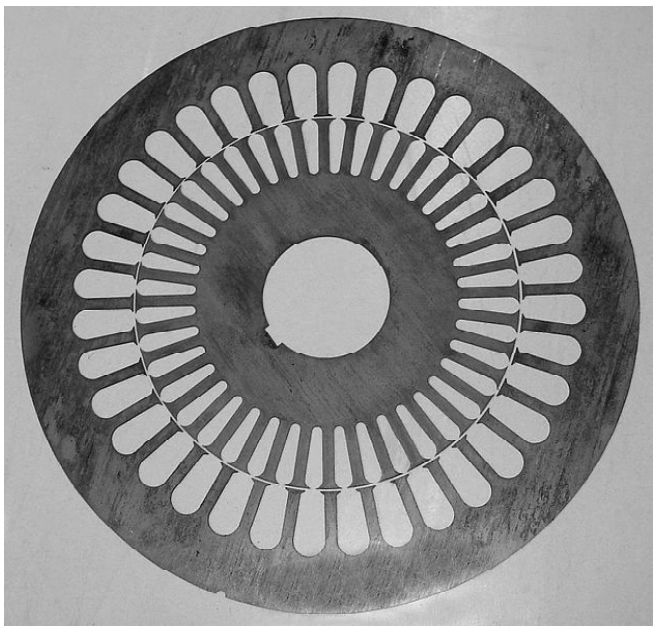

Figure: *Stator and rotor laminations*

The conductors are often skewed slightly along the length of the rotor to reduce noise and smooth out torque fluctuations that might result at some speeds due to interactions with the pole pieces of the stator. The number of bars on the squirrel cage determines to what extent the induced currents are fed back to the stator coils and hence the current through them. The constructions that offer the least feedback employ prime numbers of bars.

The iron core serves to carry the magnetic field through the rotor conductors. Because the magnetic field in the rotor is alternating with time, the core uses construction similar to a transformer core to reduce core energy losses. It is made of thin laminations, separated by varnish insulation, to reduce eddy currents circulating in the core. The material is a low carbon but high silicon iron with several times the resistivity of pure iron, further reducing eddy-current loss, and low coercivity to reduce hysteresis loss.

The same basic design is used for both single-phase and three-phase motors over a wide range of sizes. Rotors for three-phase will have variations in the depth and shape of bars to suit the design classification. Generally, thick bars have good torque and are efficient at low slip, since they present lower conductivity to the EMF. As the slip increases, skin effect starts to reduce the effective depth and increases the resistance, resulting in reduced efficiency but still maintaining torque.

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Operational Amplifiers and Applications

Himanshu Kumar

B.Tech, M.Tech (Ist Year), Electronics and Telecommunications, K.S.O.U., Manipal

Op-Amp

An operational amplifier (op-amp) is a DC-coupled high-gain electronic voltage amplifier with a differential input and, usually, a single-ended output. In this configuration, an op-amp produces an output potential (relative to circuit ground) that is typically hundreds of thousands of times larger than the potential difference between its input terminals.

Operational amplifiers had their origins in analog computers, where they were used to do mathematical operations in many linear, non-linear and frequency-dependent circuits. Characteristics of a circuit using an op-amp are set by external components with little dependence on temperature changes or manufacturing variations in the op-amp itself, which makes op-amps popular building blocks for circuit design.

Op-amps are among the most widely used electronic devices today, being used in a vast array of consumer, industrial, and scientific devices. Many standard IC op-amps cost only a few cents in moderate production volume; however some integrated or hybrid operational amplifiers with special performance specifications may cost over $100 US in small quantities. Op-amps may be packaged as components, or used as elements of more complex integrated circuits.

The op-amp is one type of differential amplifier. Other types of differential amplifier include the fully differential amplifier (similar to the op-amp, but with two outputs), the instrumentation amplifier (usually built from three op-amps), the isolation amplifier (similar to the instrumentation amplifier, but with tolerance to common-mode voltages that would destroy an ordinary op-amp), and negative feedback amplifier (usually built from one or more op-amps and a resistive feedback network).

Circuit Notation

The circuit symbol for an op-amp is shown to the right, where:

- V_+: non-inverting input
- V_-: inverting input

- V_{out}: output
- V_{S+}: positive power supply
- V_{S-}: negative power supply

The power supply pins (V_{S+} and V_{S-}) can be labelled in different ways. Often these pins are left out of the diagram for clarity, and the power configuration is described or assumed from the circuit.

Operation

The amplifier's differential inputs consist of a V_+ input and a V_- input, and ideally the op-amp amplifies only the difference in voltage between the two, which is called the *differential input voltage*. The output voltage of the op-amp is given by the equation:

$$V_{out} = A_{OL}(V_+ - V_-)$$

where V_+ is the voltage at the non-inverting terminal, V_- is the voltage at the inverting terminal and A_{OL} is the open-loop gain of the amplifier (the term "open-loop" refers to the absence of a feedback loop from the output to the input).

The magnitude of A_{OL} is typically very large—100,000 or more for integrated circuit op-amps—and therefore even a quite small difference between V_+ and V_- drives the amplifier output nearly to the supply voltage. Situations in which the output voltage is equal to or greater than the supply voltage are referred to as *saturation* of the amplifier. The magnitude of A_{OL} is not well controlled by the manufacturing process, and so it is impractical to use an operational amplifier as a stand-alone differential amplifier. Without negative feedback, and perhaps with positive feedback for regeneration, an op-amp acts as a comparator. If the inverting input is held at ground (0 V) directly or by a resistor, and the input voltage V_{in} applied to the non-inverting input is positive, the output will be maximum positive; if V_{in} is negative, the output will be maximum negative. Since there is no feedback from the output to either input, this is an *open loop* circuit acting as a comparator. The circuit's gain is just the A_{OL} of the op-amp.

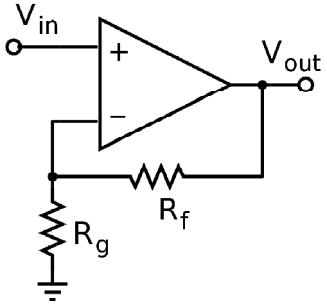

Figure: *An op-amp with negative feedback (a non-inverting amplifier)*

If predictable operation is desired, negative feedback is used, by applying a portion of the output voltage to the inverting input. The *closed loop* feedback greatly reduces the gain of the amplifier. When negative feedback is used, the circuit's overall gain and response becomes determined mostly by the feedback network rather than by the op-amp itself. If the feedback network is made of components with relatively constant, stable values, the variability of the op-amp's open loop response does not seriously affect the circuit's performance. The response of the op-amp circuit with its input, output and feedback circuits to an input is characterized mathematically by a transfer function. Designing an op-amp circuit to have a desired transfer function is in the realm of electrical engineering. The transfer functions are important in most applications of op-amps, such as in analog computers. High input impedance at the input terminals and low output impedance at the output terminal(s) are particularly useful features of an op-amp.

For example, in a non-inverting amplifier adding a negative feedback via the voltage divider R_f, R_g reduces the gain. Equilibrium will be established when V_{out} is just sufficient to reach around and "pull" the inverting input to the same voltage as V_{in}. The voltage gain of the entire circuit is determined by $1 + R_f/R_g$. As a simple example, if V_{in} = 1 V and $R_f = R_g$, V_{out} will be 2 V, the amount required to keep V_- at 1 V. Because of the feedback provided by R_f, R_g this is a *closed loop* circuit. Its overall gain V_{out} / V_{in} is called the *closed-loop gain* A_{CL}. Because the feedback is negative, in this case A_{CL} is less than the A_{OL} of the op-amp.

Another way of looking at it is to make two relatively valid assumptions.

One, that when an op-amp is being operated in linear (not saturated) mode, the difference in voltage between the non-inverting (+) pin and the inverting (–) pin is so small as to be considered negligible.

The second assumption is that the input impedance at both (+) and (–) pins is extremely high (at least several megohms with modern op-amps).

Thus, when the circuit to the right is operated as a non-inverting linear amplifier, V_{in} will appear at the (+) and (–) pins and create a current i through R_g equal to V_{in}/R_g. Since Kirchhoff's current law states that the same current must leave a node as enter it, and since the impedance into the (–) pin is near infinity, we can assume the overwhelming majority of the same current i travels through R_f, creating an output voltage equal to $V_{in} + i \times R_f$. By combining terms, we can easily determine the gain of this particular type of circuit.

$$i = \frac{V_{in}}{R_g}$$

$$V_{\text{out}} = V_{\text{in}} + i \times R_f = V_{\text{in}} + \left(\frac{V_{\text{in}}}{R_g} \times R_f \right) = V_{\text{in}} + \frac{V_{\text{in}} \times R_f}{R_g} = V_{\text{in}} \left(1 + \frac{R_f}{R_g} \right)$$

$$G = \frac{V_{\text{out}}}{V_{\text{in}}} = 1 + \frac{R_f}{R_g}$$

Op-amp Characteristics

Ideal op-amps:

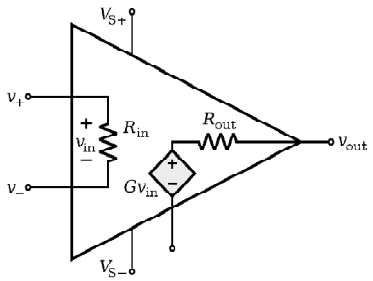

Figure: *An equivalent circuit of an operational amplifier that models some resistive non-ideal parameters.*

An ideal op-amp is usually considered to have the following properties:

- Infinite open-loop gain
- Infinite voltage range available at the output
- Infinite bandwidth with zero phase shift and infinite slew rate
- Infinite input impedance and so zero input current and zero input offset voltage
- Zero output impedance
- Zero noise
- Infinite Common-mode rejection ratio (CMRR)
- Infinite Power supply rejection ratio.

These ideals can be summarized by the two "golden rules":

I. The output attempts to do whatever is necessary to make the voltage difference between the inputs zero.

II. The inputs draw no current.

The first rule only applies in the usual case where the op-amp is used in a closed-loop design (negative feedback, where there is a signal path of some sort feeding back from the output to the inverting input). These rules are commonly used as a good first approximation for analyzing or designing op-amp circuits.

None of these ideals can be perfectly realized. A real op-amp may be modelled with non-infinite or non-zero parameters using equivalent

resistors and capacitors in the op-amp model. The designer can then include these effects into the overall performance of the final circuit. Some parameters may turn out to have negligible effect on the final design while others represent actual limitations of the final performance that must be evaluated.

Real op-amps

Real op-amps differ from the ideal model in various aspects.

DC Imperfections

Real operational amplifiers suffer from several non-ideal effects:

Finite Gain: Open-loop gain is infinite in the ideal operational amplifier but finite in real operational amplifiers. Typical devices exhibit open-loop DC gain ranging from 100,000 to over 1 million. So long as the loop gain (i.e., the product of open-loop and feedback gains) is very large, the circuit gain will be determined entirely by the amount of negative feedback (i.e., it will be independent of open-loop gain). In cases where closed-loop gain must be very high, the feedback gain will be very low, and the low feedback gain causes low loop gain; in these cases, the operational amplifier will cease to behave ideally.

Finite input impedances: The *differential input impedance* of the operational amplifier is defined as the impedance *between* its two inputs; the *common-mode input impedance* is the impedance from each input to ground. MOSFET-input operational amplifiers often have protection circuits that effectively short circuit any input differences greater than a small threshold, so the input impedance can appear to be very low in some tests. However, as long as these operational amplifiers are used in a typical high-gain negative feedback application, these protection circuits will be inactive. The input bias and leakage currents described below are a more important design parameter for typical operational amplifier applications.

Non-zero Output Impedance: Low output impedance is important for low-impedance loads; for these loads, the voltage drop across the output impedance of the amplifier will be significant. Hence, the output impedance of the amplifier limits the maximum power that can be provided. In configurations with a voltage-sensing negative feedback, the output impedance of the amplifier is effectively lowered; thus, in linear applications, op-amps usually exhibit a very low output impedance indeed. Negative feedback can not, however, reduce the limitations that R_{load} in conjunction with R_{out} place on the maximum and minimum possible output voltages; it can only reduce output errors *within* that range.

Low-impedance outputs typically require high quiescent (i.e., idle) current in the output stage and will dissipate more power, so low-power designs may purposely sacrifice low output impedance.

Input Current: Due to biasing requirements or leakage, a small amount of current (typically ~10 nanoamperes for bipolar op-amps, tens of picoamperes for JFET input stages, and only a few pA for MOSFET input stages) flows into the inputs. When large resistors or sources with high output impedances are used in the circuit, these small currents can produce large unmodelled voltage drops. If the input currents are matched, *and* the impedance looking *out* of *both* inputs are matched, then the voltages produced at each input will be equal. Because the operational amplifier operates on the *difference* between its inputs, these matched voltages will have no effect (unless the operational amplifier has poor CMRR, which is described below). It is more common for the input currents (or the impedances looking out of each input) to be slightly mismatched, and so a small *offset voltage* (different from the input offset voltage below) can be produced. This offset voltage can create offsets or drifting in the operational amplifier. It can often be nulled externally; however, many operational amplifiers include *offset null* or *balance* pins and some procedure for using them to remove this offset. Some operational amplifiers attempt to nullify this offset automatically.

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Elastic Properties of Materials and Waves and Vibrations

Bishwa Mohan Kumar

Department of Physics, R.K.D. College, Lohiya Nagar, Patna

Elasticity

In physics, elasticity is a physical property of materials which return to their original shape after they are deformed.

Overview

When an elastic material is deformed due to an external force, it experiences internal forces that oppose the deformation and restore it to its original state if the external force is no longer applied. There are various elastic moduli, such as Young's modulus, the shear modulus, and the bulk modulus, all of which are measures of the inherent stiffness of a material as a resistance to deformation under an applied load. The various moduli apply to different kinds of deformation. For instance, Young's modulus applies to uniform extension, whereas the shear modulus applies to shearing.

The elasticity of materials is described by a stress-strain curve, which shows the relation between stress (the average restorative internal force per unit area) and strain (the relative deformation). For most metals or crystaline materials, the curve is linear for small deformations, and so the stress-strain relationship can adequately be described by Hooke's law and higher-order terms can be ignored.

However, for larger stresses beyond the elastic limit, the relation is no longer linear. For even higher stresses, materials exhibit plastic behaviour, that is, they deform irreversibly and do not return to their original shape after stress is no longer applied. For rubber-like materials such as elastomers, the gradient of the stress-strain curve increases with stress, meaning that rubbers progressively become more difficult to stretch, while for most metals, the gradient decreases at very high stresses, meaning that they progressively become easier to stretch.

Elasticity is not exhibited only by only solids; non-Newtonian fluids, such as viscoelastic fluids, will also exhibit elasticity in certain conditions. In response to a small, rapidly applied and removed strain, these fluids may deform and then return to their original shape. Under larger strains,

or strains applied for longer periods of time, these fluids may start to flow like a viscous liquid.

Hooke's law

As noted above, for small deformations, most elastic materials such as springs exhibit linear elasticity. This idea was first formulated by Robert Hooke in 1675 as a Latin anagram, "ceiiinosssttuv". He published the answer in 1678: "*Ut tensio, sic vis*" meaning "*As the extension, so the force*", a linear relationship commonly referred to as Hooke's law. This law can be stated as a relationship between force F and displacement x,

$$F = -kx,$$

where k is a constant known as the *rate* or *spring constant*. It can also be stated as a relationship between stress σ and strain ε:

$$\sigma = E\varepsilon,$$

where E Is known as the elastic modulus or Young's modulus.

Although the general proportionality constant between stress and strain in three dimensions is a 4th order tensor, systems that exhibit symmetry, such as a one-dimensional rod, can often be reduced to applications of Hooke's law.

Factors Affecting Elasticity

For isotropic materials, the presence of fractures affects the Young and the shear modulus perpendicular to the planes of the cracks, which decrease (Young's modulus faster than the shear modulus) as the fracture density increases, indicating that the presence of cracks makes bodies brittler.

Microscopically, the stress-strain relationship of materials is in general governed by the Helmholtz free energy, a thermodynamic quantity. Molecules settle in the configuration which minimizes the free energy, subject to constraints derived from their structure, and, depending on whether the energy or the entropy term dominates the free energy, materials can broadly be classified as *energy-elastic* and *entropy-elastic*.

As such, microscopic factors affecting the free energy, such as the equilibriumdistance between molecules, can affect the elasticity of materials: for instance, in inorganic materials, as the equilibrium distance between molecules at 0 K increases, the bulk modulus decreases. The effect of temperature on elasticity is difficult to isolate, because there are numerous factors affecting it. For instance, the bulk modulus of a material is dependent on the form of its lattice, its behaviour under expansion, as well as the vibrations of the molecules, all of which are dependent on temperature.

Bending Moment

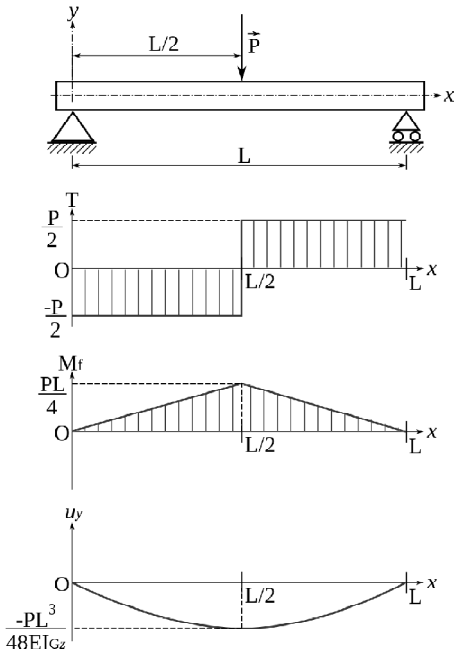

Figure: *Hinged beam with a central load: diagrams of the shearing force T, the bending moment Mf and the deflection uy; the stresses are calculated at the right of the cut.*

A bending moment exists in a structural element when a moment is applied to the element so that the element bends. Moments and torques are measured as a force multiplied by a distance so they have as unit newton-metres (N·m), or pound-foot or foot-pound (ft·lb). The concept of bending moment is very important in engineering (particularly in civil and mechanical engineering) and physics.

Discussion

Tensile and compressive stresses increase proportionally with bending moment, but are also dependent on the second moment of area of the cross-section of the structural element. Failure in bending will occur when the bending moment is sufficient to induce tensile stresses greater than the yield stress of the material throughout the entire cross-section. It is possible that failure of a structural element in shear may occur before failure in bending, however the mechanics of failure in shear and in bending are different.

The bending moment at a section through a structural element may be defined as "the sum of the moments about that section of all external forces acting to one side of that section". The forces and moments on either

side of the section must be equal in order to counteract each other and maintain a state of equilibrium so the same bending moment will result from summing the moments, regardless of which side of the section is selected.

Moments are calculated by multiplying the external vector forces (loads or reactions) by the vector distance at which they are applied. When analysing an entire element, it is sensible to calculate moments at both ends of the element, at the beginning, centre and end of any uniformly distributed loads, and directly underneath any point loads. Of course any "pin-joints" within a structure allow free rotation, and so zero moment occurs at these points as there is no way of transmitting turning forces from one side to the other. If clockwise bending moments are taken as negative, then a negative bending moment within an element will cause "sagging", and a positive moment will cause "hogging". It is therefore clear that a point of zero bending moment within a beam is a point of contraflexure—that is the point of transition from hogging to sagging or vice versa.

It is more common to use the convention that a clockwise bending moment to the left of the point under consideration is taken as positive. This then corresponds to the second derivative of a function which, when positive, indicates a curvature that is 'lower at the centre' i.e. sagging. When defining moments and curvatures in this way calculus can be more readily used to find slopes and deflections. Critical values within the beam are most commonly annotated using a bending moment diagram, where negative moments are plotted to scale above a horizontal line and positive below. Bending moment varies linearly over unloaded sections, and parabolically over uniformly loaded sections.

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Classical free Electron Theory of Metals

Kalpana Kumari Sinha

Research Scholar, Nandan Tower, B-203, Kankarbagh Colony More, Patna

This theory was developed by Drude and Lorentz and hence is also known as Drude-Lorentz theory. According to this theory, a metal consists of electrons which are free to move about in the crystal like molecules of a gas in a container. Mutual repulsion between electrons is ignored and hence potential energy is taken as zero. Therefore the total energy of the electron is equal to its kinetic energy.

Drift Velocity

If no electric field is applied on a conductor, the free electrons move in random directions. They collide with each other and also with the positive ions. Since the motion is completely random, average velocity in any direction is zero. If a constant electric field is established inside a conductor, the electrons experience a force F = -eE due to which they move in the direction opposite to direction of the field. These electrons undergo frequent collisions with positive ions. In each such collision, direction of motion of electrons undergoes random changes. As a result, in addition to the random motion, the electrons are subjected to a very slow directional motion. This motion is called drift and the average velocity of this motion is called drift velocity vd.

Consider a conductor subjected to an electric field E in the x-direction. The force on the electron due to the electric field = -eE.

By Newton's law, -eE = mdvd/dt

dvd = -eEdt/m

Integrating,

Vd = -eEt/m + Constant

When t = 0, vd = 0 Therefore Constant = 0

Vd = -eEt/m ——————— (1)

Electrical Conductivity

Consider a wire of length 'dl' and area of cross section 'A' subjected to an electric field E. If 'n' is the concentration of the electrons, the number of electrons flowing through the wire in dt seconds = nAvddt.

The quantity of charge flowing in time dt = nAvddt.e

Therefore I = dq/dt = neAvd

Current density J = I/A = nevd

Subsittuting the value of vd from (1),

J = nee Et/m = ne2Et/m ——————— (2)

By Ohm's law, J = s E

Therefore s = J/E = ne2t/m ——————— (3)

Mobility of a charge carrier is the ratio of the drift mobility to the electric field.

μ = vd/E m2/Volt-Sec

Substituting vd from (1),

μ = et/m ——————— (4)

Substituting this in equation (3),

s = neμ ——————— (5)

Relaxation Time and Mean free Path

When the field E is switched off, due to the collision of the electrons with lattice ions and lattice defects, their velocity will start to decrease. This process is called relaxation. The relaxation time(t) is the time required for the drift velocity to reduce to 1/e of its initial value. The average distance travelled by an electron between two consecutive collisions is called mean free path (l) of the electron.

l = vdt ——————— (6)

Temperature Dependence

The free electron theory is based on Maxwell-Boltzmann statistics.

Therefore Kinetic energy of electron = ½ mvd2 = 3/2 KBT

Vd = Φ 3KBT/m

Substituting this in equation (6),

t = lΦ m/3KBT ——————— (7)

Since s = ne2t/m, s is proportional to Φ1/T

Or r is proportional to ΦT.

Wiedmann-Franz Law

The ratio of thermal conductivity to electrical conductivity of a metal is directly proportional to absolute temperature.

K/s is proportional to T

Or, K/sT = L, a constant called Lorentz number.

L = 3KB2/2e2

Drawbacks of Classical free Electron Theory

1) According to this theory, r is proportional to ΦT. But experimentally it was found that r is proportional to T.
2) According to this theory, K/sT = L, a constant (Wiedmann-Franz law) for all temperatures. But this is not true at low temperatures.
3) The theoretically predicted value of specific heat of a metal does not agree with the experimentally obtained value.
4) This theory fails to explain ferromagnetism, superconductivity, photoelectric effect, Compton effect and blackbody radiation.

Quantum free Electron Theory

Classical free electron theory could not explain many physical properties. In 1928, Sommerfeld developed a new theory applying quantum mechanical concepts and Fermi-Dirac statistics to the free electrons in the metal. This theory is called quantum free electron theory. Classical free electron theory permits all electrons to gain energy. But quantum free electron theory permits only a fraction of electrons to gain energy. In order to determine the actual number of electrons in a given energy range(dE), it is necessary to know the number of states(dNs) which have energy in that range. The number of states per unit energy range is called the density of states g(E).

Therefore, g(E) = dNs/dE

According to Fermi-Dirac statistics, the probability that a particular energy statewith energy E is occupied by an electron is given by,

f(E) = 1 / [1+e(E-EF/KT)]

where EF is called Fermi level. Fermi level is the highest filled energy level at 0 K. Energy corresponding to Fermi level is known as Fermi energy. Now the actual number of electrons present in the energy range dE,

dN = f(E) g(E)dE

Effect of Temperature on Fermi-Dirac Distribution Function

Fermi-Dirac distribution function is given by,

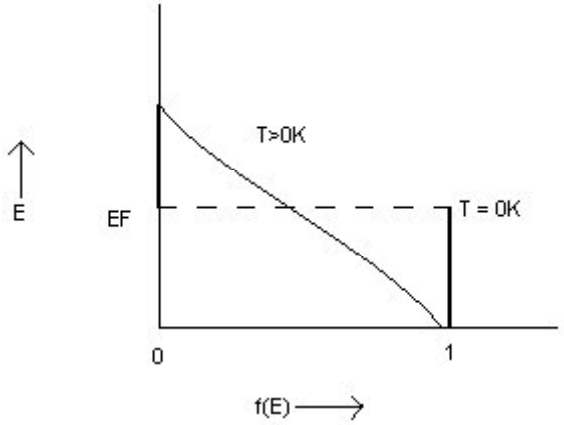
f(E) = 1 / [1+e(E-EF/KT)]

At T=0K, for EEF, f(E)=0

At T=0K, for E=EF, f(E)=indeterminate

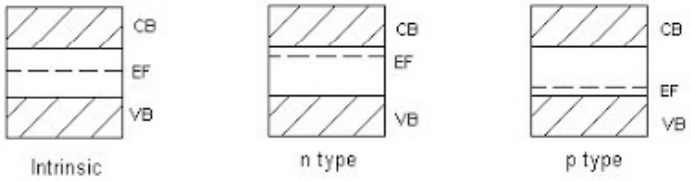
At T>0K, for E=EF, f(E)=1/2

All these results are depicted in the figure.



Classification of semiconductors on the basis of fermi level and fermi energy.

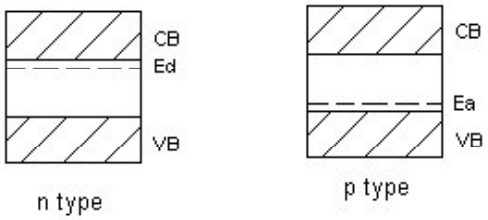
In intrinsic semiconductors, the fermilevel lie exactly at the centre of the forbidden energy gap. In n-type semiconductors fermilevel lie near the conduction band. In p-type semiconductors fermilevel lie near the valence band.



Impurity Levels

In extrinsic semiconductors, addition of impurities introduces new allowed quantum energy states in the forbidden energy band. The quantum state which appears as a single energy level is known as impurity level.

Impurity level is called donor level (Ed) in n-type semiconductors and lie just below the conduction band. Impurity level is called acceptor level (Ea) in p- type semiconductor and lie just above the valence band.



Band Theory of Solids

The atoms in the solid are very closely packed. The nucleus of an atom is so heavy that it considered to be at rest and hence the characteristic of an atom are decided by the electrons. The electrons in an isolated atom have different and discrete amounts of energy according to their occupations in different shells and sub shells. These energy values are represented by sharp lines in an energy level diagram. During the formation of a solid, energy levels of outer shell electrons got split up. As a result, closely packed energy levels are produced. The collection of such a large number of energy levels is called energy band. The electrons in the outermost shell are called valence electrons. The band formed by a series of energy levels containing the valence electrons is known as valence band. The next higher permitted band in a solid is the conduction band. The electrons occupying this band are known as conduction electrons.

Conduction band valence band are separated by a gap known as forbidden energy gap. No electrons can occupy energy levels in this band. When an electrons in the valence band absorbs enough energy, it jumps across the forbidden energy gap and enters the conduction band, creating a positively charged hole in the valence band. The hole is basically the deficiency of an electron.

Classification of solids on the basis of energy bands.

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Fundamentals of Laser

Vishwajeet Kumar Chandel

Research Scholar, Department of Physics, J.P. University, Chapra

A laser is a device that emits light through a process of optical amplification based on the stimulated emission of electromagnetic radiation. The term "laser" originated as an acronym for *Light Amplification by Stimulated Emission of Radiation*. Lasers differ from other sources of light because they emit light coherently. Its spatial coherence allows a laser to be focused to a tight spot, and this enables applications like laser cutting and laser lithography.

Its spatial coherence also keeps a laser beam collimated over long distances, and this enables laser pointers to work. Laser also have high temporal coherence which allows them to have a very narrow spectrum, i.e., they only emit a single colour of light. Their temporal coherence also allows them to emit pulses of light that only last a femtosecond.

Lasers have many important applications. They are used in common consumer devices such as DVD players, laser printers, and barcode scanners. They are used in medicine for laser surgery and various skin treatments, and in industry for cutting and welding materials.

They are used in military and law enforcement devices for marking targets and measuring range and speed. Laser lighting displays use laser light as an entertainment medium. Lasers also have many important applications in scientific research.

Fundamentals

Lasers are distinguished from other light sources by their coherence. Spatial coherence is typically expressed through the output being a narrow beam which is diffraction-limited, often a so-called "pencil beam." Laser beams can be focused to very tiny spots, achieving a very high irradiance, or they can be launched into beams of very low divergence in order to concentrate their power at a large distance. Temporal (or longitudinal) coherence implies a polarized wave at a single frequency whose phase is correlated over a relatively large distance (the coherence length) along the beam. A beam produced by a thermal or other incoherent light source has an instantaneous amplitude and phase which vary randomly with respect to time and position, and thus a very short coherence length. Most so-called "single wavelength" lasers actually produce radiation in several

modes having slightly different frequencies (wavelengths), often not in a single polarization. And although temporal coherence implies monochromaticity, there are even lasers that emit a broad spectrum of light, or emit different wavelengths of light simultaneously. There are some lasers which are not single spatial mode and consequently their light beams diverge more than required by the diffraction limit. However all such devices are classified as "lasers" based on their method of producing that light: stimulated emission. Lasers are employed in applications where light of the required spatial or temporal coherence could not be produced using simpler technologies.

Terminology

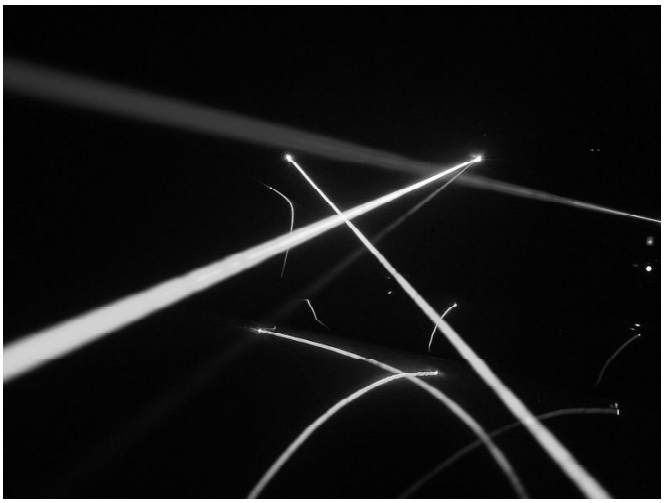


Figure: *Laser beams in fog, reflected on a car windshield*

The word *laser* started as an acronym for "light amplification by stimulated emission of radiation"; in modern usage "light" broadly denotes electromagnetic radiation of any frequency, not only visible light, hence *infrared laser*, *ultraviolet laser*, *X-ray laser*, and so on. Because the microwave predecessor of the laser, the maser, was developed first, devices of this sort operating at microwave and radio frequencies are referred to as "masers" rather than "microwave lasers" or "radio lasers". In the early technical literature, especially at Bell Telephone Laboratories, the laser was called an optical maser; this term is now obsolete.

A laser which produces light by itself is technically an optical oscillator rather than an optical amplifier as suggested by the acronym. It has been humorously noted that the acronym LOSER, for "light oscillation by stimulated emission of radiation," would have been more correct. With the widespread use of the original acronym as a common noun, actual

optical amplifiers have come to be referred to as "laser amplifiers", notwithstanding the apparent redundancy in that designation.

The back-formed verb *to lase* is frequently used in the field, meaning "to produce laser light," especially in reference to the gain medium of a laser; when a laser is operating it is said to be "lasing." Further use of the words *laser* and *maser* in an extended sense, not referring to laser technology or devices, can be seen in usages such as *astrophysical maser* and *atom laser*.

Design

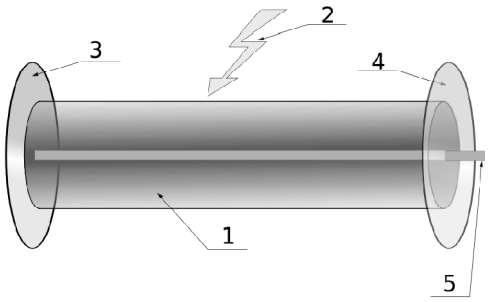


Figure: *Components of a typical laser: 1. Gain medium; 2. Laser pumping energy; 3. High reflector; 4. Output coupler; 5. Laser beam*

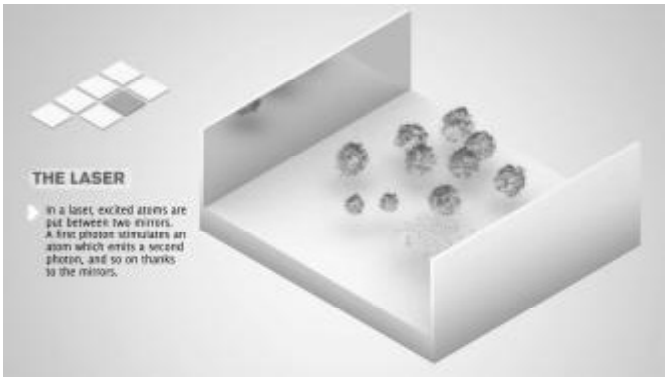


Figure: *Animation explaining the stimulated emission and the Laser principle*

A laser consists of a gain medium, a mechanism to supply energy to it, and something to provide optical feedback. The gain medium is a material with properties that allow it to amplify light by stimulated emission. Light of a specific wavelength that passes through the gain medium is amplified (increases in power).

For the gain medium to amplify light, it needs to be supplied with energy. This process is called pumping. The energy is typically supplied

as an electrical current, or as light at a different wavelength. Pump light may be provided by a flash lamp or by another laser.

The most common type of laser uses feedback from an optical cavity—a pair of mirrors on either end of the gain medium. Light bounces back and forth between the mirrors, passing through the gain medium and being amplified each time. Typically one of the two mirrors, the output coupler, is partially transparent. Some of the light escapes through this mirror. Depending on the design of the cavity (whether the mirrors are flat or curved), the light coming out of the laser may spread out or form a narrow beam. This type of device is sometimes called a *laser oscillator* in analogy to electronic oscillators, in which an electronic amplifier receives electrical feedback that causes it to produce a signal.

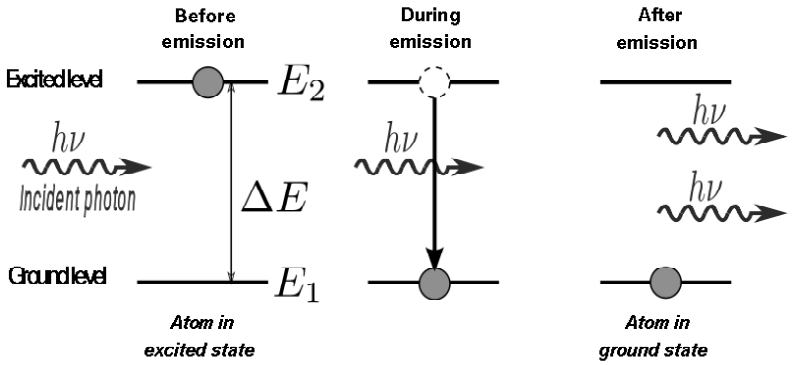
Most practical lasers contain additional elements that affect properties of the emitted light such as the polarization, the wavelength, and the shape of the beam.

Laser Physics

Electrons and how they interact with electromagnetic fields are important in our understanding of chemistry and physics.

Stimulated Emission

In the classical view, the energy of an electron orbiting an atomic nucleus is larger for orbits further from the nucleus of an atom. However, quantum mechanical effects force electrons to take on discrete positions in orbitals. Thus, electrons are found in specific energy levels of an atom, two of which are shown below:



$$E_2 - E_1 = \Delta E = h\nu$$

When an electron absorbs energy either from light (photons) or heat (phonons), it receives that incident quantum of energy. But transitions are only allowed in between discrete energy levels such as the two shown

above. This leads to emission lines and absorption lines. When an electron is excited from a lower to a higher energy level, it will not stay that way forever. An electron in an excited state may decay to a lower energy state which is not occupied, according to a particular time constant characterizing that transition. When such an electron decays without external influence, emitting a photon, that is called "spontaneous emission". The phase associated with the photon that is emitted is random. A material with many atoms in such an excited state may thus result in radiation which is very spectrally limited (centred around one wavelength of light), but the individual photons would have no common phase relationship and would emanate in random directions. This is the mechanism of fluorescence and thermal emission.

An external electromagnetic field at a frequency associated with a transition can affect the quantum mechanical state of the atom. As the electron in the atom makes a transition between two stationary states (neither of which shows a dipole field), it enters a transition state which does have a dipole field, and which acts like a small electric dipole, and this dipole oscillates at a characteristic frequency. In response to the external electric field at this frequency, the probability of the atom entering this transition state is greatly increased. Thus, the rate of transitions between two stationary states is enhanced beyond that due to spontaneous emission. Such a transition to the higher state is called absorption, and it destroys an incident photon (the photon's energy goes into powering the increased energy of the higher state). A transition from the higher to a lower energy state, however, produces an additional photon; this is the process of stimulated emission.

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Concept of de Broglie Matter Waves

Dr. Dharmendra Kumar Singh

Lecturer, D.B.S.D. Degree College, Kadna, Garkha, Saran

In quantum mechanics, the concept of matter waves or de Broglie waves reflects the wave–particle duality of matter. The theory was proposed by Louis de Broglie in 1924 in his PhD thesis. The de Broglie relations show that the wavelength is inversely proportional to the momentum of a particle and is also called de Broglie wavelength.

Also the frequency of matter waves, as deduced by de Broglie, is directly proportional to the total energy E (sum of its rest energy and the kinetic energy) of a particle.

Historical Context

At the end of the 19th century, light was thought to consist of waves of electromagnetic fields which propagated according to Maxwell's equations, while matter was thought to consist of localized particles. This division was challenged when, in his 1905 paper on the photoelectric effect, Albert Einstein postulated that light was emitted and absorbed as localized packets, or "quanta" (now called photons). These quanta would have an energy

$$E = h\nu$$

where v is the frequency of the light and h is Planck's constant. Einstein's postulate was confirmed experimentally by Robert Millikan and Arthur Compton over the next two decades. Thus it became apparent that light has both wave-like and particle-like properties. De Broglie, in his 1924 PhD thesis, sought to expand this wave-particle duality to all particles:

> *"When I conceived the first basic ideas of wave mechanics in 1923–24, I was guided by the aim to perform a real physical synthesis, valid for all particles, of the coexistence of the wave and of the corpuscular aspects that Einstein had introduced for photons in his theory of light quanta in 1905."*
>
> *—De Broglie*

In 1926, Erwin Schrödinger published an equation describing how this matter wave should evolve—the matter wave equivalent of Maxwell's equations—and used it to derive the energy spectrum of hydrogen. That

same year Max Born published his now-standard interpretation that the square of the amplitude of the matter wave gives the probability to find the particle at a given place. This interpretation was in contrast to De Broglie's own interpretation, in which the wave corresponds to the physical motion of a localized particle.

De Broglie Relations

Quantum Mechanics: The de Broglie equations relate the wavelength λ to the momentum p, and frequency f to the kinetic energy E (excluding its rest energy and any potential energy) of a particle:

$$\lambda = h / p$$

$$f = E / h$$

where h is Planck's constant. The equation can be equivalently written as

$$p = \hbar k$$

$$E = \hbar \omega$$

using the definitions

- $\hbar = h / 2\pi$ is the reduced Planck's constant (also known as Dirac's constant, pronounced "h-bar"),
- $k = 2\pi / \lambda$ is the angular wavenumber,
- $\omega = 2\pi f$ is the angular frequency.

In each pair, the second is also referred to as the Planck-Einstein relation, since it was also proposed by Planck and Einstein.

Special Relativity

Using the relativistic momentum formula from special relativity

$$p = \gamma m_0 v$$

allows the equations to be written as

$$\lambda = \frac{h}{\gamma m_0 v} = \frac{h}{m_0 v}\sqrt{1 - \frac{v^2}{c^2}}$$

$$f = \frac{\gamma m_0 c^2}{h} = \frac{m_0 c^2}{h\sqrt{1 - \frac{v^2}{c^2}}}$$

where m_0 is the particle's rest mass, v is the particle's velocity, γ is the Lorentz factor, and c is the speed of light in a vacuum. Group velocity (equal to the particle's speed) should not be confused with phase velocity (equal to the product of the particle's frequency and its wavelength). In the case of a non-dispersive medium, they happen to be equal, but otherwise they are not.

Four-vectors

Using the four-momentum P = (E/c, p) and the four-wavevector K = (ω/c, k), the De Broglie relations form a single equation:

$$P = \hbar K$$

which is frame-independent.

Experimental Confirmation

Matter waves were first experimentally confirmed to occur in the Davisson-Germer experiment for electrons, and the de Broglie hypothesis has been confirmed for other elementary particles. Furthermore, neutral atoms and even molecules have been shown to be wave-like.

Electrons

In 1927 at Bell Labs, Clinton Davisson and Lester Germer fired slow-moving electrons at a crystalline nickel target. The angular dependence of the reflected electron intensity was measured, and was determined to have the same diffraction pattern as those predicted by Bragg for x-rays. Before the acceptance of the de Broglie hypothesis, diffraction was a property that was thought to be only exhibited by waves. Therefore, the presence of any diffraction effects by matter demonstrated the wave-like nature of matter. When the de Broglie wavelength was inserted into the Bragg condition, the observed diffraction pattern was predicted, thereby experimentally confirming the de Broglie hypothesis for electrons.

This was a pivotal result in the development of quantum mechanics. Just as the photoelectric effect demonstrated the particle nature of light, the Davisson–Germer experiment showed the wave-nature of matter, and completed the theory of wave-particle duality. For physicists this idea was important because it means that not only can any particle exhibit wave characteristics, but that one can use wave equations to describe phenomena in matter if one uses the de Broglie wavelength.

This experiment were found to be in agreement with the values calculated from de-Broglie equation

Neutral Atoms

Experiments with Fresnel diffraction and specular reflection of neutral atoms confirm the application of the de Broglie hypothesis to atoms, i.e. the existence of atomic waves which undergo diffraction, interference and allow quantum reflection by the tails of the attractive potential. Advances in laser cooling have allowed cooling of neutral atoms down to nanokelvin temperatures. At these temperatures, the thermal de Broglie wavelengths come into the micrometre range. Using Bragg diffraction of atoms and a

Ramsey interferometry technique, the de Broglie wavelength of cold sodium atoms was explicitly measured and found to be consistent with the temperature measured by a different method.

This effect has been used to demonstrate atomic holography, and it may allow the construction of an atom probe imaging system with nanometer resolution. The description of these phenomena is based on the wave properties of neutral atoms, confirming the de Broglie hypothesis.

Molecules

Recent experiments even confirm the relations for molecules and even macromolecules, which are normally considered too large to undergo quantum mechanical effects. In 1999, a research team in Vienna demonstrated diffraction for molecules as large as fullerenes. The researchers calculated a De Broglie wavelength of the most probable C_{60} velocity as 2.5 pm. More recent experiments prove the quantum nature of molecules with a mass up to 6910 amu. In general, the De Broglie hypothesis is expected to apply to any well isolated object.

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Dynamic Memory Allocation

Dr. Avinash Kumar Mishra

Krishna Nagar Colony, Ishakchak, Bhagalpur

Linked List Implementation

In computer science, a linked list is a data structure consisting of a group of nodes which together represent a sequence. Under the simplest form, each node is composed of a data and a reference (in other words, a *link*) to the next node in the sequence; more complex variants add additional links. This structure allows for efficient insertion or removal of elements from any position in the sequence.

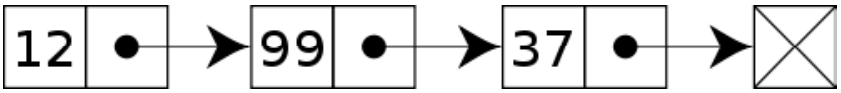


Figure: *A linked list whose nodes contain two fields: an integer value and a link to the next node. The last node is linked to a terminator used to signify the end of the list.*

Linked lists are among the simplest and most common data structures. They can be used to implement several other common abstract data types, including lists (the abstract data type), stacks, queues, associative arrays, and S-expressions, though it is not uncommon to implement the other data structures directly without using a list as the basis of implementation.

The principal benefit of a linked list over a conventional array is that the list elements can easily be inserted or removed without reallocation or reorganization of the entire structure because the data items need not be stored contiguously in memory or on disk. Linked lists allow insertion and removal of nodes at any point in the list, and can do so with a constant number of operations if the link previous to the link being added or removed is maintained during list traversal.

On the other hand, simple linked lists by themselves do not allow random access to the data, or any form of efficient indexing. Thus, many basic operations — such as obtaining the last node of the list (assuming that the last node is not maintained as separate node reference in the list structure), or finding a node that contains a given datum, or locating the place where a new node should be inserted — may require scanning most or all of the list elements.

History

Linked lists were developed in 1955-56 by Allen Newell, Cliff Shaw and Herbert A. Simon at RAND Corporation as the primary data structure

for their Information Processing Language. IPL was used by the authors to develop several early artificial intelligence programmes, including the Logic Theory Machine, the General Problem Solver, and a computer chess programme. Reports on their work appeared in IRE Transactions on Information Theory in 1956, and several conference proceedings from 1957 to 1959, including Proceedings of the Western Joint Computer Conference in 1957 and 1958, and Information Processing (Proceedings of the first UNESCO International Conference on Information Processing) in 1959. The now-classic diagram consisting of blocks representing list nodes with arrows pointing to successive list nodes appears in "Programming the Logic Theory Machine" by Newell and Shaw in Proc. WJCC, February 1957. Newell and Simon were recognized with the ACM Turing Award in 1975 for having "made basic contributions to artificial intelligence, the psychology of human cognition, and list processing". The problem of machine translation for natural language processing led Victor Yngve at Massachusetts Institute of Technology (MIT) to use linked lists as data structures in his COMIT programming language for computer research in the field of linguistics. A report on this language entitled "A programming language for mechanical translation" appeared in Mechanical Translation in 1958.

LISP, standing for list processor, was created by John McCarthy in 1958 while he was at MIT and in 1960 he published its design in a paper in the Communications of the ACM, entitled "Recursive Functions of Symbolic Expressions and Their Computation by Machine, Part I". One of LISP's major data structures is the linked list. By the early 1960s, the utility of both linked lists and languages which use these structures as their primary data representation was well established. Bert Green of the MIT Lincoln Laboratory published a review article entitled "Computer languages for symbol manipulation" in IRE Transactions on Human Factors in Electronics in March 1961 which summarized the advantages of the linked list approach. A later review article, "A Comparison of list-processing computer languages" by Bobrow and Raphael, appeared in Communications of the ACM in April 1964.

Several operating systems developed by Technical Systems Consultants (originally of West Lafayette Indiana, and later of Chapel Hill, North Carolina) used singly linked lists as file structures. A directory entry pointed to the first sector of a file, and succeeding portions of the file were located by traversing pointers. Systems using this technique included Flex (for the Motorola 6800 CPU), mini-Flex (same CPU), and Flex9 (for the Motorola 6809 CPU). A variant developed by TSC for and marketed by Smoke Signal Broadcasting in California, used doubly linked lists in the same manner.

The TSS/360 operating system, developed by IBM for the System 360/370 machines, used a double linked list for their file system catalogue. The directory structure was similar to Unix, where a directory could contain files and/or other directories and extend to any depth. A utility flea was created to fix file system problems after a crash, since modified portions of the file catalogue were sometimes in memory when a crash occurred. Problems were detected by comparing the forward and backward links for consistency. If a forward link was corrupt, then if a backward link to the infected node was found, the forward link was set to the node with the backward link. A humorous comment in the source code where this utility was invoked stated "Everyone knows a flea collar gets rid of bugs in cats".

Basic Concepts and Nomenclature

Each record of a linked list is often called an element or node. The field of each node that contains the address of the next node is usually called the *next* link or *next* pointer. The remaining fields are known as the data, information, value, cargo, or payload fields.

The head of a list is its first node. The tail of a list may refer either to the rest of the list after the head, or to the last node in the list. In Lisp and some derived languages, the next node may be called the cdr (pronounced *could-er*) of the list, while the payload of the head node may be called the car.

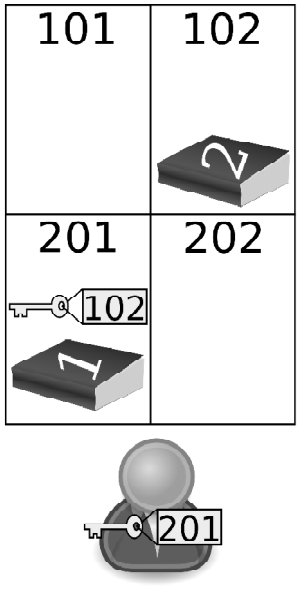


Figure: *Bob (bottom) has the key to box 201, which contains the first half of the book and a key to box 102, which contains the rest of the book.*

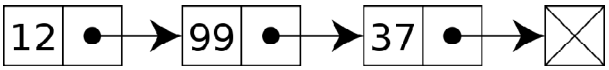
Post Office Box Analogy

The concept of a linked list can be explained by a simple analogy to real-world post office boxes. Suppose Alice is a spy who wishes to give a codebook to Bob by putting it in a post office box and then giving him the key. However, the book is too thick to fit in a single post office box, so instead she divides the book into two halves and purchases two post office boxes. In the first box, she puts the first half of the book and a key to the second box, and in the second box she puts the second half of the book. She then gives Bob a key to the first box. No matter how large the book is, this scheme can be extended to any number of boxes by always putting the key to the next box in the previous box.

In this analogy, the boxes correspond to *elements* or *nodes*, the keys correspond to *pointers*, and the book itself is the *data*. The key given to Bob is the *head* pointer, while those stored in the boxes are *next* pointers. The scheme as described above is a *singly linked list*.

Singly Linked List

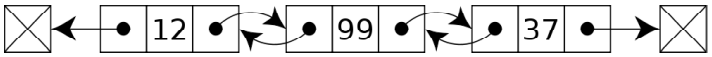
Singly linked lists contain nodes which have a data field as well as a *next* field, which points to the next node in the linked list.



A singly linked list whose nodes contain two fields: an integer value and a link to the next node

Doubly Linked List

In a doubly linked list, each node contains, besides the next-node link, a second link field pointing to the *previous* node in the sequence. The two links may be called forward(s) and backwards, or next and prev(ious).



A doubly linked list whose nodes contain three fields: an integer value, the link forward to the next node, and the link backward to the previous node

A technique known as XOR-linking allows a doubly linked list to be implemented using a single link field in each node. However, this technique requires the ability to do bit operations on addresses, and therefore may not be available in some high-level languages.

Multiply Linked List

In a multiply linked list, each node contains two or more link fields, each field being used to connect the same set of data records in a different order (e.g., by name, by department, by date of birth, etc.). While doubly linked lists can be seen as special cases of multiply linked list, the fact that

the two orders are opposite to each other leads to simpler and more efficient algorithms, so they are usually treated as a separate case.

Circular List

In the last node of a list, the link field often contains a null reference, a special value used to indicate the lack of further nodes. A less common convention is to make it point to the first node of the list; in that case the list is said to be circular or circularly linked; otherwise it is said to be open or linear.

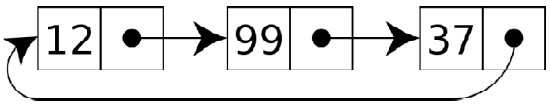


Figure: *A circular linked list*

In the case of a circular doubly linked list, the only change that occurs is that the end, or "tail", of the said list is linked back to the front, or "head", of the list and vice versa.

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Chemistry of Surfactants and Lubricants

Sanju Kumari

Lecturer in Chemistry, Mahila Polytechnic, Muzaffarpur

Surface Active Agents

Surfactants are compounds that lower the surface tension of a liquid, the interfacial tension between two liquids, or that between a liquid and a solid. Surfactants may act as detergents, wetting agents, emulsifiers, foaming agents, and dispersants.

Etymology and Definition

The term *surfactant/surfactants* is a blend of *surface active agents*.

In Index Medicus and the United States National Library of Medicine, *surfactant/surfactants* is reserved for the meaning pulmonary surfactant. For the more general meaning, *surface active agent/s* is the heading.

Composition and Structure

Surfactants are usually organic compounds that are amphiphilic, meaning they contain both hydrophobic groups (their *tails*) and hydrophilic groups (their *heads*). Therefore, a surfactant contains both a water insoluble (or oil soluble) component and a water soluble component. Surfactants will diffuse in water and adsorb at interfaces between air and water or at the interface between oil and water, in the case where water is mixed with oil. The insoluble hydrophobic group may extend out of the bulk water phase, into the air or into the oil phase, while the water soluble head group remains in the water phase. This alignment of surfactants at the surface modifies the surface properties of water at the water/air or water/oil interface.

World production of surfactants is estimated at 15 Mton/y, of which about half are soaps. Other surfactants produced on a particularly large scale are linear alkylbenzenesulfonates (1700 kton/y), lignin sulfonates (600 kton/y), fatty alcohol ethoxylates (700 ktons/y), and alkylphenol ethoxylates (500 kton/y).

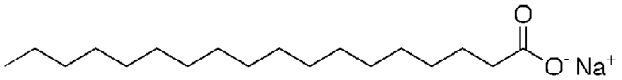

Figure: *Sodium stearate, the most common component of most soap, which comprise about 50% of commercial surfactants.*

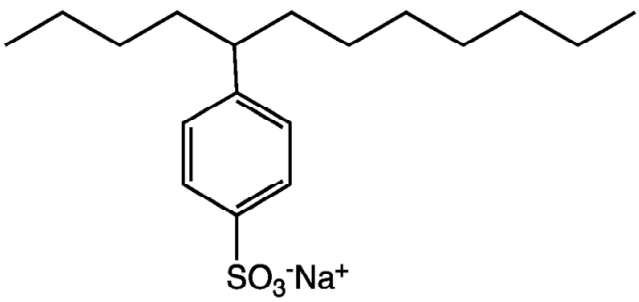

Figure: *4-(5-Dodecyl) benzenesulfonate, a linear dodecylbenzenesulfonate, one of the most common surfactants.*

Structure of Surfactant Phases in Water

In the bulk aqueous phase, surfactants form aggregates, such as micelles, where the hydrophobic tails form the core of the aggregate and the hydrophilic heads are in contact with the surrounding liquid. Other types of aggregates such as spherical or cylindrical micelles or bilayers can be formed. The shape of the aggregates depends on the chemical structure of the surfactants, depending on the balance of the sizes of the hydrophobic tail and hydrophilic head.

This is known as the HLB, Hydrophilic-lipophilic balance. Surfactants reduce the surface tension of water by adsorbing at the liquid-gas interface. The relation that links the surface tension and the surface excess is known as the Gibbs isotherm.

Dynamics of Surfactants at Interfaces

The dynamics of adsorption of surfactants is of great importance for practical applications such as foaming, emulsifying or coating processes, where bubbles or drops are rapidly generated and need to be stabilized. The dynamics of adsorption depends on the diffusion coefficient of the surfactants. Indeed, as the interface is created, the adsorption is limited by the diffusion of the surfactants to the interface. In some cases, there exists a barrier of energy for the adsorption or the desorption of the surfactants, then the adsorption dynamics is known as 'kinetically limited'. Such energy barrier can be due to steric or electrostatic repulsions. The surface rheology of surfactant layers, including the elasticity and viscosity of the surfactant layers plays a very important role in foam or emulsion stability.

Characterization of Interfaces and Surfactant Layers

Interfacial and surface tension can be characterized by classical methods such as the -pendant or spinning drop method Dynamic surface

tensions, i.e. surface tension as a function of time, can be obtained by the Maximum Bubble Pressure apparatus

The structure of surfactant layers can be studied by ellipsometry or X-Ray reflectivity.

Surface rheology can be characterized by the oscillating drop method or shear surface rheometers such as double-cone, double-ring or magnetic rod shear surface rheometer.

Detergents in Biochemistry and Biotechnology

In solution, detergents help solubilize a variety of chemical species by dissociating aggregates and unfolding proteins. Popular surfactants in the biochemistry laboratory are SDS and CTAB. Detergents are key reagents to extract protein by lysis of the cells and tissues: They disorganize the membrane's lipidic bilayer (SDS, Triton X-100, X-114, CHAPS, DOC, and NP-40), and solubilize proteins. Milder detergents such as (OctylThioGlucosides) are used to solubilize sensible proteins (enzymes, receptors). Non-solubilized material is harvested by centrifugation or other means. For electrophoresis, for example, proteins are classically treated with SDS to denature the native tertiary and quaternary structures, allowing the separation of proteins according to their molecular weight.

Detergents have also been used to decellularise organs. This process maintains a matrix of proteins that preserves the structure of the organ and often the microvascular network. The process has been successfully used to prepare organs such as the liver and heart for transplant in rats. Pulmonary surfactants are also naturally secreted by type II cells of the lung alveoli in mammals.

Classification of Surfactants

The "tail" of most surfactants are fairly similar, consisting of a hydrocarbon chain, which can be branch, linear, or aromatic. Fluorosurfactants have fluorocarbon chains. Siloxane surfactants have siloxane chains. Many important surfactants include a polyether chain terminating in a highly polar anionic group. The polyether groups often comprise ethoxylated (polyethylene oxide-like) sequences inserted to increase the hydrophilic character of a surfactant. Polypropylene oxides conversely, may be inserted to increase the lipophilic character of a surfactant.

Surfactant molecules have either one tail or two; those with two tails are said to be double-chained.

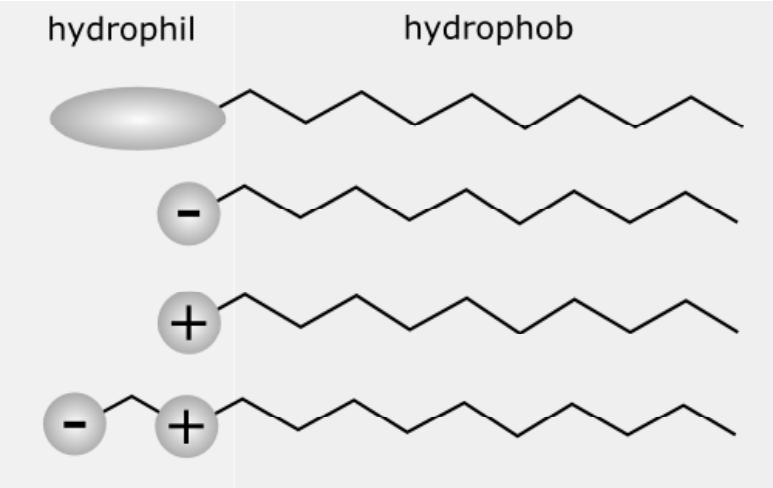

Figure: *Surfactant classification according to the composition of their head: nonionic, anionic, cationic, amphoteric.*

Most commonly, surfactants are classified according to polar head group. A non-ionic surfactant has no charge groups in its head. The head of an ionic surfactant carries a net charge. If the charge is negative, the surfactant is more specifically called anionic; if the charge is positive, it is called cationic. If a surfactant contains a head with two oppositely charged groups, it is termed zwitterionic. Commonly encountered surfactants of each type include:

Anionic

Sulphate, Sulfonate, and Phosphate Esters: Anionic surfactants contain anionic functional groups at their head, such as sulphate, sulfonate, phosphate, and carboxylates. Prominent alkyl sulphates include ammonium lauryl sulphate, sodium lauryl sulphate (SDS, sodium dodecyl sulphate, another name for the compound) and the related alkyl-ether sulphates sodium laureth sulphate, also known as sodium lauryl ether sulphate (SLES), and sodium myreth sulphate.

Docusates: dioctyl sodium sulfosuccinate, perfluorooctanesulf-onate (PFOS), perfluorobutanesulfonate, linear alkylbenzene sulfonates (LABs). These include alkyl-aryl ether phosphates and the alkyl ether phosphate

Carboxylates

These are the most common surfactants and comprise the alkyl carboxylates (soaps), such as sodium stearate. More specialized species include sodium lauroyl sarcosinate and carboxylate-based fluorosurfactants such as perfluorononanoate, perfluorooctanoate (PFOA or PFO).

Cationic head groups

- pH-dependent primary, secondary, or tertiary amines: Primary and secondary amines become positively charged at pH < 10:
  - o Octenidine dihydrochloride;
- Permanently charged quaternary ammonium cation:
  - o Alkyltrimethylammonium salts: cetyl trimethylammonium bromide (CTAB) a.k.a. hexadecyl trimethyl ammonium bromide, cetyl trimethylammonium chloride (CTAC
  - o Cetylpyridinium chloride (CPC)
  - o Benzalkonium chloride (BAC)
  - o Benzethonium chloride (BZT)
  - o 5-Bromo-5-nitro-1,3-dioxane
  - o Dimethyldioctadecylammonium chloride
  - o Cetrimonium bromide
  - o Dioctadecyldimethylammonium bromide (DODAB)

Zwitterionic Surfactants

Zwitterionic (amphoteric) surfactants have both cationic and anionic centres attached to the same molecule. The cationic part is based on primary, secondary, or tertiary amines or quaternary ammonium cations. The anionic part can be more variable and include sulfonates, as in CHAPS (3-[(3-Cholamidopropyl)dimethylammonio]-1-propanesulfonate). Other anionic groups are sultaines illustrated by cocamidopropyl hydroxysultaine. Betaines, e.g., cocamidopropyl betaine.

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Chemistry of Metals and Alloys

Dr. Hari Mohan Kumar

Dahiyawan Tola, Chapra

Gibbs' phase rule was proposed by Josiah Willard Gibbs in the 1870s as the equality

$$F = C - P + 2$$

where C is the number of components and P (alternatively π or Φ) is the number of phases in thermodynamic equilibrium with each other. Typical phases are solids, liquids and gases. A system involving one pure chemical is an example of a one-component system. Two-component systems, such as mixtures of water and ethanol, have two chemically independent components. F is the number of degrees of freedom, which means the number of intensive properties such as temperature or pressure, which are independent of other intensive variables. There are rare cases when it does not work, for instance when two substances have the same triple point, in which case, the equation predicts that there are -2 degrees of freedom.

Foundations

- A phase is a form of matter that is homogeneous in chemical composition and physical state. Typical phases are solid, liquid and gas. Two immiscible liquids (or liquid mixtures with different compositions) separated by a distinct boundary are counted as two different phases, as are two immiscible solids.
- The number of components (C) is the number of chemically independent constituents of the system, i.e. the minimum number of independent species necessary to define the composition of all phases of the system.
- The number of degrees of freedom (F) in this context is the number of intensive variables which are independent of each other.

The basis for the rule (Atkins and de Paula, justification 6.1) is that equilibrium between phases places a constraint on the intensive variables. More rigorously, since the phases are in thermodynamic equilibrium with each other, the chemical potentials of the phases must be equal. The number of equality relationships determines the number of degrees of freedom. For example, if the chemical potentials of a liquid and of its vapour depend on temperature (T) and pressure (p), the equality of chemical potentials will mean that each of those variables will be dependent on the other. Mathematically, the equation $\mu_{liq}(T, p) = \mu_{vap}(T, p)$,

where μ = chemical potential, defines temperature as a function of pressure or vice versa. (Caution: do not confuse p = pressure with *P* = number of phases.)

To be more specific, the composition of each phase is determined by *C – 1* intensive variables (such as mole fractions) in each phase. The total number of variables is *(C–1)P + 2*, where the extra two are temperature T and pressure p. The number of constraints are *C(P–1)*, since the chemical potential of each component must be equal in all phases. Subtract the number of constraints from the number of variables to obtain the number of degrees of freedom as $F = (C–1)P + 2 – C(P–1) = C – P + 2$.

The rule is valid provided the equilibrium between phases is not influenced by gravitational, electrical or magnetic forces, or by surface area, and only by temperature, pressure, and concentration.

Consequences and Examples

Pure substances (one component):

For pure substances $C = 1$ so that $F = 3 – P$. In a single phase ($P = 1$) condition of a pure component system, two variables ($F = 2$), such as temperature and pressure, can be chosen independently to be any pair of values consistent with the phase. However, if the temperature and pressure combination ranges to a point where the pure component undergoes a separation into two phases ($P = 2$), F decreases from 2 to 1. When the system enters the two-phase region, it becomes no longer possible to independently control temperature and pressure.

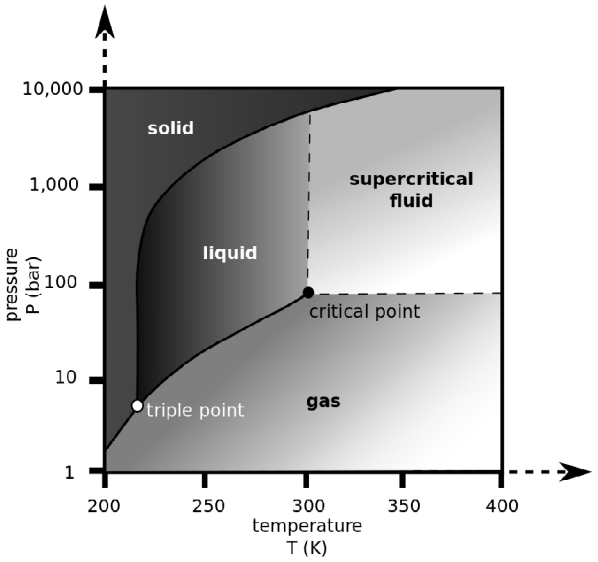

Figure: *Carbon dioxide pressure-temperature phase diagram showing the triple point and critical point of carbon dioxide*

In the phase diagram to the right, the boundary curve between the liquid and gas regions maps the constraint between temperature and pressure when the single-component system has separated into liquid and gas phases at equilibrium. If the pressure is increased by compression, some of the gas condenses and the temperature goes up. If the temperature is decreased by cooling, some of the gas condenses, decreasing the pressure. Throughout both processes, the temperature and pressure stay in the relationship shown by this boundary curve unless one phase is entirely consumed by evaporation or condensation, or unless the critical point is reached. As long as there are two phases, there is only one degree of freedom, which corresponds to position along the phase line.

The critical point is the black dot at the end of the liquid-gas boundary. As this point is approached, the liquid and gas phases become progressively more similar until, at the critical point, there is no longer a separation into two phases. Above the critical point and away from the phase boundary curve, $F = 2$ and the temperature and pressure can be controlled independently. Hence there is only one phase, and it has the physical properties of a dense gas, but is also referred to as a supercritical fluid.

Of the other two-boundary curves, one is the solid–liquid boundary or melting point curve which indicates the conditions for equilibrium between these two phases, and the other at lower temperature and pressure is the solid–gas boundary. Even for a pure substance, it is possible that three phases, such as solid, liquid and vapour, can exist together in equilibrium ($P = 3$). If there is only one component, there are no degrees of freedom ($F = 0$) when there are three phases. Therefore, in a single-component system, this three-phase mixture can only exist at a single temperature and pressure, which is known as a triple point. Here there are two equations $\mu_{sol}(T, p) = \mu_{liq}(T, p) = \mu_{vap}(T, p)$, which are sufficient to determine the two variables T and p. In the diagram for CO_2 the triple point is the point at which the solid, liquid and gas phases come together, at 5.2 bar and 217 K. It is also possible for other sets of phases to form a triple point, for example in the water system there is a triple point where ice I, ice III and liquid can coexist. If four phases of a pure substance were in equilibrium ($P = 4$), the phase rule would give $F = -1$, which is meaningless, since there cannot be –1 independent variables. This explains the fact that four phases of a pure substance (such as ice I, ice III, liquid water and water vapour) are not found in equilibrium at any temperature and pressure. In terms of chemical potentials there are now three equations, which cannot in general be satisfied by any values of the two variables T and p, although in principle they might be solved in a special case where one equation is mathematically dependent on the other two.

In practice, however, the coexistence of more phases than allowed by the phase rule normally means that the phases are not all in true equilibrium.

Two-component Systems

For binary mixtures of two chemically independent components, $C = 2$ so that $F = 4 - P$. In addition to temperature and pressure, the other degree of freedom is the composition of each phase, often expressed as mole fraction or mass fraction of one component.

As an example, consider the system of two completely miscible liquids such as toluene and benzene, in equilibrium with their vapours. This system may be described by a boiling-point diagram which shows the composition (mole fraction) of the two phases in equilibrium as functions of temperature (at a fixed pressure).

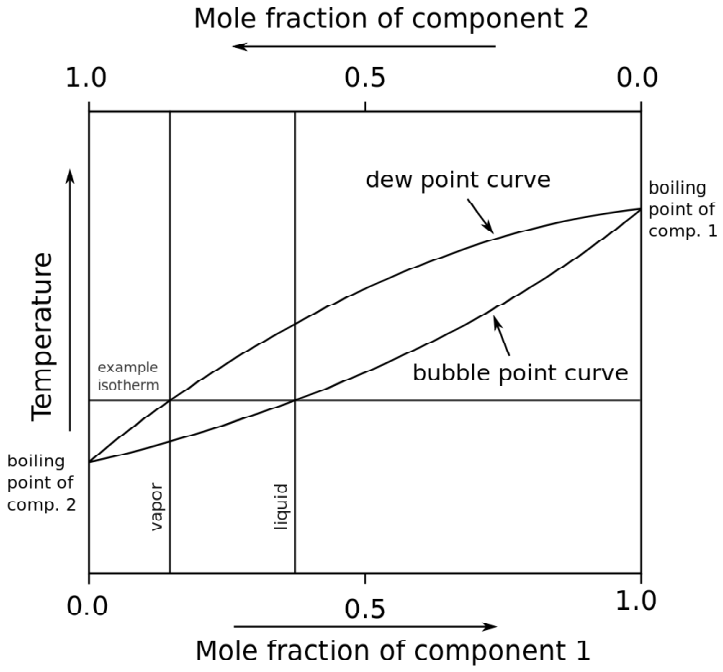

Figure: *Boiling Point Diagram*

Four thermodynamic variables which may describe the system include temperature (T), pressure (p), mole fraction of component 1 (toluene) in the liquid phase (x_{1L}), and mole fraction of component 1 in the vapour phase (x_{1V}). However since two phases are in equilibrium, only two of these variables can be independent ($F = 2$). This is because the four variables are constrained by two relations: the equality of the chemical potentials of liquid toluene and toluene vapour, and the corresponding equality for benzene.

For given T and p, there will be two phases at equilibrium when the overall composition of the system (system point) lies in between the two curves. A horizontal line (isotherm or tie line) can be drawn through any such system point, and intersects the curve for each phase at its equilibrium composition. The quantity of each phase is given by the lever rule (expressed in the variable corresponding to the x-axis, here mole fraction).

For the analysis of fractional distillation, the two independent variables are instead considered to be liquid-phase composition (x_{1L}) and pressure. In that case the phase rule implies that the equilibrium temperature (boiling point) and vapour-phase composition are determined. Liquid-vapour phase diagrams for other systems may have azeotropes (maxima or minima) in the composition curves, but the application of the phase rule is unchanged. The only difference is that the compositions of the two phases are equal exactly at the azeotropic composition.

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Internal Energy and Enthalpy in Thermal Physics

Dr. Amrendra Kumar Singh

Lecturer, Department of Physics, Jai Prakash College, Ara

For a closed system (a system from which no matter can enter or exit), the first law of thermodynamics states that the change in internal energy ΔU of the system is equal to the amount of heat Q supplied to the system minus the amount of work W done by system on its surroundings.

$$\Delta U = Q - W \quad \text{(first law)}.$$

The expressions 'Q' and 'W' represent amounts of energy but energy in different forms, heat and work. It is fundamental to the science of heat that conversion between *heat* (Q) and *work* (W) is never 100% efficient, so the ' = ' sign cannot have its usual mathematical meaning.

$$Q = \Delta U + W.$$

The work done by the system includes boundary work (when the system increases its volume against an external force, such as that exerted by a piston) and other work (e.g. shaft work performed by a compressor fan), which is called isochoric work:

$$Q = \Delta U + W_{\text{boundary}} + W_{\text{other}}.$$

The internal energy, U, is a state function. In cyclical processes, such as the operation of a heat engine, state functions return to their initial values after completing one cycle. Then the differential, or infinitesimal increment, for the internal energy in an infinitesimal process is an exact differential dU. The symbol for exact differentials is the lowercase letter d.

In contrast, neither of the infintestimal increments δQ nor δW in an infinitesimal process represents the state of the system. Thus, infinitesimal increments of heat and work are inexact differentials. The lowercase Greek letter delta, δ, is the symbol for inexact differentials.

The integral of any inexact differential over the time it takes for a system to leave and return to the same thermodynamic state does not necessarily equal zero.

The second law of thermodynamics observes that if heat is supplied to a system in which no irreversible processes take place and which has a well-defined temperature T, the increment of heat δQ and the temperature T form the exact differential

$$\mathrm{d}S = \frac{\delta Q}{T},$$

and that S, the entropy of the working body, is a function of state. Likewise, with a well-defined pressure, P, behind the moving boundary, the work differential, δW, and the pressure, P, combine to form the exact differential

$$\mathrm{d}V = \frac{\delta W}{P},$$

with V the volume of the system, which is a state variable. In general, for homogeneous systems,

$$\mathrm{d}U = T\mathrm{d}S - P\mathrm{d}V.$$

Associated with this differential equation is that the internal energy may be considered to be a function $U(S,V)$ of its natural variables S and V. The internal energy representation of the fundamental thermodynamic relation is written

$$U = U(S,V).$$

If V is constant

$$T\mathrm{d}S = \mathrm{d}U \quad (V \text{ constant})$$

and if P is constant

$$T\mathrm{d}S = \mathrm{d}H \quad (P \text{ constant})$$

with H the enthalpy defined by

$$H = U + PV.$$

The enthalpy may be considered to be a function $H(S,P)$ of its natural variables S and P. The enthalpy representation of the fundamental thermodynamic relation is written

$$H = H(S,P).$$

The internal energy representation and the enthalpy representation are partial Legendre transforms of one another. They contain the same physical information, written in different ways.

Chemical Reactions

For a closed system in which a chemical reaction is of interest, the extent of reaction, denoted by ξ, states the degree of advancement of the reaction and is included as a further natural variable for internal energy and for enthalpy. This is written

$$U = U(S,V,\xi) \text{ and } H = H(S,P,\xi).$$

In practice, chemists often use tables of a special but unnamed thermodynamic potential that is not the enthalpy expressed in its natural variables; instead they use the enthalpy expressed as a function of temperature instead of entropy. This special potential is related to the

natural form of the enthalpy $H(S,P,\xi)$ by another partial Legendre transform, that makes its natural variables T, P, and ξ. The special unnamed potential is still usually called the enthalpy. It can be written

$$H = H(T, P, \xi).$$

This enthalpy is used to report the enthalpy change of reaction, also called the heat of reaction.

Latent and Sensible Heat

In an 1847 lecture entitled *On Matter, Living Force, and Heat*, James Prescott Joule characterized the terms latent heat and sensible heat as components of heat each affecting distinct physical phenomena, namely the potential and kinetic energy of particles, respectively. He described latent energy as the energy possessed via a distancing of particles where attraction was over a greater distance, i.e. a form of potential energy, and the sensible heat as an energy involving the motion of particles or what was known as a *living force*. At the time of Joule kinetic energy either held 'invisibly' internally or held 'visibly' externally was known as a *living force*.

Latent heat is the heat released or absorbed by a chemical substance or a thermodynamic system during a change of state that occurs without a change in temperature. Such a process may be a phase transition, such as the melting of ice or the boiling of water. The term was introduced around 1750 by Joseph Black as derived from the Latin *latere* (*to lie hidden*), characterizing its effect as not being directly measurable with a thermometer.

Sensible heat, in contrast to latent heat, is the heat exchanged by a thermodynamic system that has as its sole effect a change of temperature. Sensible heat therefore only increases the thermal energy of a system.

Consequences of Black's distinction between sensible and latent heat are examined in the Wikipedia article on calorimetry.

Specific Heat

Specific heat, also called specific heat capacity, is defined as the amount of energy that has to be transferred to or from one unit of mass (kilogram) or amount of substance (mole) to change the system temperature by one degree. Specific heat is a physical property, which means that it depends on the substance under consideration and its state as specified by its properties.

The specific heats of monatomic gases (e.g., helium) are nearly constant with temperature. Diatomic gases such as hydrogen display some temperature dependence, and triatomic gases (e.g., carbon dioxide) still more.

Rigorous Mathematical Definition of Quantity of Energy Transferred as Heat

It is sometimes convenient to have a very rigorous mathematically stated definition of quantity of energy transferred as heat. Such definition is customarily based on the work of Carathéodory (1909), referring to processes in a closed system, as follows.

The internal energy U_X of a body in an arbitrary state X can be determined by amounts of work adiabatically performed by the body on its surrounds when it starts from a reference state O, allowing that sometimes the amount of work is calculated by assuming that some adiabatic process is virtually though not actually reversible. Adiabatic work is defined in terms of adiabatic walls, which allow the frictionless performance of work but no other transfer, of energy or matter. In particular they do not allow the passage of energy as heat. According to Carathéodory (1909), passage of energy as heat is allowed, by walls which are "permeable only to heat".

For the definition of quantity of energy transferred as heat, it is customarily envisaged that an arbitrary state of interest Y is reached from state O by a process with two components, one adiabatic and the other not adiabatic. For convenience one may say that the adiabatic component was the sum of work done by the body through volume change through movement of the walls while the non-adiabatic partition was excluded, and of isochoric adiabatic work. Then the non-adiabatic component is a process of energy transfer through the wall that passes only heat, newly made accessible for the purpose of this transfer, from the surroundings to the body. The change in internal energy to reach the state Y from the state O is the difference of the two amounts of energy transferred.

Although Carathéodory himself did not state such a definition, following his work it is customary in theoretical studies to define the quantity of energy transferred as heat, Q, to the body from its surroundings, in the combined process of change to state Y from the state O, as the change in internal energy, ΔU_Y, minus the amount of work, W, done by the body on its surrounds by the adiabatic process, so that $Q = \Delta U_Y - W$.

In this definition, for the sake of mathematical rigour, the quantity of energy transferred as heat is not specified directly in terms of the non-adiabatic process. It is defined through knowledge of precisely two variables, the change of internal energy and the amount of adiabatic work done, for the combined process of change from the reference state O to the arbitrary state Y. It is important that this does not explicitly involve

the amount of energy transferred in the non-adiabatic component of the combined process.

It is assumed here that the amount of energy required to pass from state *O* to state *Y*, the change of internal energy, is known, independently of the combined process, by a determination through a purely adiabatic process, like that for the determination of the internal energy of state *X* above. The mathematical rigour that is prized in this definition is that there is one and only one kind of energy transfer admitted as fundamental: energy transferred as work. Energy transfer as heat is considered as a derived quantity. The uniqueness of work in this scheme is considered to provide purity of conception, which is considered as guaranteeing mathematical rigour. The conceptual purity of this definition, based on the concept of energy transferred as work as an ideal notion, relies on the idea that some frictionless and otherwise non-dissipative processes of energy transfer can be realized in physical actuality. The second law of thermodynamics, on the other hand, assures us that such processes are not found in nature.

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Analysis of Vectors

Dr. Saryug Mandal

Lecturer in Math, B.E.C., Bhagalpur

Vector analysis, a branch of mathematics that deals with quantities that have both magnitude and direction. Some physical and geometric quantities, called scalars, can be fully defined by specifying their magnitude in suitable units of measure. Thus, mass can be expressed in grams, temperature in degrees on some scale, and time in seconds. Scalars can be represented graphically by points on some numerical scale such as a clock or thermometer. There also are quantities, called vectors, that require the specification of direction as well as magnitude. Velocity, force, and displacement are examples of vectors. A vector quantity can be represented graphically by a directed line segment, symbolized by an arrow pointing in the direction of the vector quantity, with the length of the segment representing the magnitude of the vector.

Vector Algebra

A prototype of a vector is a directed line segment AB that can be thought to represent the displacement of a particle from its initial position A to a new position B. To distinguish vectors from scalars it is customary to denote vectors by boldface letters. Thus the vector AB can be denoted by a and its length (or magnitude) by $|a|$. In many problems the location of the initial point of a vector is immaterial, so that two vectors are regarded as equal if they have the same length and the same direction.

The equality of two vectors a and b is denoted by the usual symbolic notation $a = b$, and useful definitions of the elementary algebraic operations on vectors are suggested by geometry. Thus, if $AB = a$ represents a displacement of a particle from A to B and subsequently the particle is moved to a position C, so that $BC = b$, it is clear that the displacement from A to C can be accomplished by a single displacement $AC = c$. Thus, it is logical to write $a + b = c$. This construction of the sum, c, of a and b yields the same result as the parallelogram law in which the resultant c is given by the diagonal AC of the parallelogram constructed on vectors AB and AD as sides. Since the location of the initial point B of the vector $BC = b$ is immaterial, it follows that $BC = AD$. Figure shows that $AD + DC = AC$, so that the commutative law

$$a + b = b + a \tag{1}$$

holds for vector addition. Also, it is easy to show that the associative law

$$(a+b)+c=a+(b+c) \tag{2}$$

is valid, and hence the parentheses in (2) can be omitted without any ambiguities.

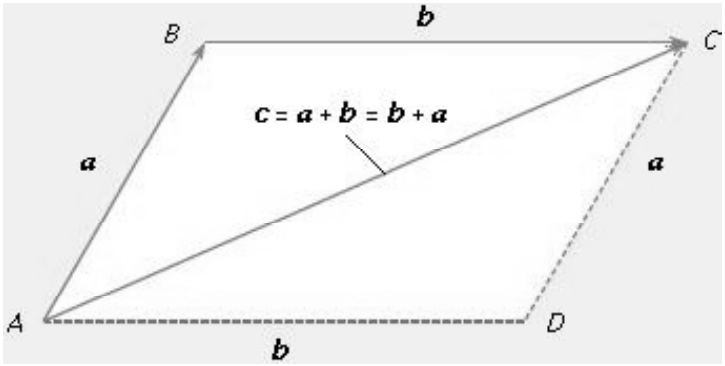

Figure: *Parallelogram law for addition of vectors*

If s is a scalar, sa or as is defined to be a vector whose length is $|s||a|$ and whose direction is that of a when s is positive and opposite to that of a if s is negative. Thus, a and $-a$ are vectors equal in magnitude but opposite in direction. The foregoing definitions and the well-known properties of scalar numbers (represented by s and t) show that

$$\begin{aligned} &s(ta)=(st)a \\ &(s+t)a=sa+ta \\ &s(a+b=sa+sb). \end{aligned} \tag{3}$$

Inasmuch as the laws (1), (2), and (3) are identical with those encountered in ordinary algebra, it is quite proper to use familiar algebraic rules to solve systems of linear equations containing vectors. This fact makes it possible to deduce by purely algebraic means many theorems of synthetic Euclidean geometry that require complicated geometric constructions.

Products of Vectors

The multiplication of vectors leads to two types of products, the dot product and the cross product.

The dot or scalar product of two vectors a and b, written $a\cdot b$, is a real number $|a||b|\cos(a,b)$, where (a,b) denotes the angle between the directions of a and b. Geometrically,

$$\begin{aligned} \mathrm{a}.\mathrm{b} &= |\mathrm{a}||\mathrm{b}|\cos(\mathrm{a},\mathrm{b}) \\ &= |\mathrm{a}| \times \text{projection of b on a}. \end{aligned} \tag{4}$$

If a and b are at right angles then $a \cdot b = 0$, and if neither a nor b is a zero vector then the vanishing of the dot product shows the vectors to be perpendicular. If $a = b$ then $\cos(a,b) = 1$, and $a \cdot a = |a|^2$ gives the square of the length of a.

The associative, commutative, and distributive laws of elementary algebra are valid for the dot multiplication of vectors. The cross or vector product of two vectors a and b, written $a \times b$, is the vector

$$a \times b = n\,|a|\,|b|\sin(a.b), \tag{5}$$

where n is a vector of unit length perpendicular to the plane of a and b and so directed that a right-handed screw rotated from a toward b will advance in the direction of n. If a and b are parallel, $a \times b = 0$. The magnitude of $a \times b$ can be represented by the area of the parallelogram having a and b as adjacent sides. Also, since rotation from b to a is opposite to that from a to b,

$$a \times b = -b \times a.$$

This shows that the cross product is not commutative, but the associative law $(sa) \times b = s(a \times b)$ and the distributive law

$$a \times b(b + c) = a \times b + a \times c \tag{6}$$

are valid for cross products.

Coordinate Systems

Since empirical laws of physics do not depend on special or accidental choices of reference frames selected to represent physical relations and geometric configurations, vector analysis forms an ideal tool for the study of the physical universe. The introduction of a special reference frame or coordinate system establishes a correspondence between vectors and sets of numbers representing the components of vectors in that frame, and it induces definite rules of operation on these sets of numbers that follow from the rules for operations on the line segments.

If some particular set of three noncollinear vectors (termed base vectors) is selected, then any vector A can be expressed uniquely as the diagonal of the parallelepiped whose edges are the components of A in the directions of the base vectors. In common use is a set of three mutually orthogonal unit vectors (*i.e.,* vectors of length 1) i, j, k directed along the axes of the familiar Cartesian reference frame. In this system the expression takes the form

$$A = xi + yj + zk,$$

where x, y, and z are the projections of A upon the coordinate axes. When two vectors A_1 and A_2 are represented as

$$A_1 = x_1 i + x_2 j + x_3 k$$
$$A_2 = y_1 i + y_2 j + y_3 k,$$

then the use of laws (3) yields for their sum

$$A_1 + A_2 = (x_1 + y_1)i + (x_2 + y_2)j + (x_3 + y_3)k. \quad (7)$$

Thus, in a Cartesian frame, the sum of A_1 and A_2 is the vector determined by $(x_1 + y_1, x_2 + y_2, x_3 + y_3)$. Also, the dot product can be written

$$A_1.A_2 = x_1 y_1 + x_2 y_2 + x_3 y_3. \quad (8)$$

since

$$i.i = j.j = k.k = 1,$$
$$i.j = j.k = k.i = 0.$$

The use of law (6) yields for

$$A_1 \times A_2 = (x_2 y_3 - x_3 y_2)i + (x_3 y_1 - x_1 y_3)j + (x_1 y_2 - x_2 y_1)k, \quad (9)$$

so that the cross product is the vector determined by the triple of numbers appearing as the coefficients of *i*, *j*, and *k* in (9).

If vectors are represented by 1 × 3 (or 3 × 1) matrices consisting of the components (x_1, x_2, x_3) of the vectors, it is possible to rephrase formulas (7) through (9) in the language of matrices. Such rephrasing suggests a generalization of the concept of a vector to spaces of dimensionality higher than three.

For example, the state of a gas generally depends on the pressure *p*, volume *v*, temperature *T*, and time *t*. A quadruple of numbers (*p*,*v*,*T*,*t*) cannot be represented by a point in a three-dimensional reference frame. But since geometric visualization plays no role in algebraic calculations, the figurative language of geometry can still be used by introducing a four-dimensional reference frame determined by the set of base vectors a_1, a_2, a_3, a_4 with components determined by the rows of the matrix

$$\begin{bmatrix} 1000 \\ 0100 \\ 0010 \\ 0001 \end{bmatrix}$$

A vector *x* is then represented in the form

$$x = x_1 a_1 + a_2 a_2 + x_3 a_3 + x_4 a_4.$$

so that in a four-dimensional space, every vector is determined by the quadruple of the components (x_1, x_2, x_3, x_4).

Calculus of Vectors

A particle moving in three-dimensional space can be located at each instant of time t by a position vector r drawn from some fixed reference point O. Since the position of the terminal point of r depends on time, r is a vector function of t. Its components in the directions of Cartesian axes, introduced at O, are the coefficients of i, j, and k in the representation

$$r = x(t)i + y(t)j + z(t)k.$$

If these components are differentiable functions, the derivative of r with respect to t is defined by the formula

$$\frac{dr}{dt} = \frac{ds}{dt}i + \frac{dy}{dt}j + \frac{dz}{dt}k = \nu, \qquad (10)$$

which represents the velocity v of the particle. The Cartesian components of v appear as coefficients of i, j, and k in (10). If these components are also differentiable, the acceleration $a = dv/dt$ is obtained by differentiating (10):

$$\frac{d^2r}{dt^2} = \frac{d^2x}{dt^2}i + \frac{d^2y}{dt^2}j + \frac{d^2z}{dt^2}k = a, \quad (11)$$

The rules for differentiating products of scalar functions remain valid for derivatives of the dot and cross products of vector functions, and suitable definitions of integrals of vector functions allow the construction of the calculus of vectors, which has become a basic analytic tool in physical sciences and technology.

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The Phyto-chemistry of Essential Oils

Bhupendra Kumar Singh
Research Scholar, Monad University, Hapur

The Chemistry of Essential Oils, and their Chemical Components

Most people either use essential oils for their therapeutic effect or for the fragrance alone but it is also interesting to take note of the chemistry, of which the oils are made up from. It is interesting to know the chemical components that nature combines to make up the oils, but it is also humbling to take note of the fact that even with the best human efforts, should you in a laboratory combine all the chemicals in the correct proportions, you would still not have an identical oil. Such a copy of an oil will not have the same therapeutic effect as the natural and pure essential oil. And though we pride ourselves on being a technology advanced society, modern science can still not unlock the secrets of essential oils and why they can do what they do. If you for instance took all the correct chemical components, which will include lavandulol, borneol, terpineol, geraniol and linalol, and try to make up lavender essential oil in a laboratory, you will not have an oil that can successfully treat burns the way that true lavender oil can.

Essential oils, like all organic compounds, are made up of hydrocarbon molecules and can further be classified as terpenes, alcohols, esters, aldehydes, ketones and phenols etc. To help you in the understanding of the chemical constituents of the oils, it may be a good idea just to have a look at what an isoprene unit is. Every single oil normally has more than a hundred components, but this figure can also run into thousands, depending on the oil in question.

When you analyse essential oils with a chromatograph various organic components are found and the primary ones are as follows:

- Terpene hydrocarbons
  - o Monoterpene hydrocarbons
  - o Sesquiterpenes
- Oxygenated compounds

- Phenols
- Alcohols
    - Monoterpene alcohols
    - Sesquiterpene alcohols
- Aldehydes
- Ketones
- Esters
- Lactones
- Coumarins
- Ethers
- Oxides.

Terpenes Hydrocarbons

Monoterpene:

- These monoterpene compounds are found in nearly all essential oils and have a structure of 10 carbon atoms and at least one double bond. The 10 carbon atoms are derived from two isoprene units.
- They react readily to air and heat sources. For this reason citrus oils do not last well, since they are high in monoterpene hydrocarbons and have a quick reaction to air, and are readily oxidised.
- Although some quarters may simply state that these components have anti-inflammatory, antiseptic, antiviral and antibacterial therapeutic properties while some can be analgesic or stimulating with a tonic effect, it could be seen as a very broad generalisation, since this large group of chemicals vary greatly. Since some have a stimulating effect on the mucus membranes they are also often used as decongestants.

Sesquiterpenes:

- These sesquiterpenes consist of 15 carbon atoms and have complex pharmacological actions and here we can look at chamazulene, which is found in German chamomile.
- It has anti-inflammatory and anti-allergy properties. Another sesquiterpene often found in chamomile and rose, as well as other floral oils is farnesene.
- History highlight of terpene research
- The 1910 Nobel prize winner for Chemistry was Professor Otto Wallach for his work on terpenes which influenced the essential oil industry.

Oxygenated compounds:

- Phenols
  - The phenols found in essential oils normally have a carbon side chain and here we can look at compounds such as thymol, eugenol and carvacrol. These components have great antiseptic, anti-bacterial and disinfectant qualities and also have greatly stimulating therapeutic properties.
  - Due to the nature of phenols, essential oils that are high in them should be used in low concentrations and for short periods of time, since they can lead to toxicity if used over long periods of time, as the liver will be required to work harder to excrete them.
  - Phenols are also classified as skin and mucus membrane irritants and although they have great antiseptic qualities, like cinnamon and clove oil, they can cause severe skin reactions.
- Alcohols
  - Monoterpene alcohols
  - These oils have good antiseptic, anti-viral and anti-fungal properties with very few side effects such as skin irritation or toxicity and have an uplifting energising effect.
  - Examples of these alcohols are linalool, citronellol and terpineol found respectively in lavender, rose and geranium, and in juniper and tea tree oil.
  - Sesquiterpene alcohols
  - These alcohols are not commonly found in essential oils, but when found, like bisabolol in German chamomile, have great properties, which include liver and glandular stimulant, anti-allergen and anti-inflammatory.
  - Other oils that contain sesquiterpene alcohols are sandalwood (a-santalol) as well as ginger, patchouli, vetiver, carrot seed, everlasting and valerian.
- Aldehydes
  - These aldehydes have anti-fungal, anti-inflammatory, disinfectant, sedative yet uplifting therapeutic qualities and are the component that imparts the citrus-like fragrance in melissa, lemongrass and citronella. These properties are best used in aromatherapy when the essential oil is used in low dilutions - around 1%.
  - Should oils high in this component be used, it could cause skin irritation and sensitivity as for instance lemongrass oil. Aldehydes are also unstable and will easily oxidise in the presence of oxygen and even low heat.

- Ketones
  - Although ketones can be toxic, as in the case of thujone found in thuja and wormwood oil as well as pinocamphone found in others, they also have some great therapeutic benefits - especially in the field of easing the secretion of mucus as well as cell and tissue regeneration.
  - Other oils, such as hyssop, eucalyptus and rosemary have moderate amounts of ketones, and when used properly in aromatherapy can be greatly beneficial to the body.
  - The ketone italidone found in everlasting, not only has the mucolytic (mucus easing) properties, but is also useful in skin regeneration, wound healing and reducing old scar tissue such as in wounds, stretch marks and adhesions.
  - Essential oils high in ketones need to be used with care in pregnancy.
- Esters
  - Esters are formed from alcohols and acids, and are named after both their original molecules with the alcohols dropping the "ol" and gaining an "yl" and the acids dropping the "ic" and gaining an "ate".
  - The esters found in essential oils are normally very fragrant and tend to be fruity and their therapeutic effects include being sedative and antispasmodic. Some esters also have anti-fungal and anti-microbial properties - like the anti-fungal properties in geranium oil.
  - The most well known ester must be lineally acetate, which is found in lavender, clary sage as well as petitgrain.
  - These components are normally gentle in their actions and can be used with great ease.
- Lactones and coumarins
  - Lactones contain an ester group integrated into a carbon ring system and coumarins are also types of lactones. There are similarities between the actions of lactones, coumarins and ketones since they also have some neurotoxic effects and can cause skin sensitising and irritation.
  - Yet the sesquiterpene lactone, called helenalin found in arnica oil, seems to be responsible for the anti-inflammatory action of arnica oil.
  - The amount of lactones and coumarins normally found in essential oils is very low, and does not pose a huge problem.

Lactones also have great mucus moving and expectorant properties and for this reason elecampane is often used in the treatment of bronchitis and chest complaints.

- Some coumarins, like furocoumarin - bergaptene - found in bergamot oil are severely skin UV sensitive and should be used with great care should you be exposed to sunlight.

- Ethers
  - Phenolic ethers are the most widely found ethers in essential oils with anethole found in aniseed, the only real ether of importance together with methyl chavicol found in basil and tarragon.
- Oxides
  - The main therapeutic effect of oxides are that of expectorant, with 1,8-cineole - commonly known as eucalyptol being the most well known.

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Plant Tissue Culture: An Overview

Uma Shankar Prasad

Department of Botany, Murarka College, Sultanganj, Bhagalpur

Plant tissue culture is a practice used to propagate plants under sterile conditions, often to produce clones of a plant. Different techniques in plant tissue culture may offer certain advantages over traditional methods of propagation, including:

- The production of exact copies of plants that produce particularly good flowers, fruits, or have other desirable traits.
- To quickly produce mature plants.
- The production of multiples of plants in the absence of seeds or necessary pollinators to produce seeds.
- The regeneration of whole plants from plant cells that have been genetically modified.
- The production of plants in sterile containers that allows them to be moved with greatly reduced chances of transmitting diseases, pests, and pathogens.
- The production of plants from seeds that otherwise have very low chances of germinating and growing, i.e.: orchids and nepenthes.
- To clean particular plant of viral and other infections and to quickly multiply these plants as 'cleaned stock' for horticulture and agriculture.

Plant tissue culture relies on the fact that many plant cells have the ability to regenerate a whole plant (totipotency). Single cells, plant cells without cell walls (protoplasts), pieces of leaves, or (less commonly) roots can often be used to generate a new plant on culture media given the required nutrients and plant hormones.

Techniques

Modern plant tissue culture is performed under aseptic conditions under filtered air. Living plant materials from the environment are naturally contaminated on their surfaces (and sometimes interiors) with microorganisms, so surface sterilization of starting materials (explants) in chemical solutions (usually alcohol or bleach) is required. Mercuric chloride is seldom used as a plant sterilant today, as it is dangerous to use, and is difficult to dispose of. Explants are then usually placed on the surface of a solid culture medium, but are sometimes placed directly into a liquid medium, particularly when cell suspension cultures are desired.

Solid and liquid media are generally composed of inorganic salts plus a few organic nutrients, vitamins and plant hormones. Solid media are prepared from liquid media with the addition of a gelling agent, usually purified agar. The composition of the medium, particularly the plant hormones and the nitrogen source (nitrate versus ammonium salts or amino acids) have profound effects on the morphology of the tissues that grow from the initial explant. For example, an excess of auxin will often result in a proliferation of roots, while an excess of cytokinin may yield shoots. A balance of both auxin and cytokinin will often produce an unorganised growth of cells, or callus, but the morphology of the outgrowth will depend on the plant species as well as the medium composition. As cultures grow, pieces are typically sliced off and transferred to new media (subcultured) to allow for growth or to alter the morphology of the culture. The skill and experience of the tissue culturist are important in judging which pieces to culture and which to discard.

As shoots emerge from a culture, they may be sliced off and rooted with auxin to produce plantlets which, when mature, can be transferred to potting soil for further growth in the greenhouse as normal plants.

Choice of Explant

The tissue obtained from the plant to culture is called an explant. Based on work with certain model systems, particularly tobacco, it has often been claimed that a totipotent explant can be grown from any part of the plant. However, this concept has been vitiated in practice. In many species explants of various organs vary in their rates of growth and regeneration, while some do not grow at all. The choice of explant material also determines if the plantlets developed via tissue culture are haploid or diploid. Also the risk of microbial contamination is increased with inappropriate explants. Thus it is very important that an appropriate choice of explant be made prior to tissue culture. The specific differences in the regeneration potential of different organs and explants have various explanations. The significant factors include differences in the stage of the cells in the cell cycle, the availability of or ability to transport endogenous growth regulators, and the metabolic capabilities of the cells. The most commonly used tissue explants are the meristematic ends of the plants like the stem tip, auxiliary bud tip and root tip. These tissues have high rates of cell division and either concentrate or produce required growth regulating substances including auxins and cytokinins.

Some explants, like the root tip, are hard to isolate and are contaminated with soil microflora that become problematic during the tissue culture process. Certain soil microflora can form tight associations with the root systems, or even grow within the root. Soil particles bound

to roots are difficult to remove without injury to the roots that then allows microbial attack. These associated microflora will generally overgrow the tissue culture medium before there is significant growth of plant tissue. Aerial (above soil) explants are also rich in undesirable microflora. However, they are more easily removed from the explant by gentle rinsing, and the remainder usually can be killed by surface sterilization. Most of the surface microflora do not form tight associations with the plant tissue. Such associations can usually be found by visual inspection as a mosaic, de-colorization or localized necrosis on the surface of the explant.

An alternative for obtaining uncontaminated explants is to take explants from seedlings which are aseptically grown from surface-sterilized seeds. The hard surface of the seed is less permeable to penetration of harsh surface sterilizing agents, such as hypochlorite, so the acceptable conditions of sterilization used for seeds can be much more stringent than for vegetative tissues. Tissue cultured plants are clones, if the original mother plant used to produce the first explants is susceptible to a pathogen or environmental condition, the entire crop would be susceptible to the same problem, conversely any positive traits would remain within the line also.

Applications

Plant tissue culture is used widely in plant science; it also has a number of commercial applications. Applications include:

- Micropropagation is widely used in forestry and in floriculture. Micropropagation can also be used to conserve rare or endangered plant species.
- A plant breeder may use tissue culture to screen cells rather than plants for advantageous characters, e.g. herbicide resistance/tolerance.
- Large-scale growth of plant cells in liquid culture in bioreactors for production of valuable compounds, like plant-derived secondary metabolites and recombinant proteins used as biopharmaceuticals .
- To cross distantly related species by protoplast fusion and regeneration of the novel hybrid.
- To cross-pollinate distantly related species and then tissue culture the resulting embryo which would otherwise normally die (Embryo Rescue).
- For production of doubled monoploid (dihaploid) plants from haploid cultures to achieve homozygous lines more rapidly in breeding programmes, usually by treatment with colchicine which causes doubling of the chromosome number.
- As a tissue for transformation, followed by either short-term testing of genetic constructs or regeneration of transgenic plants.

- Certain techniques such as meristem tip culture can be used to produce clean plant material from virused stock, such as potatoes and many species of soft fruit.
- micropropagation using meristem and shoot culture to produce large numbers of identical individuals.

Laboratories

Although some growers and nurseries have their own labs for propagating plants by the technique of tissue culture, a number of independent laboratories provide custom propagation services. The Plant Tissue Culture Information Exchange lists many commercial tissue culture labs. Since plant tissue culture is a very labour intensive process, this would be an important factor in determining which plants would be commercially viable to propagate in a laboratory.

Hairy Root Culture

Hairy root culture, also called transformed root culture, is a type of plant tissue culture that is used to study plant metabolic processes or to produce valuable secondary metabolites, often with plant genetic engineering. A naturally occurring soil bacterium *Agrobacterium rhizogenes* that contains root inducing plasmids (also called Ri plasmids) can infect plant roots and cause them to produce a food source for the bacterium, opines, and to grow abnormally.

The abnormal roots are particularly easy to culture in artificial media because hormones are not needed, and they are neoplastic, with indefinite growth. The neoplastic roots produced by *A. rhizogenes* infection have a high growth rate (compared to untransformed adventitious roots), as well as genetic and biochemical stability. Currently the main constraint for commercial utilization of hairy root culture is the development and up-scaling of appropriate (bioreactors) vessels for the delicate and sensitive hairy roots.

Metabolic Studies

Hairy root cultures can be used for phytoremediation, and are particularly valuable for studies of the metabolic processes involved in phytoremediation. Further applications include detailed studies of fundamental molecular, genetic and biochemical aspects of genetic transformation and of hairy root induction.

Genetically Transformed Cultures

The Ri plasmids can be engineered to also contain T-DNA, used for genetic transformation (biotransformation) of the plant cells. The resulting genetically transformed root cultures can produce high levels of secondary metabolites, comparable or even higher than those of intact plants.

Use in Plant Propagation

Hairy root culture can also be used for regeneration of whole plants and for production of artificial seeds.

History of Plant Tissue and Cell Culture

History of Plant Tissue Culture

Plant tissue culture, or the aseptic culture of cells, tissues, organs, and their components under defined physical and chemical conditions in vitro, is an important tool in both basic and applied studies as well as in commercial application. It owes its origin to the ideas of the German scientist, Haberlandt, at the begining of the 20th century. The early studies led to root cultures, embryo cultures, and the first true callus/tissue cultures. The period between the 1940s and the 1960s was marked by the development of new techniques and the improvement of those that were already in use. It was the availability of these techniques that led to the application of tissue culture to five broad areas, namely, cell behaviour (including cytology, nutrition, metabolism, morphogenesis, embryogenesis, and pathology), plant modification and improvement, pathogen-free plants and germplasm storage, clonal propagation, and product (mainly secondary metabolite) formation, starting in the mid-1960s. The 1990s saw continued expansion in the application of the in vitro technologies to an increasing number of plant species. Cell cultures have remained an important tool in the study of basic areas of plant biology and biochemistry and have assumed major significance in studies in molecular biology and agricultural biotechnology.

An important aspect of all biotechnology processes is the culture of either the plant cells or animal cells or microorganisms. The cells in culture can be used for recombinant DNA technology, genetic manipulations etc. Plant cell culture is based on the unique property of the cell-totipotency. Cell-totipotency is the ability of the plant cell to regenerate into whole plant. This property of the plant cells has been exploited to regenerate plant cells under the laboratory conditions using artificial nutrient mediums. With the advances made in genetic engineering, it became possible to introduce foreign genes into cell and tissue culture systems. This led to the development of Genetically Modified (GM) or Transgenic Crops which had improved traits and characteristics.

In the early 19th century, Schleiden and Schwann proposed the concept of the 'cell theory'. In 1902, Gottlieb Haberlandt, the german botanist and regarded as the father of plant tissue culture, first attempted to cultivate the mechanically isolated plant leaf cells on a simple nutrient medium. He did not succeed in achieving the growth and differentiation of the

cultured cells, however, he predicted the concept of growth hormones, the use of embryo sac fluids, the cultivation of artificial embryos from somatic cells, etc.

During the period 1902 - 1930, attempts were made to culture the isolated plant organs such as roots and shoot apices (organ culture). Hanning (1904) isolated embryos of some crucifers and successfully grew on mineral salts and sugar solutions. Simon (1908) successfully regenerated a bulky callus, buds, roots from a poplar tree on the surface of medium containing IAA which proliferated cell division. Gautheret, White and Nobecourt (1934-1940) largely contributed to the developments made in plant tissue culture. White (1939) cultured tobacco tumour tissue from the hybrid Nicotiana glauca, and N. Langsdorffii. The period of 1940 - 1970s saw the development of suitable nutrient media to culture plant tissues, embryos, anthers, pollen, cells and protoplasts, and the regeneration of complete plants (in vitro morphogenesis) from cultured tissues and cells. In 1941, van Overbek and co-workers used coconut milk (embryo sac fluid) for embryo development and callus formation in Datura. Steward and Reinert (1959) first discovered somatic embryo production in vitro. Maheswari and Guha (1964) developed the anther culture for the production of haplid plants.

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Plant Diseases: Development and Management

Dr. Amit Kumar Dwivedy

Assistant Professor, Department of Botany, S.T.S.M. College, Panwari (V.K.S.U., Ara)

The interactions between plants and disease organisms are complex, and commercial growers and home gardeners alike may have difficulty understanding plant diseases. Confusion can be reduced by learning a few basic concepts and principles of how diseases develop and how they are managed. This publication presents these concepts and is dedicated to commercial growers, commercial applicators, home gardeners and others who want more information on how plant diseases develop and are managed. To serve these diverse groups, examples of common North Dakota diseases were selected from both the commercial farm and from the home garden. Various circulars available at offices of the NDSU Extension Service provide detailed information on many of these diseases.

How Disease Develops

What is a Disease?

A disease is any abnormal condition that damages a plant and reduces its productivity or usefulness to man. Under this definition, air pollution can cause disease, as can many fungi and other infectious living organisms. This illustrates the first important concept: there are two basic types of diseases, non-infectious (abiotic) and infectious (biotic).

Types of Diseases

Non-infectious (Abiotic)

Non-infectious diseases are caused by some environmental factor that produces an abnormal plant; that is, one that has an abnormal appearance. Non-infectious diseases are *not* caused by a living, parasitic organism (an organism that gets its food by attacking other organisms), but are abiotic in nature.

Nutrition

Nutrition is a frequent cause of non-infectious disease. Either too much (excess) or too little (deficiency) can cause problems. For example, plants that are deficient in nitrogen develop a general yellowing, beginning with the lower leaves and progressing upward. Trace element deficiencies such as iron chlorosis, caused by iron deficiency, are common. Iron chlorosis occurs in many North Dakota trees and shrubs, especially silver maple,

oak, and spirea. Iron chlorosis is recognized by progressively smaller leaves on the new growth; these leaves are yellow with green veins. When iron chlorosis is severe, leaves may turn brown and become brittle as well. Lime-induced chlorosis is common in our alkaline soils because the iron in the soil is not readily available to plants. Iron chlorosis also is common on certain soybean varieties. Zinc deficiency is common on dry beans and fairly common on flax, causing yellow leaves and stunted growth. Excess trace elements may also cause growth problems, but these are rare in North Dakota.

Moisture

Deficient or excessive moisture (water) can cause disease. Moisture deficiency produces stunted, stressed or wilted plants. In addition, this stress may predispose (weaken) plants to infection by infectious organisms or increase the effects of infectious disease. For example, some tree canker organisms commonly infect trees stressed by drought or extreme cold. The effects of stem rust and root rot on small grains are greater when plants are moisture stressed (deficient in water). Excess moisture also has adverse effects, such as suffocation of roots due to lack of oxygen or predisposing plants to water mold infections.

Temperature

Frost is a common problem in spring and fall, affecting tender farm crops and garden vegetables. Extremely high temperatures in summer can also cause problems. For example, heat sterility in small grains is common in North Dakota. In oats, this is referred to as "blast."

Other Meteorological Conditions

High soil temperatures early in the season may injure or kill plant tissues at the soil surface, resulting in a constricted stem; this is called heat canker. Bright sun, high temperatures, and strong dry winds may suddenly desiccate (dry) leaves of crops and garden plants, resulting in sunscald. When lightning strikes the ground it may kill plants in somewhat circular patches up to 50 feet in diameter.

Toxic Chemicals

Toxic chemicals injure plants. Salt may damage or kill farm crops growing in saline seeps; road salt may severely damage boulevard trees and other vegetation. Air pollution also damages vegetation. Bronzing of beans caused by ozone is common in the state. The source of the ozone is not known.

Infectious (Biotic)

Infectious diseases are caused by organisms that attack plants and get their nutrition from them. The plant attacked is called the host plant. The organism causing the disease is called a pathogen. The pathogen can spread from a diseased plant to a healthy plant. There are five common

groups of pathogens A few other kinds of microorganisms may cause plant disease but are not common in North Dakota.

Fungi

Fungi are the most common pathogens in North Dakota. They produce tiny thread-like filaments called hyphae. Most pathogenic fungi produce spores which serve to reproduce and disseminate them. Spores function similarly to the seeds of higher plants. Some spores are formed in masses, like the orange pustules of rust fungi. Other spores develop in specialized fruiting structures. These structures are called signs of the pathogen and are useful in field identification of disease. Symptoms are also useful in identification of a disease. Symptoms are visible abnormalities such as wilts, rots, and other types of tissue death, stunting, excessive growth, or abnormal colour. Examples of common pathogenic fungal diseases in North Dakota include: rusts of small grains, sunflower and dry beans; cereal smuts and head scab of small grains; Cercospora leafspot of sugarbeet; white mold or Sclerotinia of dry beans, sunflowers, and canola; early blight and late blight of potatoes; root rots of small grains, sugarbeets and dry beans; apple scab; anthracnose of muskmelon; tree cankers; Septoria leafspot of tomato; peony blight; powdery mildews of ornamentals, and plum pockets.

Most fungi that cause plant diseases are parasites, organisms that get their food from other living organisms. However, not all fungi are parasites. Many live on dead or decaying organic matter and are called saprophytes. Mushrooms that spring up in lawns are among the most spectacular saprophytic fungi. There are also many inconspicuous ones that rot organic matter. The sooty molds seen on wheat heads at harvest also are saprophytic, living on the already ripe or senescing glumes and awns.

Bacteria

Bacteria are tiny one-celled organisms that multiply by cell division. They can be seen only with a microscope. Most are saprophytes, but there are a few common and serious bacterial pathogens that attack North Dakota plants. Examples of common bacterial diseases include bacterial blights of dry beans, bacterial blights and black chaff of wheat and barley, ring rot and blackleg of potato, fireblight of apples and related plants, bacterial wilt of cucumber and muskmelon, angular leafspot of cucumber, and bacterial speck and spot of tomato.

Viruses

Viruses are 1,000 times smaller than the tiniest living cell. Most viruses have a core of nucleic acid, the basic unit of heredity, and have a protein coat covering the core. Viruses are usually in the form of rods or spheres and alter the activities of the host to manufacture more virus. Some viruses are transmitted mechanically (by contact with another plant, or

contaminated workers' hands or tools); others are transmitted (carried) by insects and by eriophyid mites. Examples of virus diseases that can cause serious losses are wheat streak mosaic, barley yellow dwarf, bean common mosaic, potato virus diseases, tobacco mosaic, cucumber mosaic, and squash mosaic.

Phytoplasmas

Phytoplasmas lack a rigid cell wall, have no defined shape and can only be seen with an electron microscope. They are usually systemic in the host (distributed internally throughout the host) and are transmitted by leafhoppers. Phytoplasmas cause growth abnormalities such as witches' brooms (a broom-like mass of plant branches) or excessive tillering (stooling of small grains). Our most common phytoplasma is aster yellows. Aster yellows produces witches' brooms and greenish flowers on marigold; carrots develop yellow tops and hairy roots. The disease produces greenish flowers on flax, "purple top" on potato and tomato, and bladder-like pods on canola. Purple top of potato is often accompanied by the formation of small aerial tubers in the leaf axils (the point where the leaf joins the stem).

Nematodes

Nematodes are tiny roundworms. Most can be seen only with a microscope, but a few can be seen with the naked eye. Reproduction is by formation of eggs. Some parasitic forms attack plant roots and can cause severe damage. Nematode problems are common in warm climates but are rare in North Dakota. Two potentially serious nematode diseases occur in neighboring states but are not yet confirmed (2000) in North Dakota. They are the soybean cyst nematode and the pine wilt nematode. The pine wilt nematode is unusual in that it invades the vascular (water-conducting) tissues of pine trees instead of attacking the roots.

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Physiology: Plant Growth and Development

Rashmi Rani

Research Scholar, Ratanpura, Chapra, Saran

The three major plant functions that are the basics for plant growth and development are photosynthesis, respiration, and transpiration.

Photosynthesis

One of the major differences between plants and animals on earth is the ability of plants to internally manufacture their own food. To produce food for itself a plant requires energy from sunlight, carbon dioxide from the air and water from the soil.

If any of these ingredients is lacking, photosynthesis, or food production, will stop. If any factor is removed for a long period of time, the plant will die. Photosynthesis literally means "to put together with light."

$$\text{Carbon dioxide} + \text{Water} \xrightarrow{(\text{Light})} \text{Sugar} + \text{Oxygen}$$

$$6CO_2\,6H_2O \xrightarrow{(\text{Light})} C_6H_{12}O_6 + 6O_2$$

Any green plant tissue is capable of photosynthesis. Chloroplasts in these cells contain the green pigment called chlorophyll which traps the light energy. However, leaves are generally the site of most food production due to their special structure. The internal tissue (mesophyll) contains cells with abundant chloroplasts in an arrangement that allows easy movement of water and air. The protective upper and lower epidermis (skin) layers of the leaf include many stomata that are openings in the leaf formed by two specialized guard cells on either side. Guard cells regulate movement of the gases, (i.e. CO_2 into and O_2 and H_2O out of the leaf), involved in photosynthesis. The lower epidermis of the leaf normally contains the largest percentage of stomata.

Light Reaction

Photosynthesis is the process of turning the energy of sunlight into chemical energy from the raw products of CO_2 and H_2O. This process is necessary to sustain nearly all forms of life. Photosynthesis is divided in to two separate reactions known as the light and dark reactions. They take place when light is present but the dark reaction does not require light. The whole process is begun by light reacting with pigments in the leaf causing the splitting of water molecules. This is called photolysis or

the *Hill Reaction* which is not completely understood. Three products are produced in this reaction. Electrons from the hydrogen molecules and remaining H^+ ions are used to form two separate energy storage molecules. The air we breath is from the remaining oxygen portion of H_2O. The carbon dioxide molecules are transformed into sugars during the dark reaction using the energy that was formed during the light reaction.

Dark Reaction

This part of the photosynthetic process is also called the Calvin Cycle. With one cycle of this reaction 3 carbon atoms are fixed or placed in a sugar molecule. This pathway is called C-3 photosynthesis. This is the way that most dicots or broadleaf plants make sugars during the dark reaction. C-3 photosynthesis has a disadvantage though. Oxygen competes with CO_2 for a binding site during the dark reaction. Sometimes sugars are not formed, but energy is still expended to complete the cycle. This is called photorespiration.

Another dark reaction pathway is called C-4 photosynthesis because 4 carbons are fixed or placed in a sugar molecule each time the cycle is completed. The dark reaction of C-4 photosynthesis occurs inside of specialized parts of leaf cells in the leaf called the bundle sheath, which exclude the presence of O_2. Because there is no oxygen present photorespiration does not occur. The C-4 photosynthetic pathway is what occurs in most monocots or grasses. This is a more efficient pathway and allows grasses to grow faster than broadleaf plants. Crassulacean acid metabolism or CAM photosynthesis is the dark reaction type found in many cactus, succulents, bromeliads, and orchids as well as a few other plants. CAM photosynthesis is similar to C-4 photosynthesis. However, CAM plants open their stomata only during the night to collect CO_2, when air temperatures are cooler, thus conserving water because of reduced transpiration. The CO_2 is converted into malic acid and then converted back to CO_2 during the day when light is present, thus producing sugars, while the stomata are closed and greatly reducing water loss.

Plants convert the energy from light into simple sugars, such as glucose. This food may be converted back to water and carbon dioxide, releasing the stored energy through a process called respiration. This energy is required for growth in nearly all organisms. Simple sugars are also converted to other sugars and starches (carbohydrates) which may be transported to the stems and roots for use or storage, or they may be used as building blocks for more complex structures, e.g. oils, pigments, proteins, cell walls, etc.

Photosynthesis is dependent on the availability of light. Generally speaking, as sunlight increases in intensity photosynthesis increases. This results in greater food production. Many garden crops, such as tomatoes,

respond best to maximum sunlight. Tomato production is cut drastically as light intensities drop. Only two or three varieties of "greenhouse" tomatoes will produce any fruit when sunlight is minimal in fall and spring.

Water plays an important role in photosynthesis in several ways. First, it maintains a plant's turgor or the firmness or fullness of plant tissue. Turgor pressure in a cell can be compared to air in an inflated balloon. Water pressure or turgor is needed in plant cells to maintain shape and ensure cell growth. Second, water is split into hydrogen and oxygen by the energy of the sun that has been absorbed by the chlorophyll in the plant leaves. The oxygen is released into the atmosphere and the hydrogen is used in manufacturing carbohydrates. Third, water dissolves minerals from the soil and transports them up from the roots and throughout the plant, where they serve as raw materials in the growth of new plant tissues.

The soil surrounding a plant should be moist, not too wet or too dry. Water is pulled through the plant by evaporation of water through the leaves (transpiration). Photosynthesis also requires carbon dioxide (CO_2) which enters the plant through the stomata. Carbon and oxygen are used in the manufacture of carbohydrates. Carbon dioxide in the air is 350 parts per million (ppm) or 0.035% at sea level and is plentiful enough so that it is not a limiting factor in plant growth. However, since carbon dioxide is consumed in making sugars and is not replenished by plants at a rapid rate, a tightly closed greenhouse in midwinter may not let in enough outside air to maintain an adequate carbon dioxide level. Under these conditions, improved crops of roses, carnations, tomatoes and certain other crops can be produced if the carbon dioxide level is raised with CO_2, generators or, in small greenhouses, with dry ice or a natural gas flame. Although not a direct component in photosynthesis, temperature is an important factor. Photosynthesis occurs at its highest rate in the temperature range of 65° to 85°F (18° to 27°C) and decreases when temperatures are above or below this range.

Respiration

Carbohydrates made during photosynthesis are of value to the plant when they are converted into energy. This energy is used in the process of building new tissues. The chemical process by which sugars and starches produced by photosynthesis are converted into energy is called respiration. It is similar to the burning of wood or coal to produce heat or energy. This process in cells is shown most simply as:

$$C_6H_{12}O_6 + 6O_2 \rightarrow 6CO_2 + 6H_2O + \text{Energy}$$

$$\text{Sugar} + \text{Oxygen} \rightarrow \text{Carbon dioxide} + \text{Water}$$

This equation is precisely the opposite of that used to illustrate photosynthesis, although more is involved than just reversing the reaction. However, it is appropriate to relate photosynthesis to a building process, while respiration is a breaking-down process.

Respiration

1. Uses food for plant energy.
2. Releases energy.
3. Occurs in all cells.
4. Uses oxygen.
5. Produces water.
6. Produces carbon dioxide.
7. Occurs in darkness as well as light.

If oxygen is limited or not present then anaerobic respiration or metabolism occurs. The by products of this reaction are ethyl alcohol or lactic acid and CO_2. This process is also know as fermentation or the Pasteur effect, (Louis Pasteur was the first to describe the effect), which is used to manufacture brewing and dairy products. It also occurs in muscle tissue when they are over exerted. The muscle burning we feel doing exercises is the accumulated lactic acid that forms in our tissue because of limited oxygen. Plant tissues undergo the same process, for example waterlogged soils limit the oxygen available to roots and may cause them to rot because of fermentation.

Photosynthesis

1. Produces food.
2. Stores energy.
3. Occurs in cells containing chloroplasts.
4. Releases oxygen.
5. Uses water.
6. Uses carbon dioxide.
7. Occurs in sunlight.

By now, it should be clear that respiration is the reverse of photosynthesis. Unlike photosynthesis, respiration occurs at night as well as during the day. Respiration occurs in all life forms and in all cells. The release of accumulated carbon dioxide and the uptake of oxygen occurs at the cell level. In animals, blood carries both oxygen and carbon dioxide to and from the atmosphere by means of the lungs, gills, spiracles etc. In plants there is simple diffusion into the open spaces within the leaf and exchange occurs through the stomata.

Transpiration

Transpiration is the process by which a plant loses water, primarily through leaf stomata. Transpiration is a necessary process that involves the use of about 90% of the water that enters the plant through the roots. The other 10% of the water is used in chemical reactions and in plant tissues. Transpiration is necessary for mineral transport from the soil to the plant for the cooling of the plant through evaporation, to move sugars and plant chemicals, and for the maintenance of turgor pressure. The amount of water lost from the plant depends on several environmental factors such as temperature, humidity and wind or air movement. An increase in temperature or air movement decreases relative humidity and causes the guard cells in the leaf to shrink, opening the stomata and increasing the rate of transpiration.

Environmental Factors that affect Plant Growth

Plant growth and distribution are limited by the environment. If any one environmental factor is less than ideal it will become a limiting factor in plant growth. Limiting factors are also responsible for the geography of plant distribution. For example, only plants adapted to limited amounts of water can live in deserts. Most plant problems are caused by environmental stress, either directly or indirectly. Therefore, it is important to understand the environmental aspects that affect plant growth. These factors are light, temperature, water (humidity), and nutrition.

Light

Light has three principal characteristics that affect plant growth: quantity, quality, and duration.

Light quantity refers to the intensity or concentration of sunlight and varies with the season of the year. The maximum is present in the summer and the minimum in winter. The more sunlight a plant receives (up to a point), the better capacity it has to produce plant food through photosynthesis. As the sunlight quantity decreases the photosynthetic process decreases. Light quantity can be decreased in a garden or greenhouse by using shade-cloth or shading paint above the plants. It can be increased by surrounding plants with white or reflective material or supplemental lights.

Light quality refers to the colour or wavelength reaching the plant surface. Sunlight can be broken up by a prism into respective colours of red, orange, yellow, green, blue, indigo, and violet. On a rainy day, raindrops act as tiny prisms and break the sunlight into these colours producing a rainbow. Red and blue light have the greatest effect on plant growth. Green light is least effective to plants as most plants reflect green

light and absorb very little. It is this reflected light that makes them appear green. Blue light is primarily responsible for vegetative growth or leaf growth. Red light when combined with blue light, encourages flowering in plants. Fluorescent or cool-white light is high in the blue range of light quality and is used to encourage leafy growth. These lights are excellent for starting seedlings. Incandescent light is high in the red or orange range but generally produces too much heat to be a valuable light source. Fluorescent "grow" lights have a mixture of red and blue colours that attempts to imitate sunlight as closely as possible. They are costly and generally not of any greater value than regular fluorescent lights.

Light duration or photoperiod refers to the amount of time that a plant is exposed to sunlight. When the concept of photoperiod was first recognized it was thought that the length of periods of light triggered flowering. The various categories of response were named according to the light length (i.e., short-day and long-day). It was then discovered that it is not the length of the light period but the length of uninterrupted dark periods that is critical to floral development. The ability of many plants to flower is controlled by photoperiod.

Plants can be classified into three categories, depending upon their flowering response to the duration of darkness. These are short-day, long-day, or day-neutral plants. Short-day, (long nights) plants form their flowers only when the day length is less than about 12 hours in duration. Short-day plants include many spring and fall flowering plants such as chrysanthemum and poinsettia. Long-day, (short nights) plants form flowers only when day lengths exceed 12 hours. They include almost all of the summer-flowering plants, as well as many vegetables including beet, radish, lettuce, spinach, and potato. Day-neutral plants form flowers regardless of day length. Some plants do not really fit into any category but may be responsive to combinations of day lengths. The petunia will flower regardless of day length, but flowers earlier and more profusely under long daylight. Since chrysanthemums flower under the short-day conditions of spring or fall the method for manipulating the plant into experiencing short days is very simple. If long days are predominant, a black plastic sheet is drawn over the chrysanthemum for 12 hours daily to block out light until flower buds are initiated. To bring a long-day plant into flower when sunlight is not present longer than 12 hours artificial light is added until flower buds are initiated.

Temperature

Temperature affects the productivity and growth of a plant depending upon whether the plant variety is a warm-season or cool-season crop. If temperatures are high and day length is long, cool-season crops such as

broccoli and spinach will bolt rather than produce the desired flower. Temperatures that are too low or high for a warm-season crop will prevent fruit set. Temperatures that are too high for warm-season crops such as pepper or tomato can cause pollen to become inviable and not pollinate flowers. Adverse temperatures also cause stunted growth and poor quality. For example, the bitterness in lettuce is caused by high temperatures.

Sometimes temperatures are used in connection with day length to manipulate the flowering of plants. Chrysanthemums will flower for a longer period of time if daylight temperatures are 59°F (15°C). The Christmas cactus forms flowers as a result of short days and low temperatures. Temperatures alone also influence flowering. Daffodils are forced to flower by putting the bulbs in cold storage in October at 35° to 40°F (2° to 4°C). The cold temperatures allow the bulb to mature. The bulbs are transferred to the greenhouse in midwinter where growth begins. The flowers are then ready for cutting in 3 to 4 weeks.

Thermoperiod refers to daily temperature change. Plants produce maximum growth when exposed to a day temperature that is about 10 to 15° F. (5.5 to 8°C) higher than the night temperature. This allows the plant to photosynthesize and respire during an optimum daytime temperature and to curtail the rate of respiration during a cooler night.

High temperatures cause increased respiration sometimes above the rate of photosynthesis. This means that the products of photosynthesis are being used more rapidly than they are being produced. For growth to occur photosynthesis must be greater than respiration.

Low temperatures can result in poor growth. Photosynthesis slows at low temperatures. Since photosynthesis is slowed, growth is slowed and this results in lower yields. Not all plants grow best in the same temperature range. For example, snapdragons grow best when nighttime temperatures are 55°F (12°C); the poinsettia prefers 62°F (17°C). Florist cyclamen does well under very cool conditions while many bedding plants prefer a higher temperature. Recently it has been found that roses can tolerate much lower nighttime temperatures than was previously believed. This has meant a conservation in energy for greenhouse growers. However, in some cases a certain number of days of low temperatures are needed by plants to grow properly. This is true of crops growing in cold regions of the country. Peaches are a prime example; most varieties require 700 to 1,000 hours below 45°F (7°C) and above 32°F (0°C) before they break their rest period and begin flowering and growth. If this cold requirement is not met then small, misshapen leaves and fruit will result. Many times fruit will not set. In low desert areas where these temperatures are not

experienced low chill peach trees should be planted. Lilies need 6 weeks at 33°F (1°C or below) before they will bloom.

Plants can be classified as either hardy or non-hardy depending upon their ability to withstand cold temperatures. Winter injury can occur to non-hardy plants if temperatures are too low or if unseasonably low temperatures occur late in the spring or early in the fall. Winter injury may also occur because of desiccation (drying out).

Plant roots need moist soil during the winter. When the soil is frozen the movement of water into the plant is severely restricted. On a windy winter day broad-leaved evergreens can become water-deficient in a few minutes, turning the leaves or needles brown. Wide variations in winter temperatures can cause premature bud break in some plants and consequent freezing damage. Late spring frost damage can ruin entire crops. If temperatures drop too low during the winter, entire trees of some species are killed by the freezing of plant cells and tissue.

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Substitution and Elimination Reactions of Amines

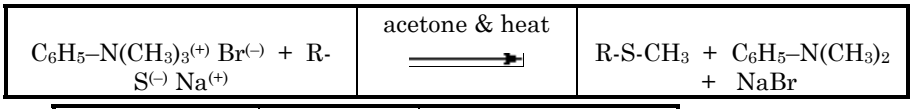
Dr. Kumar Mukesh Ranjan

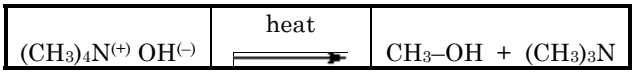
Dahiyawan, Chapra, Saran

Amine functions seldom serve as leaving groups in nucleophilic substitution or base-catalysed elimination reactions. Indeed, they are even less effective in this role than are hydroxyl and alkoxyl groups. In the case of alcohols and ethers, a useful technique for enhancing the reactivity of the oxygen function was to modify the leaving group ($OH^{(-)}$ or $OR^{(-)}$) to improve its stability as an anion (or equivalent). This stability is conveniently estimated from the strength of the corresponding conjugate acids.

As noted earlier, 1º and 2º-amines are much weaker acids than alcohols, so it is not surprising that it is difficult to force the nitrogen function to assume the role of a nucleophilic leaving group. For example, heating an amine with HBr or HI does not normally convert it to the corresponding alkyl halide, as in the case of alcohols and ethers. In this context we note that the acidity of the putative ammonium leaving group is at least ten powers of ten less than that of an analogous oxonium species. The loss of nitrogen from diazonium intermediates is a notable exception in this comparison, due to the extreme stability of this leaving group (the conjugate acid of N2 would be an extraordinarily strong acid).

One group of amine derivatives that have proven useful in SN2 and E2 reactions is that composed of the tetraalkyl (4º-) ammonium salts. Most applications involving this class of compounds are eliminations, but a few examples of SN2 substitution have been reported.

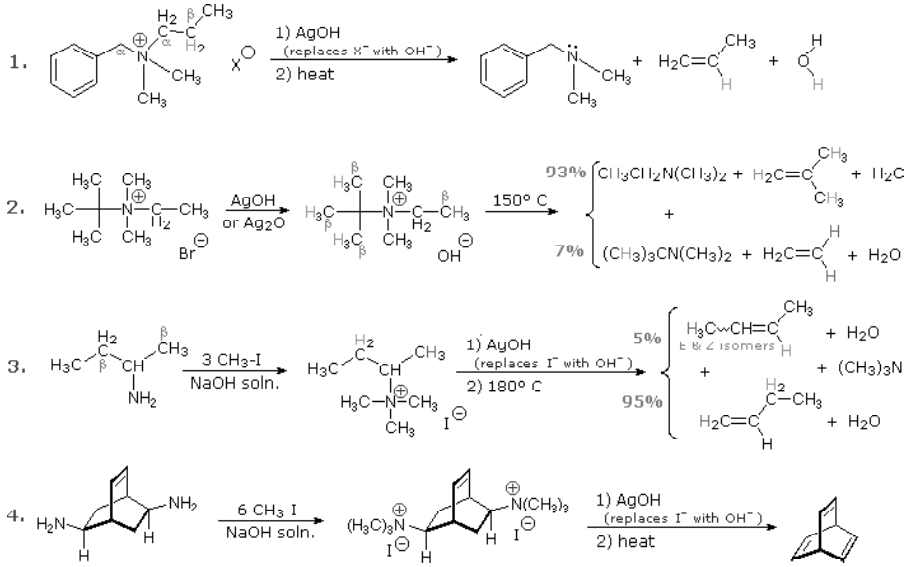




Hofmann Elimination

Elimination reactions of 4º-ammonium salts are termed Hofmann eliminations. Since the counter anion in most 4º-ammonium salts is halide, this is often replaced by the more basic hydroxide ion through reaction with silver hydroxide (or silver oxide). The resulting hydroxide salt must

then be heated (100 - 200 ºC) to effect the E2-like elimination of a 3º-amine. Example #1 below shows a typical Hofmann elimination. Obviously, for an elimination to occur one of the alkyl substituents on nitrogen must have one or more beta-hydrogens, as noted earlier in examining elimination reactions of alkyl halides.



In example #2 above, two of the alkyl substituents on nitrogen have beta-hydrogens, all of which are on methyl groups (coloured orange & magenta). The chief product from the elimination is the alkene having the more highly substituted double bond, reflecting not only the 3:1 numerical advantage of those beta-hydrogens, but also the greater stability of the double bond.

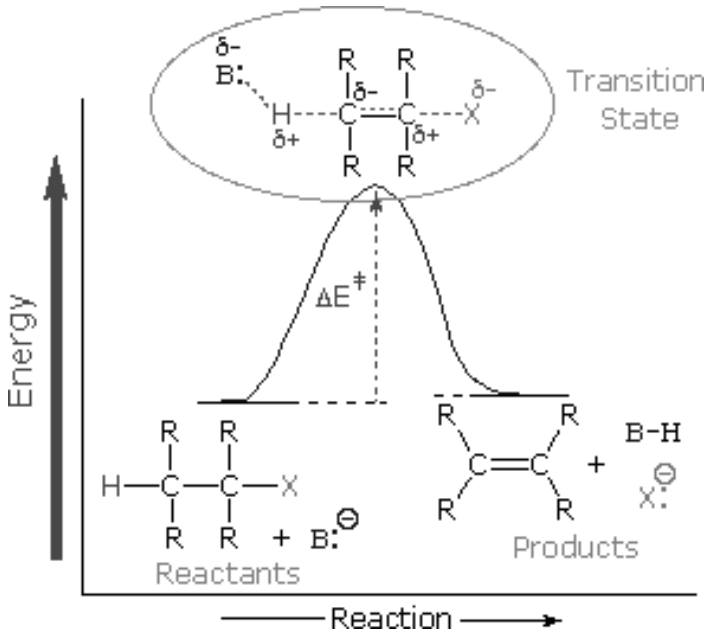
Example #3 illustrates two important features of the Hofmann elimination:

First, simple amines are easily converted to the necessary 4º-ammonium salts by exhaustive alkylation, usually with methyl iodide (methyl has no beta-hydrogens and cannot compete in the elimination reaction). Exhaustive methylation is shown again in example #4.

Second, when a given alkyl group has two different sets of beta-hydrogens available to the elimination process (coloured orange & magenta here), the major product is often the alkene isomer having the less substituted double bond.

The tendency of Hofmann eliminations to give the less-substituted double bond isomer is commonly referred to as the Hofmann Rule, and contrasts strikingly with the Zaitsev Rule formulated for dehydrohalogenations and dehydrations. In cases where other activating

groups, such as phenyl or carbonyl, are present, the Hofmann Rule may not apply. Thus, if 2-amino-1-phenylpropane is treated in the manner of example #3, the product consists largely of 1-phenylpropene (E & Z-isomers).

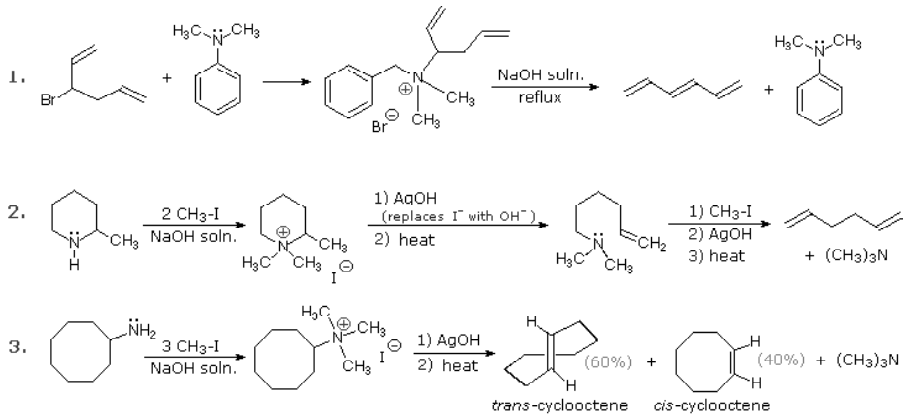


To understand why the base-induced elimination of 4º-ammonium salts behaves differently from that of alkyl halides it is necessary to reexamine the nature of the E2 transition state, first described for dehydrohalogenation. The energy diagram shown earlier for a single-step bimolecular E2 mechanism is repeated on the right. The E2 transition state is less well defined than is that of SN2 reactions. More bonds are being broken and formed, with the possibility of a continuum of states in which the extent of C-H and C-X bond-breaking and C=C bond-making varies. For example, if the bond to the leaving group (X) is substantially broken relative to the other bond changes, the transition state approaches that for an E1 reaction (initial ionization followed by a fast second step). At the other extreme, if the acidity of the beta-hydrogens is enhanced, then substantial breaking of C-H may occur before the other bonds begin to be affected. For most simple alkyl halides it was proper to envision a balanced transition state, in which there was a synchronous change in all the bonds. Such a model was consistent with the Zaitsev Rule.

When the leaving group X carries a positive charge, as do the 4º-ammonium compounds discussed here, the inductive influence of this charge will increase the acidity of both the alpha and the beta-hydrogens.

Furthermore, the 4º-ammonium substituent is much larger than a halide or hydroxyl group and may perturb the conformations available to substituted beta-carbons. It seems that a combination of these factors acts to favour base attack at the least substituted (least hindered and most acidic) set of beta-hydrogens. The favoured anti orientation of the leaving group and beta-hydrogen, noted for dehydrohalogenation, is found for many Hofmann eliminations; but syn-elimination is also common, possibly because the attraction of opposite charges orients the hydroxide base near the 4º-ammonium leaving group.

Three additional examples of the Hofmann elimination are shown in the following diagram. Example #1 is interesting in two respects. First, it generates a 4º-ammonium halide salt in a manner different from exhaustive methylation. Second, this salt is not converted to its hydroxide analog prior to elimination. A concentrated aqueous solution of the halide salt is simply dropped into a refluxing sodium hydroxide solution, and the volatile hydrocarbon product is isolated by distillation.



Example #2 illustrates an important aspect of the Hofmann elimination. If the nitrogen atom is part of a ring, then a single application of this elimination procedure does not remove the nitrogen as a separate 3º-amine product. In order to sever the nitrogen function from the molecule, a second Hofmann elimination must be carried out. Indeed, if the nitrogen atom was a member of two rings (fused or spiro), then three repetitions of the Hofmann elimination would be required to sever the nitrogen from the remaining molecular framework.

Example #3 is noteworthy because the less stable trans-cyclooctene is the chief product, accompanied by the cis-isomer. An anti-E2-transition state would necessarily give the cis-cycloalkene, so the trans-isomer must be generated by a syn-elimination. The cis-cyclooctene produced in this reaction could also be formed by a syn-elimination. Cyclooctane is a

conformationally complex structure. Several puckered conformations that avoid angle strain are possible, and one of the most stable of these is shown on the right. Some eclipsed bonds occur in all these conformers, and transannular hydrogen crowding is unavoidable.

Since the trimethylammonium substituent is large (about the size of tert-butyl) it will probably assume an equatorial-like orientation to avoid steric crowding. An anti-E2 transition state is likely to require an axial-like orientation of this bulky group, making this an unfavourable path.

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Applications in Geometric Measure Theory

Dr. Ashutosh Kumar
Lecturer, Department of Mathematics, Jai Prakash College, Ara

Numerous minimality results for complex analytic manifolds are based on the Wirtinger inequality for 2-forms. A succinct proof may be found in Herbert Federer's classic text Geometric Measure Theory. The Wirtinger inequality is also a key ingredient in Gromov's inequality for complex projective space in systolic geometry.

Divergence Theorem

In vector calculus, the divergence theorem, also known as Gauss's theorem or Ostrogradsky's theorem, is a result that relates the flow (that is, flux) of a vector field through a surface to the behaviour of the vector field inside the surface. More precisely, the divergence theorem states that the outward flux of a vector field through a closed surface is equal to the volume integral of the divergence over the region inside the surface. Intuitively, it states that the sum of all sources minus the sum of all sinks gives the net flow out of a region.

The divergence theorem is an important result for the mathematics of engineering, in particular in electrostatics and fluid dynamics.

In physics and engineering, the divergence theorem is usually applied in three dimensions. However, it generalizes to any number of dimensions. In one dimension, it is equivalent to the fundamental theorem of calculus. In two dimensions, it is equivalent to Green's theorem. The theorem is a special case of the more general Stokes' theorem.

Intuition

If a fluid is flowing in some area, and we wish to know how much fluid flows out of a certain region within that area, then we need to add up the sources inside the region and subtract the sinks. The fluid flow is represented by a vector field, and the vector field's divergence at a given point describes the strength of the source or sink there. So, integrating the field's divergence over the interior of the region should equal the integral of the vector field over the region's boundary. The divergence theorem says that this is true. The divergence theorem is thus a conservation law which states that the volume total of all sinks and sources, that is the volume integral of the divergence, is equal to the net flow across the volume's boundary.

Mathematical Statement

Suppose V is a subset of R^n (in the case of $n = 3$, V represents a volume in 3D space) which is compact and has a piecewise smooth boundary S (also indicated with $\partial V=S$). If F is a continuously differen-tiable vector field defined on a neighbourhood of V, then we have

$$\iiint_V (\nabla \cdot F) dV = \oiint_s (F \cdot n) dS.$$

The left side is a volume integral over the volume V, the right side is the surface integral over the boundary of the volume V. The closed manifold ∂V is quite generally the boundary of V oriented by outward-pointing normals, and n is the outward pointing unit normal field of the boundary ∂V. (dS may be used as a shorthand for n dS.) By the symbol within the two integrals it is stressed once more that ∂V is a closed surface. In terms of the intuitive description above, the left-hand side of the equation represents the total of the sources in the volume V, and the right-hand side represents the total flow across the boundary ∂V.

Corollaries

By applying the divergence theorem in various contexts, other useful identities can be derived (cf. vector identities).

- Applying the divergence theorem to the product of a scalar function g and a vector field F, the result is

$$\iiint_V \left[F \cdot (\nabla g) + g (\nabla \cdot F) \right] dV = \oiint_{S.} gF \cdot n dS.$$

A special case of this is $F = \nabla f$, in which case the theorem is the basis for Green's identities.

- Applying the divergence theorem to the cross-product of two vector fields $F \times G$, the result is

$$\iiint_V \left[G \cdot (\nabla \times F) - F \cdot (\nabla \times G) \right] dV = \oiint_{SF \times G \cdot dS.}$$

- Applying the divergence theorem to the product of a scalar function, f, and a non-zero constant vector, the following theorem can be proven:

$$\iiint_V \nabla f \, dV = \oiint_{S.} \oiint_{sfdS}$$

- Applying the divergence theorem to the cross-product of a vector field F and a non-zero constant vector, the following theorem can be proven:

$$\iiint_V \nabla \times F \, dV = \oiint_{sdS \times F}$$

Analytic Geometry

Analytic geometry, or analytical geometry, has two different meanings in mathematics. The modern and advanced meaning refers to the geometry of analytic varieties. This article focuses on the classical and elementary meaning.

In classical mathematics, analytic geometry, also known as coordinate geometry, or Cartesian geometry, is the study of geometry using a coordinate system and the principles of algebra and analysis. This contrasts with the synthetic approach of Euclidean geometry, which treats certain geometric notions as primitive, and uses deductive reasoning based on axioms and theorems to derive truth. Analytic geometry is widely used in physics and engineering, and is the foundation of most modern fields of geometry, including algebraic, differential, discrete, and computational geometry.

Usually the Cartesian coordinate system is applied to manipulate equations for planes, straight lines, and squares, often in two and sometimes in three dimensions. Geometrically, one studies the Euclidean plane (2 dimensions) and Euclidean space (3 dimensions). As taught in school books, analytic geometry can be explained more simply: it is concerned with defining and representing geometrical shapes in a numerical way and extracting numerical information from shapes' numerical definitions and representations. The numerical output, however, might also be a vector or a shape. That the algebra of the real numbers can be employed to yield results about the linear continuum of geometry relies on the Cantor–Dedekind axiom.

History

The Greek mathematician Menaechmus solved problems and proved theorems by using a method that had a strong resemblance to the use of coordinates and it has sometimes been maintained that he had introduced analytic geometry. Apollonius of Perga, in On Determinate Section, dealt with problems in a manner that may be called an analytic geometry of one dimension; with the question of finding points on a line that were in a ratio to the others. Apollonius in the Conics further developed a method that is so similar to analytic geometry that his work is sometimes thought to have anticipated the work of Descartes — by some 1800 years. His application of reference lines, a diameter and a tangent is essentially no different from our modern use of a coordinate frame, where the distances measured along the diameter from the point of tangency are the abscissas, and the segments parallel to the tangent and intercepted between the axis

and the curve are the ordinates. He further developed relations between the abscissas and the corresponding ordinates that are equivalent to rhetorical equations of curves.

However, although Apollonius came close to developing analytic geometry, he did not manage to do so since he did not take into account negative magnitudes and in every case the coordinate system was superimposed upon a given curve a posteriori instead of a priori. That is, equations were determined by curves, but curves were not determined by equations. Coordinates, variables, and equations were subsidiary notions applied to a specific geometric situation.

The eleventh century Persian mathematician Omar Khayyám saw a strong relationship between geometry and algebra, and was moving in the right direction when he helped to close the gap between numerical and geometric algebra with his geometric solution of the general cubic equations, but the decisive step came later with Descartes.

Analytic geometry has traditionally been attributed to René Descartes. Descartes made significant progress with the methods in an essay entitled La Geometrie (Geometry), one of the three accompanying essays (appendices) published in 1637 together with his Discourse on the Method for Rightly Directing One's Reason and Searching for Truth in the Sciences, commonly referred to as Discourse on Method. This work, written in his native French tongue, and its philosophical principles, provided a foundation for Infinitesimal calculus in Europe. Initially the work was not well received, due, in part, to the many gaps in arguments and complicated equations. Only after the translation into Latin and the addition of commentary by van Schooten in 1649 (and further work thereafter) did Descarte's masterpiece receive due recognition.

Pierre Fermat also pioneered the development of analytic geometry. Although not published in his lifetime, a manuscript form of Ad locos planos et solidos isagoge (Introduction to Plane and Solid Loci) was circulating in Paris in 1637, just prior to the publication of Descartes' Discourse. Clearly written and well received, the Introduction also laid the groundwork for analytical geometry. The key difference between Fermat's and Descartes' treatments is a matter of viewpoint. Fermat always started with an algebraic equation and then described the geometric curve which satisfied it, while Descartes starts with geometric curves and produces their equations as one of several properties of the curves. As a consequence of this approach, Descartes had to deal with more complicated equations and he had to develop the methods to work with polynomial equations of higher degree.

Basic Principles

Coordinates

In analytic geometry, the plane is given a coordinate system, by which every point has a pair of real number coordinates. The most common coordinate system to use is the Cartesian coordinate system, where each point has an x-coordinate representing its horizontal position, and a y-coordinate representing its vertical position. These are typically written as an ordered pair (x, y). This system can also be used for three-dimensional geometry, where every point in Euclidean space is represented by an ordered triple of coordinates (x, y, z).

Other coordinate systems are possible. On the plane the most common alternative is polar coordinates, where every point is represented by its radius r from the origin and its angle θ. In three dimensions, common alternative coordinate systems include cylindrical coordinates and spherical coordinates.

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Organic Chemistry and Organic Compound

Smt. Nisha Kumari Mishra
M.Sc. Chemistry, M.U. Bodh Gaya,
Principal, Sanjay Gandhi Mahila (Inter Section) Gosai Bagh, Gaya

Organic chemistry is a chemistry subdiscipline involving the scientific study of the structure, properties, and reactions of organic compounds and organic materials, i.e., matter in its various forms that contain carbon atoms. Study of structure includes using spectroscopy (e.g., NMR), mass spectrometry, and other physical and chemical methods to determine the chemical composition and constitution of organic compounds and materials. Study of properties includes both physical properties and chemical properties, and uses similar methods as well as methods to evaluate chemical reactivity, with the aim to understand the behaviour of the organic matter in its pure form (when possible), but also in solutions, mixtures, and fabricated forms. The study of organic reactions includes probing their scope through use in preparation of target compounds (e.g., natural products, drugs, polymers, etc.) by chemical synthesis, as well as the focused study of the reactivities of individual organic molecules, both in the laboratory and via theoretical (in silico) study.

An organic compound is any member of a large class of gaseous, liquid, or solid chemical compounds whose molecules contain carbon. For historical reasons discussed below, a few types of carbon-containing compounds such as carbides, carbonates, simple oxides of carbon (such as CO and CO_2), and cyanides are considered inorganic. The distinction between *organic* and *inorganic* carbon compounds, while "useful in organizing the vast subject of chemistry... is somewhat arbitrary."

Organic chemistry is the science concerned with all aspects of organic compounds. Organic synthesis is the methodology of their preparation.

Characterization

Since organic compounds often exist as mixtures, a variety of techniques have also been developed to assess purity, especially important being chromatography techniques such as HPLC and gas chromatography. Traditional methods of separation include distillation, crystallization, and solvent extraction.

Organic compounds were traditionally characterized by a variety of chemical tests, called "wet methods", but such tests have been largely displaced by spectroscopic or other computer-intensive methods of analysis. Listed in approximate order of utility, the chief analytical methods are:

- Nuclear magnetic resonance (NMR) spectroscopy is the most commonly used technique, often permitting complete assignment of atom connectivity and even stereochemistry using correlation spectroscopy. The principal constituent atoms of organic chemistry - hydrogen and carbon - exist naturally with NMR-responsive isotopes, respectively ^{1}H and ^{13}C.
- Elemental analysis: A destructive method used to determine the elemental composition of a molecule. See also mass spectrometry, below.
- Mass spectrometry indicates the molecular weight of a compound and, from the fragmentation patterns, its structure. High resolution mass spectrometry can usually identify the exact formula of a compound and is used in lieu of elemental analysis. In former times, mass spectrometry was restricted to neutral molecules exhibiting some volatility, but advanced ionization techniques allow one to obtain the "mass spec" of virtually any organic compound.
- Crystallography is an unambiguous method for determining molecular geometry, the proviso being that single crystals of the material must be available and the crystal must be representative of the sample. Highly automated software allows a structure to be determined within hours of obtaining a suitable crystal.

Traditional spectroscopic methods such as infrared spectroscopy, optical rotation, UV/VIS spectroscopy provide relatively nonspecific structural information but remain in use for specific classes of compounds.

Properties

Physical properties of organic compounds typically of interest include both quantitative and qualitative features. Quantitative information includes melting point, boiling point, and index of refraction. Qualitative properties include odour, consistency, solubility, and colour.

Melting and Boiling Properties

Organic compounds typically melt and many boil. In contrast, while inorganic materials generally can be melted, many do not boil, tending instead to degrade. In earlier times, the melting point (m.p.) and boiling

point (b.p.) provided crucial information on the purity and identity of organic compounds. The melting and boiling points correlate with the polarity of the molecules and their molecular weight. Some organic compounds, especially symmetrical ones, sublime, that is they evaporate without melting. A well-known example of a sublimable organic compound is para-dichlorobenzene, the odiferous constituent of modern mothballs. Organic compounds are usually not very stable at temperatures above 300 °C, although some exceptions exist.

Solubility

Neutral organic compounds tend to be hydrophobic; that is, they are less soluble in water than in organic solvents. Exceptions include organic compounds that contain ionizable groups as well as low molecular weight alcohols, amines, and carboxylic acids where hydrogen bonding occurs. Organic compounds tend to dissolve in organic solvents. Solvents can be either pure substances like ether or ethyl alcohol, or mixtures, such as the paraffinic solvents such as the various petroleum ethers and white spirits, or the range of pure or mixed aromatic solvents obtained from petroleum or tar fractions by physical separation or by chemical conversion. Solubility in the different solvents depends upon the solvent type and on the functional groups if present.

Solid State Properties

Various specialized properties of molecular crystals and organic polymers with conjugated systems are of interest depending on applications, e.g. thermo-mechanical and electro-mechanical such as piezoelectricity, electrical conductivity, and electro-optical (e.g. non-linear optics) properties. For historical reasons, such properties are mainly the subjects of the areas of polymer science and materials science.

Nomenclature

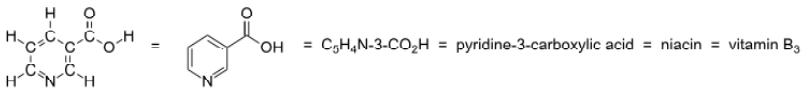

Figure: *Various names and depictions for one organic compound.*

The names of organic compounds are either systematic, following logically from a set of rules, or nonsystematic, following various traditions. Systematic nomenclature is stipulated by specifications from IUPAC. Systematic nomenclature starts with the name for a parent structure within the molecule of interest. This parent name is then modified by prefixes, suffixes, and numbers to unambiguously convey the structure. Given that millions of organic compounds are known, rigorous use of systematic names can be cumbersome. Thus, IUPAC recommendations are more

closely followed for simple compounds, but not complex molecules. To use the systematic naming, one must know the structures and names of the parent structures. Parent structures include unsubstituted hydrocarbons, heterocycles, and monofunctionalized derivatives thereof.

Nonsystematic nomenclature is simpler and unambiguous, at least to organic chemists. Nonsystematic names do not indicate the structure of the compound. They are common for complex molecules, which includes most natural products. Thus, the informally named lysergic acid diethylamide is systematically named (6a*R*,9*R*)-*N*,*N*-diethyl-7-methyl-4,6,6a,7,8,9-hexahydroindolo-[4,3-*fg*] quinoline-9-carboxamide.

With the increased use of computing, other naming methods have evolved that are intended to be interpreted by machines. Two popular formats are SMILES and InChI.

Structural Drawings

Organic molecules are described more commonly by drawings or structural formulas, combinations of drawings and chemical symbols. The line-angle formula is simple and unambiguous. In this system, the endpoints and intersections of each line represent one carbon, and hydrogen atoms can either be notated explicitly or assumed to be present as implied by tetravalent carbon. The depiction of organic compounds with drawings is greatly simplified by the fact that carbon in almost all organic compounds has four bonds, nitrogen three, oxygen two, and hydrogen one.

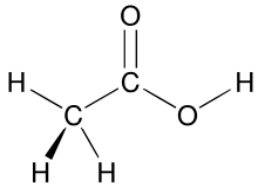

Figure: *The family of carboxylic acids contains a carboxyl (-COOH) functional group. Acetic acid, shown here, is an example.*

The concept of functional groups is central in organic chemistry, both as a means to classify structures and for predicting properties. A functional group is a molecular module, and the reactivity of that functional group is assumed, within limits, to be the same in a variety of molecules. Functional groups can have decisive influence on the chemical and physical properties of organic compounds. Molecules are classified on the basis of their functional groups. Alcohols, for example, all have the subunit C-O-H. All alcohols tend to be somewhat hydrophilic, usually form esters, and usually can be converted to the corresponding halides. Most functional groups feature heteroatoms (atoms other than C and H).

Organic compounds are classified according to functional groups, alcohols, carboxylic acids, amines, etc.

Aliphatic Compounds

The aliphatic hydrocarbons are subdivided into three groups of homologous series according to their state of saturation:

- paraffins, which are alkanes without any double or triple bonds,
- olefins or alkenes which contain one or more double bonds, i.e. di-olefins (dienes) or poly-olefins.
- alkynes, which have one or more triple bonds.

The rest of the group is classed according to the functional groups present. Such compounds can be "straight-chain", branched-chain or cyclic. The degree of branching affects characteristics, such as the octane number or cetane number in petroleum chemistry.

Both saturated (alicyclic) compounds and unsaturated compounds exist as cyclic derivatives. The most stable rings contain five or six carbon atoms, but large rings (macrocycles) and smaller rings are common. The smallest cycloalkane family is the three-membered cyclopropane ($(CH_2)_3$). Saturated cyclic compounds contain single bonds only, whereas aromatic rings have an alternating (or conjugated) double bond. Cycloalkanes do not contain multiple bonds, whereas the cycloalkenes and the cycloalkynes do.

Aromatic Compounds

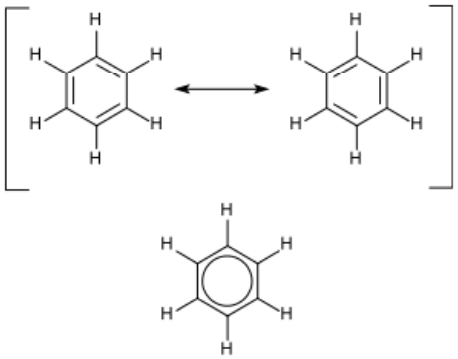

Figure: *Benzene is one of the best-known aromatic compounds as it is one of the simplest and most stable aromatics.*

Aromatic hydrocarbons contain conjugated double bonds. This means that every carbon atom in the ring is sp2 hybridized, allowing for added stability. The most important example is benzene, the structure of which was formulated by Kekulé who first proposed the delocalization or resonance principle for explaining its structure. For "conventional" cyclic compounds, aromaticity is conferred by the presence of 4n + 2 delocalized

pi electrons, where n is an integer. Particular instability (antiaromaticity) is conferred by the presence of 4n conjugated pi electrons.

Heterocyclic Compounds

The characteristics of the cyclic hydrocarbons are again altered if heteroatoms are present, which can exist as either substituents attached externally to the ring (exocyclic) or as a member of the ring itself (endocyclic). In the case of the latter, the ring is termed a heterocycle. Pyridine and furan are examples of aromatic heterocycles while piperidine and tetrahydrofuran are the corresponding alicyclic heterocycles. The heteroatom of heterocyclic molecules is generally oxygen, sulfur, or nitrogen, with the latter being particularly common in biochemical systems.

Examples of groups among the heterocyclics are the aniline dyes, the great majority of the compounds discussed in biochemistry such as alkaloids, many compounds related to vitamins, steroids, nucleic acids (e.g. DNA, RNA) and also numerous medicines. Heterocyclics with relatively simple structures are pyrrole (5-membered) and indole (6-membered carbon ring). Rings can fuse with other rings on an edge to give polycyclic compounds. The purine nucleoside bases are notable polycyclic aromatic heterocycles. Rings can also fuse on a "corner" such that one atom (almost always carbon) has two bonds going to one ring and two to another. Such compounds are termed spiro and are important in a number of natural products.

Polymers

One important property of carbon is that it readily forms chains, or networks, that are linked by carbon-carbon (carbon to carbon) bonds. The linking process is called polymerization, while the chains, or networks, are called polymers. The source compound is called a monomer.

Two main groups of polymers exist: synthetic polymers and biopolymers. Synthetic polymers are artificially manufactured, and are commonly referred to as industrial polymers. Biopolymers occur within a respectfully natural environment, or without human intervention.

Since the invention of the first synthetic polymer product, bakelite, synthetic polymer products have frequently been invented. Common synthetic organic polymers are polyethylene (polythene), polypropylene, nylon, teflon (PTFE), polystyrene, polyesters, polymethylmethacrylate (called perspex and plexiglas), and polyvinylchloride (PVC).

Both synthetic and natural rubber are polymers.

Varieties of each synthetic polymer product may exist, for purposes of a specific use. Changing the conditions of polymerization alters the

chemical composition of the product and its properties. These alterations include the chain length, or branching, or the tacticity.

With a single monomer as a start, the product is a homopolymer. Secondary component(s) may be added to create a heteropolymer (co-polymer) and the degree of clustering of the different components can also be controlled. Physical characteristics, such as hardness, density, mechanical or tensile strength, abrasion resistance, heat resistance, transparency, colour, etc. will depend on the final composition.

Biomolecules

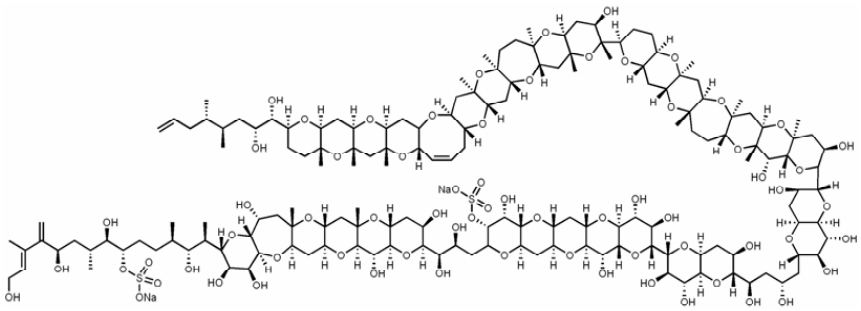

Figure: *Maitotoxin, a complex organic biological toxin.*

Biomolecular chemistry is a major category within organic chemistry which is frequently studied by biochemists. Many complex multi-functional group molecules are important in living organisms. Some are long-chain biopolymers, and these include peptides, DNA, RNA and the polysaccharides such as starches in animals and celluloses in plants. The other main classes are amino acids (monomer building blocks of peptides and proteins), carbohydrates (which includes the polysaccharides), the nucleic acids (which include DNA and RNA as polymers), and the lipids. In addition, animal biochemistry contains many small molecule intermediates which assist in energy production through the Krebs cycle, and produces isoprene, the most common hydrocarbon in animals. Isoprenes in animals form the important steroid structural (cholesterol) and steroid hormone compounds; and in plants form terpenes, terpenoids, some alkaloids, and a class of hydrocarbons called biopolymer polyisoprenoids present in the latex of various species of plants, which is the basis for making rubber.

Small Molecules

In pharmacology, an important group of organic compounds is small molecules, also referred to as 'small organic compounds'. In this context, a small molecule is a small organic compound that is biologically active,

but is not a polymer. In practice, small molecules have a molar mass less than approximately 1000 g/mol.

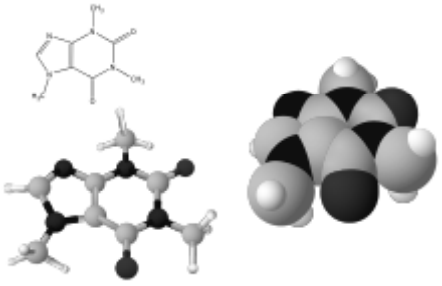

Figure: *Molecular models of caffeine.*

Others

Organic compounds containing bonds of carbon to nitrogen, oxygen and the halogens are not normally grouped separately. Others are sometimes put into major groups within organic chemistry and discussed under titles such as organosulfur chemistry, organometallic chemistry, organophosphorus chemistry and organosilicon chemistry.

Organic Synthesis

Organic synthesis of a novel compound is a problem solving task, where a synthesis is designed for a target molecule by selecting optimal reactions from optimal starting materials. Complex compounds can have tens of reaction steps that sequentially build the desired molecule. The synthesis proceeds by utilizing the reactivity of the functional groups in the molecule. For example, a carbonyl compound can be used as a nucleophile by converting it into an enolate, or as an electrophile; the combination of the two is called the aldol reaction. Designing practically useful syntheses always requires conducting the actual synthesis in the laboratory. The scientific practice of creating novel synthetic routes for complex molecules is called total synthesis.

Strategies to design a synthesis include retrosynthesis, popularized by E.J. Corey, starts with the target molecule and splices it to pieces according to known reactions. The pieces, or the proposed precursors, receive the same treatment, until available and ideally inexpensive starting materials are reached. Then, the retrosynthesis is written in the opposite direction to give the synthesis. A "synthetic tree" can be constructed, because each compound and also each precursor has multiple syntheses.

Organic Reactions

Organic reactions are chemical reactions involving organic compounds. Many of these reactions are associated with functional groups.

The general theory of these reactions involves careful analysis of such properties as the electron affinity of key atoms, bond strengths and steric hindrance. These factors can determine the relative stability of short-lived reactive intermediates, which usually directly determine the path of the reaction.

The basic reaction types are: addition reactions, elimination reactions, substitution reactions, pericyclic reactions, rearrangement reactions and redox reactions. An example of a common reaction is a substitution reaction written as:

$$Nu^- + C\text{-}X \rightarrow C\text{-}Nu + X^-$$

where X is some functional group and Nu is a nucleophile.

The number of possible organic reactions is basically infinite. However, certain general patterns are observed that can be used to describe many common or useful reactions. Each reaction has a stepwise reaction mechanism that explains how it happens in sequence—although the detailed description of steps is not always clear from a list of reactants alone.

The stepwise course of any given reaction mechanism can be represented using arrow pushing techniques in which curved arrows are used to track the movement of electrons as starting materials transition through intermediates to final products.

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Product Lifecycle Management

Rajiv Ranjan

Research Scholar, Department of Mechanical,
Shri Venkateshwara University, Gajraula, UP

In industry, product lifecycle management (PLM) is the process of managing the entire lifecycle of a product from inception, through engineering design and manufacture, to service and disposal of manufactured products. PLM integrates people, data, processes and business systems and provides a product information backbone for companies and their extended enterprise.

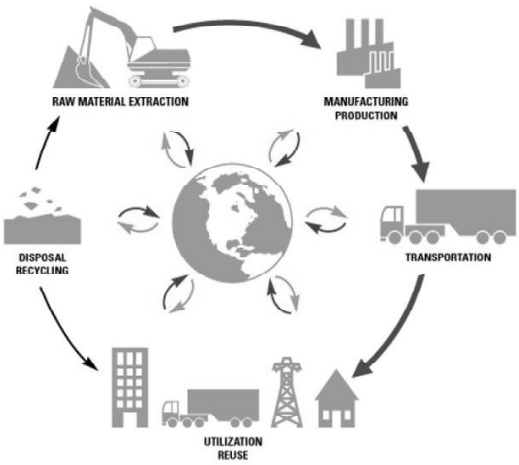


Figure: *A generic lifecycle of products*

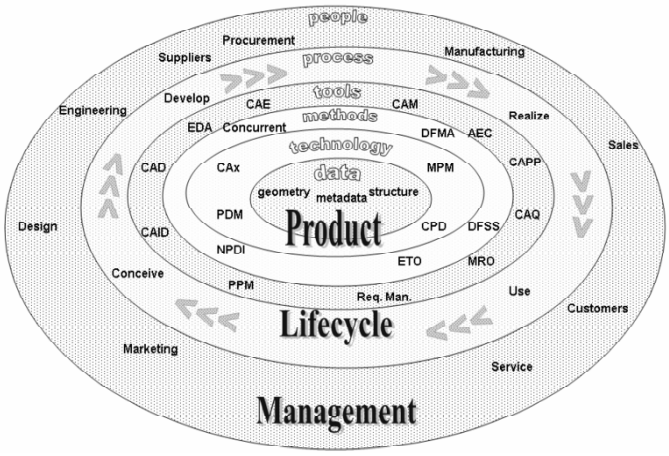
The inspiration for the burgeoning business process now known as PLM came from American Motors Corporation (AMC). The automaker was looking for a way to speed up its product development process to compete better against its larger competitors in 1985, according to François Castaing, Vice President for Product Engineering and Development. After introducing its compact Jeep Cherokee (XJ), the vehicle that launched the modern sport utility vehicle (SUV) market, AMC began development of a new model, that later came out as the Jeep Grand Cherokee. The first part in its quest for faster product development was computer-aided design (CAD) software system that make engineers more productive. The second part in this effort was the new communication system that allowed conflicts to be resolved faster, as well as reducing costly engineering changes because all drawings and documents were in a central database. The product data management was so effective that after AMC was purchased

by Chrysler, the system was expanded throughout the enterprise connecting everyone involved in designing and building products. While an early adopter of PLM technology, Chrysler was able to become the auto industry's lowest-cost producer, recording development costs that were half of the industry average by the mid-1990s.

Forms

PLM systems help organizations in coping with the increasing complexity and engineering challenges of developing new products for the global competitive markets. Product lifecycle management (PLM) should be distinguished from 'product life cycle management (marketing)' (PLCM). PLM describes the engineering aspect of a product, from managing descriptions and properties of a product through its development and useful life; whereas, PLCM refers to the commercial management of life of a product in the business market with respect to costs and sales measures.

Product lifecycle management can be considered one of the four cornerstones of a manufacturing corporation's information technology structure. All companies need to manage communications and information with their customers (CRM-customer relationship management), their suppliers and fulfillment (SCM-supply chain), their resources within the enterprise (ERP-enterprise resource planning) and their product planning and development (PLM).



One form of PLM is called people-centric PLM. While traditional PLM tools have been deployed only on release or during the release phase, people-centric PLM targets the design phase.

The core of PLM (product lifecycle management) is in the creation and central management of all product data and the technology used to

access this information and knowledge. PLM as a discipline emerged from tools such as CAD, CAM and PDM, but can be viewed as the integration of these tools with methods, people and the processes through all stages of a product's life. It is not just about software technology but is also a business strategy.

For simplicity the stages described are shown in a traditional sequential engineering workflow. The exact order of event and tasks will vary according to the product and industry in question but the main processes are:

- Conceive
  - o Specification
  - o Concept design
- Design
  - o Detailed design
  - o Validation and analysis (simulation)
  - o Tool design
- Realize
  - o Plan manufacturing
  - o Manufacture
  - o Build/Assemble
  - o Test (quality check)
- Service
  - o Sell and deliver
  - o Use
  - o Maintain and support
  - o Dispose

The major key point events are:

- Order
- Idea
- Kickoff
- Design freeze
- Launch

The reality is however more complex, people and departments cannot perform their tasks in isolation and one activity cannot simply finish and the next activity start. Design is an iterative process, often designs need to be modified due to manufacturing constraints or conflicting requirements. Where a customer order fits into the time line depends on the industry type and whether the products are for example, built to order, engineered to order, or assembled to order.

Phases of Product Lifecycle and Corresponding Technologies

Many software solutions have developed to organize and integrate the different phases of a product's lifecycle. PLM should not be seen as a single software product but a collection of software tools and working methods integrated together to address either single stages of the lifecycle or connect different tasks or manage the whole process. Some software providers cover the whole PLM range while others single niche application. Some applications can span many fields of PLM with different modules within the same data model. An overview of the fields within PLM is covered here. It should be noted however that the simple classifications do not always fit exactly, many areas overlap and many software products cover more than one area or do not fit easily into one category. It should also not be forgotten that one of the main goals of PLM is to collect knowledge that can be reused for other projects and to coordinate simultaneous concurrent development of many products. It is about business processes, people and methods as much as software application solutions. Although PLM is mainly associated with engineering tasks it also involves marketing activities such as product portfolio management (PPM), particularly with regards to new product development (NPD). There are several life-cycle models in industry to consider, but most are rather similar.

These techniques include:-

- Concurrent engineering workflow
- Industrial design
- Bottom–up design
- Top–down design
- Front-loading design workflow
- Design in context
- Modular design
- NPD new product development
- DFSS design for Six Sigma
- DFMA design for manufacture / assembly
- Digital simulation engineering
- Requirement-driven design
- Specification-managed validation
- Configuration management

Market Size

Total spending on PLM software and services was estimated in 2006 to be above $30 billion a year. Market growth estimates are in the area of 10%.

Pyramid of Production Systems

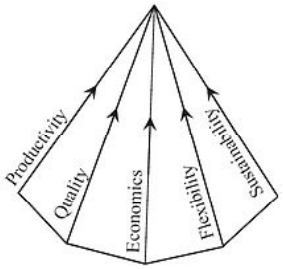


Figure: *Pyramid of Production Systems*

According to Malakooti (2013), there are five long-term objectives that should be considered in production systems:

- Cost which can be measured in terms of monetary units and usually consists of fixed and variable cost.
- Productivity which can be measured in terms of the number of products produced during a period of time.
- Quality which can be measured, for example, in terms of customers' satisfaction.
- Flexibility, for example, ability of the system to produce variety of products.
- Ecological Soundness which can be measured in terms of biological and environmental impacts of the production system.

The relation between these five objects can be presented as pyramid which its tip is associated with the highest productivity, the highest quality, the most economical, the most flexibility, and the most sustainability. The points inside of this pyramid are associated with different combinations of five criteria. The tip of the pyramid is the ideal point but it is infeasible and the base of pyramid is consists of the worst points.

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Information and Communications Technology

Neeraj Aggarwal

Research Scholar, Department of Computer/IT, CMJ University, Meghalaya

Information and communications technology (ICT) is often used as an extended synonym for information technology (IT), but is a more specific term that stresses the role of unified communications and the integration of telecommunications (telephone lines and wireless signals), computers as well as necessary enterprise software, middleware, storage, and audio-visual systems, which enable users to access, store, transmit, and manipulate information. The term *ICT* is also used to refer to the convergence of audio-visual and telephone networks with computer networks through a single cabling or link system. There are large economic incentives (huge cost savings due to elimination of the telephone network) to merge the telephone network with the computer network system using a single unified system of cabling, signal distribution and management.

The term infocommunications is sometimes used interchangeably with ICT. Infocommunications is the expansion of telecommunications with information processing and content handling functions on a common digital technology base.

Global Costs of IT

The money spent on IT worldwide has been most recently estimated as US $3.5 trillion and is currently growing at 5% per year – doubling every 15 years. The 2014 IT budget of US federal government is nearly $82 billion. IT costs, as a percentage of corporate revenue, have grown 50% since 2002, putting a strain on IT budgets. When looking at current companies' IT budgets, 75% are recurrent costs, used to "keep the lights on" in the IT department, and 25% are cost of new initiatives for technology development.

The average IT budget has the following breakdown:

- 31% personnel costs (internal)
- 29% software costs (external/purchasing category)
- 26% hardware costs (external/purchasing category)
- 14% costs of external service providers (external/services).

Information and Communication Technologies for Development

Information and Communication Technologies for Development (ICT4D) refers to the use of Information and Communication Technologies

(ICTs) in the fields of socioeconomic development, international development and human rights. The theory behind this is that more and better information and communication furthers the development of a society.

Aside from its reliance on technology, ICT4D also requires an understanding of community development, poverty, agriculture, healthcare, and basic education. This makes ICT4D appropriate technology and if it is shared openly open source appropriate technology. Richard Heeks suggests that the I in ICT4D is related with "library and information sciences", the C is associated with "communication studies", the T is linked with "information systems", and the D for "development studies". It is aimed at bridging the digital divide and aid economic development by fostering equitable access to modern communications technologies. It is a powerful tool for economic and social development. Other terms can also be used for "ICT4D" or "ICT4Dev" ("ICT for development") like ICTD ("ICT and development", which is used in a broader sense) and development informatics.

ICT4D can refer to assisting disadvantaged populations anywhere in the world, but it is usually associated with applications in developing countries. It is concerned with directly applying information technology approaches to poverty reduction. ICTs can be applied directly, wherein its use directly benefits the disadvantaged population, or indirectly, wherein it can assist aid organisations or non-governmental organizations or governments or businesses to improve socio-economic conditions.

The field is an interdisciplinary research area through the growing number of conferences, workshops and publications. This is partly due to the need for scientifically validated benchmarks and results, that can measure the effectiveness of current projects. This field has also produced an informal community of technical and social science researchers who rose out of the annual ICT4D conferences.

In line with the Schumpeterian school of thought, the first enabling factor for the associated socio-economic transformations is the existence technological infrastructure: hardware infrastructure and generic software services. Additionally, capacity and knowledge are the human requirements to make use of these technologies. These foundations are the basis for the digitization of information flows and communication mechanisms in different sectors of society. When part of the information flows and communication processes in these sectors are carried out in electronic networks, the prefix "e-" is often added to the sector's name, resulting in e-government, e-business and e-commerce, e-health, and e-learning, etc. (vertical blue dimension in Figure). This process of transformation represent the basic requirements and building blocks, but they are not sufficient for development. The mere existence of technology is not enough to achieve positive outcomes.

ICT4D and Mobile Technologies

In recent years, development in mobile computing and communication led to the proliferation of mobile phones, tablet computers, smartphones, and netbooks. Some of these consumer electronic products, like netbooks and entry-level tablet computers are often priced lower as compared to notebooks/laptops and desktop computer since the target market for these products are those living in the emerging markets. This made the Internet and computing more accessible to people, especially in emerging markets and developing countries where most of the world's poor reside. Furthermore, these consumer electronic products are equipped with basic mobile communication hardware like, WiFi and 2.5G/3G Internet USB sticks. These allowed users to connect to the Internet via mobile and wireless networks without having to secure a landline or an expensive broadband connection via DSL, cable Internet or fiber optics.

According to International Telecommunication Union, mobile communications and technology has emerged as the primary technology that will bridge in the least developed countries. This trend can be further supported by the rosy sales reports of technology companies selling these electronic devices in emerging markets which includes some of the least developed countries. In fact, some multinational computer manufacturers like Acer and Lenovo are focusing in bringing cheaper netbooks to emerging markets like China, Indonesia and India.

Moreover, data from the ITU's Measuring the Information Society 2011 report shows that mobile phones and other mobile devices are replacing computers and laptops in accessing the Internet. Countries in Africa have also recorded growth in using mobile phones to access the Internet. In Nigeria, for example, 77% of individuals aged 16 and above use their mobile phones to access the Internet as compared to a mere 13% who use computers to go online. These developments and growth in mobile communication and its penetration in developing countries are expected to bridge the digital divide between least-developed countries and developed countries although there are still challenges in making these services affordable.

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Consumer Electronics and Electronic Waste

Sachin Vijay

Department of Electronics, Kalinga University, Raipur, CG

Consumer electronics (abbreviated CE) are electronic equipment intended for everyday use, most often in entertainment, communications and office productivity. Main products include radio receivers, television sets, MP3 players, video recorders, DVD players, digital cameras, camcorders, personal computers, video game consoles, telephones and mobile phones. Increasingly these products have become based on digital technologies, and have largely merged with the computer industry in what is increasingly referred to as the consumerization of information technology. The electronics industry, especially meaning consumer electronics emerged in the 20th century and has now become a global industry worth billions of dollars. Contemporary society uses all manner of electronic devices built in automated or semi-automated factories operated by the industry.

Trends

One overriding characteristic of consumer electronic products is the trend of ever-falling prices. This is driven by gains in manufacturing efficiency and automation, lower labor costs as manufacturing has moved to lower-wage countries, and improvements in semiconductor design. Semiconductor components benefit from Moore's Law, an observed principle which states that, for a given price, semiconductor functionality doubles every two years. While consumer electronics continues in its trend of convergence, combining elements of many products, consumers face different decisions when purchasing. There is an ever increasing need to keep product information updated and comparable, for the consumer to make an informed choice. Style, price, specification, and performance are all relevant. There is a gradual shift towards e-commerce web-storefronts.

Many products include Internet connectivity using technologies such as Wi-Fi, Bluetooth or Ethernet. Products not traditionally associated with computer use (such as TVs or Hi-Fi equipment) now provide options to connect to the Internet or to a computer using a home network to provide access to digital content. The desire for high-definition (HD) content has led the industry to develop a number of technologies, such as WirelessHD

or ITU-T G.hn, which are optimized for distribution of HD content between CE devices in a home.

Environmental Impact

Standby power used by consumer electronics and appliance while they are turned off accounts for 5–10% of total household energy consumption, adding an estimated $3 billion to annual energy costs in the United States. "In the average home, 75% of the electricity used to power home electronics is consumed while the products are turned off."

Electrical waste contains hazardous but also valuable and scarce materials and up to 60 elements can be found in complex electronics.

The processes of dismantling and disposing of electronic waste in the third world lead to a number of environmental impacts as illustrated in the graphic. Liquid and atmospheric releases end up in bodies of water, groundwater, soil, and air and therefore in land and sea animals – both domesticated and wild, in crops eaten by both animals and human, and in drinking water.

Electronic Waste

Electronic waste describes discarded electrical or electronic devices. There is a lack of consensus as to whether the term should apply to resale, reuse, and refurbishing industries, or only to a product that cannot be used for its intended purpose. Informal processing of electronic waste in developing countries may cause serious health and pollution problems, though these countries are also most likely to reuse and repair electronics. Rapid changes in technology, changes in media (tapes, software, MP3), falling prices, and planned obsolescence have resulted in a fast-growing surplus of electronic waste around the globe. Technical solutions are available, but in most cases a legal framework, a collection, logistics, and other services need to be implemented before a technical solution can be applied.

Display units (CRT, LCD, LED monitors), processors (CPU, GPU, or APU chips), memory (DRAM or SRAM), and audio components have different useful lives. Processors are most frequently out-dated (by software no longer being optimized) and are more likely to become "e-waste", while display units are most often replaced while working without repair attempts, due to changes in wealthy nation appetites for new display technology.

An estimated 50 million tons of E-waste are produced each year.

The United States is the world leader in producing electronic waste, tossing away about 3 million tons each year. China already produces about 2.3 million tons (2010 estimate) domestically, second only to the United

States. And, despite having banned e-waste imports, China remains a major e-waste dumping ground for developed countries. The UNEP estimate that the amount of e-waste being produced - including mobile phones and computers - could rise by as much as 500 percent over the next decade in some countries, such as India.

Information Security

E-waste presents a potential security threat to individuals and exporting countries. Hard drives that are not properly erased before the computer is disposed of can be reopened, exposing sensitive information. Credit card numbers, private financial data, account information, and records of online transactions can be accessed by most willing individuals. Organized criminals in Ghana commonly search the drives for information to use in local scams.

Processing Techniques

In many developed countries, electronic waste processing usually first involves dismantling the equipment into various parts (metal frames, power supplies, circuit boards, plastics), often by hand, but increasingly by automated shredding equipment. A typical example is the NADIN electronic waste processing plant in Novi Iskar, Bulgaria—the largest facility of its kind in Eastern Europe. The advantages of this process are the human's ability to recognize and save working and repairable parts, including chips, transistors, RAM, etc. The disadvantage is that the labor is cheapest in countries with the lowest health and safety standards.

Sustainable Electronics

Increasing environmental awareness has led to changes in electronics design to reduce or eliminate toxic materials and to reduce energy consumption. The landmark Restriction of Hazardous Substances Directive (RoHS) and Waste Electrical and Electronic Equipment Directive (WEEE) were released by the European Commission in 2002.

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Drug Discovery

Sayed Azeem Haider Abidi

Research Scholar, Department of Pharmecy,
Shri Venkateshwara University, Gajraula, UP

In the fields of medicine, biotechnology and pharmacology, drug discovery is the process by which new candidate medications are discovered. Historically, drugs were discovered through identifying the active ingredient from traditional remedies or by serendipitous discovery. Later chemical libraries of synthetic small molecules, natural products or extracts were screened in intact cells or whole organisms to identify substances that have a desirable therapeutic effect in a process known as classical pharmacology. Since sequencing of the human genome which allowed rapid cloning and synthesis of large quantities of purified proteins, it has become common practice to use high throughput screening of large compounds libraries against isolated biological targets which are hypothesized to be disease modifying in a process known as reverse pharmacology. Hits from these screens are then tested in cells and then in animals for efficacy.

Modern drug discovery involves the identification of screening hits, medicinal chemistry and optimization of those hits to increase the affinity, selectivity (to reduce the potential of side effects), efficacy/potency, metabolic stability (to increase the half-life), and oral bioavailability. Once a compound that fulfills all of these requirements has been identified, it will begin the process of drug development prior to clinical trials. One or more of these steps may, but not necessarily, involve computer-aided drug design. Despite advances in technology and understanding of biological systems, drug discovery is still a lengthy, "expensive, difficult, and inefficient process" with low rate of new therapeutic discovery. In 2010, the research and development cost of each new molecular entity (NME) was approximately US$1.8 billion. Drug discovery is done by pharmaceutical companies, with research assistance from universities. The "final product" of drug discovery is a patent on the potential drug. The drug requires very expensive Phase I, II and III clinical trials, and most of them fail. Small companies have a critical role, often then selling the rights to larger companies that have the resources to run the clinical trials.

Discovering drugs that may be a commercial success, or a public health success, involves a complex interaction between investors, industry, academia, patent laws, regulatory exclusivity, marketing and the need to balance secrecy with communication.

Historical Background

The idea that the effect of a drug in the human body is mediated by specific interactions of the drug molecule with biological macromolecules, (proteins or nucleic acids in most cases) led scientists to the conclusion that individual chemicals are required for the biological activity of the drug. This made for the beginning of the modern era in pharmacology, as pure chemicals, instead of crude extracts, became the standard drugs. Examples of drug compounds isolated from crude preparations are morphine, the active agent in opium, and digoxin, a heart stimulant originating from Digitalis lanata. Organic chemistry also led to the synthesis of many of the natural products isolated from biological sources.

Historically substances, whether crude extracts or purified chemicals were screened for biological activity without knowledge of the biological target. Only after an active substance was identified was an effort made to identify the target. This approach is known as classical pharmacology, forward pharmacology, or phenotypic drug discovery.

Later, small molecules were synthesized to specifically target a known physiological/pathological pathway, rather than adopt the mass screening of banks of stored compounds. This led to great success, such as the work of Gertrude Elion and George H. Hitchings on purine metabolism, the work of James Black on beta blockers and cimetidine, and the discovery of statins by Akira Endo. Another champion of the approach of developing chemical analogues of known active substances was Sir David Jack at Allen and Hanbury's, later Glaxo, who pioneered the first inhaled selective beta2-adrenergic agonist for asthma, the first inhaled steroid for asthma, ranitidine as a successor to cimetidine, and supported the development of the triptans.

Gertrude Elion, working mostly with a group of fewer than 50 people on purine analogues, contributed to the discovery of the first anti-viral; the first immunosuppressant (azathioprine) that allowed human organ transplantation; the first drug to induce remission of childhood leukaemia; pivotal anti-cancer treatments; an anti-malarial; an anti-bacterial; and a treatment for gout. Cloning of human proteins made possible the screening of large libraries of compounds against specific targets thought to be linked to specific diseases. This approach is known as reverse pharmacology and is the most frequently used approach today.

Drug Targets

Generally, the "target" is the naturally existing cellular or molecular structure involved in the pathology of interest that the drug-in-development is meant to act on. However, the distinction between a "new" and "established" target can be made without a full understanding of just

what a "target" is. This distinction is typically made by pharmaceutical companies engaged in discovery and development of therapeutics. In an estimate from 2011, 435 human genome products were identified as therapeutic drug targets of FDA-approved drugs.

"Established targets" are those for which there is a good scientific understanding, supported by a lengthy publication history, of both how the target functions in normal physiology and how it is involved in human pathology. This does not imply that the mechanism of action of drugs that are thought to act through a particular established targets is fully understood. Rather, "established" relates directly to the amount of background information available on a target, in particular functional information. The more such information is available, the less investment is (generally) required to develop a therapeutic directed against the target. The process of gathering such functional information is called "target validation" in pharmaceutical industry parlance. Established targets also include those that the pharmaceutical industry has had experience mounting drug discovery campaigns against in the past; such a history provides information on the chemical feasibility of developing a small molecular therapeutic against the target and can provide licensing opportunities and freedom-to-operate indicators with respect to small-molecule therapeutic candidates.

In general, "new targets" are all those targets that are not "established targets" but which have been or are the subject of drug discovery campaigns. These typically include newly discovered proteins, or proteins whose function has now become clear as a result of basic scientific research.

The majority of targets currently selected for drug discovery efforts are proteins. Two classes predominate: G-protein-coupled receptors (or GPCRs) and protein kinases.

Screening and Design

The process of finding a new drug against a chosen target for a particular disease usually involves high-throughput screening (HTS), wherein large libraries of chemicals are tested for their ability to modify the target. For example, if the target is a novel GPCR, compounds will be screened for their ability to inhibit or stimulate that receptor: if the target is a protein kinase, the chemicals will be tested for their ability to inhibit that kinase. Another important function of HTS is to show how selective the compounds are for the chosen target. The ideal is to find a molecule which will interfere with only the chosen target, but not other, related targets. To this end, other screening runs will be made to see whether the "hits" against the chosen target will interfere with other related targets - this is the process of cross-screening. Cross-screening is important, because the more unrelated targets a compound hits, the more likely that off-target

toxicity will occur with that compound once it reaches the clinic. It is very unlikely that a perfect drug candidate will emerge from these early screening runs. It is more often observed that several compounds are found to have some degree of activity, and if these compounds share common chemical features, one or more pharmacophores can then be developed. At this point, medicinal chemists will attempt to use structure-activity relationships (SAR) to improve certain features of the lead compound:

- increase activity against the chosen target
- reduce activity against unrelated targets
- improve the druglikeness or ADME properties of the molecule.

This process will require several iterative screening runs, during which, it is hoped, the properties of the new molecular entities will improve, and allow the favoured compounds to go forward to in vitro and in vivo testing for activity in the disease model of choice.

Amongst the physico-chemical properties associated with drug absorption include ionization (pKa), and solubility; permeability can be determined by PAMPA and Caco-2. PAMPA is attractive as an early screen due to the low consumption of drug and the low cost compared to tests such as Caco-2, gastrointestinal tract (GIT) and Blood–brain barrier (BBB) with which there is a high correlation. A range of parameters can be used to assess the quality of a compound, or a series of compounds, as proposed in the Lipinski's Rule of Five. Such parameters include calculated properties such as cLogP to estimate lipophilicity, molecular weight, polar surface area and measured properties, such as potency, in-vitro measurement of enzymatic clearance etc. Some descriptors such as ligand efficiency (LE) and lipophilic efficiency (LiPE) combine such parameters to assess druglikeness.

While HTS is a commonly used method for novel drug discovery, it is not the only method. It is often possible to start from a molecule which already has some of the desired properties. Such a molecule might be extracted from a natural product or even be a drug on the market which could be improved upon (so-called "me too" drugs). Other methods, such as virtual high throughput screening, where screening is done using computer-generated models and attempting to "dock" virtual libraries to a target, are also often used.

Another important method for drug discovery is drug design, whereby the biological and physical properties of the target are studied, and a prediction is made of the sorts of chemicals that might (e.g.) fit into an active site. One example is fragment-based lead discovery (FBLD). Novel pharmacophores can emerge very rapidly from these exercises. In general, computer-aided drug design is often but not always used to try to improve the potency and properties of new drug leads.

Once a lead compound series has been established with sufficient target potency and selectivity and favourable drug-like properties, one or two compounds will then be proposed for drug development. The best of these is generally called the lead compound, while the other will be designated as the "backup".

Nature as Source of Drugs

Traditionally many drugs and other chemicals with biological activity have been discovered by studying allelopathy - chemicals that organisms create that affect the activity of other organisms in the fight for survival.

Despite the rise of combinatorial chemistry as an integral part of lead discovery process, natural products still play a major role as starting material for drug discovery. A report was published in 2007, covering years 1981-2006 details the contribution of biologically occurring chemicals in drug development. According to this report, of the 974 small molecule new chemical entities, 63% were natural derived or semisynthetic derivatives of natural products. For certain therapy areas, such as antimicrobials, antineoplastics, antihypertensive and anti-inflammatory drugs, the numbers were higher. In many cases, these products have been used traditionally for many years.

Natural products may be useful as a source of novel chemical structures for modern techniques of development of antibacterial therapies.

Despite the implied potential, only a fraction of Earth's living species has been tested for bioactivity.

Plant-derived

Prior to Paracelsus, the vast majority of traditionally used crude drugs in Western medicine were plant-derived extracts. This has resulted in a pool of information about the potential of plant species as an important source of starting material for drug discovery. A different set of metabolites is sometimes produced in the different anatomical parts of the plant (e.g. root, leaves and flower), and botanical knowledge is crucial also for the correct identification of bioactive plant materials.

Microbial Metabolites

Microbes compete for living space and nutrients. To survive in these conditions, many microbes have developed abilities to prevent competing species from proliferating. Microbes are the main source of antimicrobial drugs. *Streptomyces* species have been a valuable source of antibiotics. The classical example of an antibiotic discovered as a defense mechanism against another microbe is the discovery of penicillin in bacterial cultures contaminated by *Penicillium* fungi in 1928.

Marine Invertebrates

Marine environments are potential sources for new bioactive agents. Arabinose nucleosides discovered from marine invertebrates in 1950s, demonstrating for the first time that sugar moieties other than ribose and deoxyribose can yield bioactive nucleoside structures. However, it was 2004 when the first marine-derived drug was approved. The cone snail toxin ziconotide, also known as Prialt, was approved by the Food and Drug Administration to treat severe neuropathic pain. Several other marine-derived agents are now in clinical trials for indications such as cancer, anti-inflammatory use and pain. One class of these agents are bryostatin-like compounds,under investigation as anti-cancer therapy.

Chemical Diversity of Natural Products

As above mentioned, combinatorial chemistry was a key technology enabling the efficient generation of large screening libraries for the needs of high-throughput screening. However, now, after two decades of combinatorial chemistry, it has been pointed out that despite the increased efficiency in chemical synthesis, no increase in lead or drug candidates has been reached. This has led to analysis of chemical characteristics of combinatorial chemistry products, compared to existing drugs or natural products. The chemoinformatics concept chemical diversity, depicted as distribution of compounds in the chemical space based on their physicochemical characteristics, is often used to describe the difference between the combinatorial chemistry libraries and natural products. The synthetic, combinatorial library compounds seem to cover only a limited and quite uniform chemical space, whereas existing drugs and particularly natural products, exhibit much greater chemical diversity, distributing more evenly to the chemical space. The most prominent differences between natural products and compounds in combinatorial chemistry libraries is the number of chiral centers (much higher in natural compounds), structure rigidity (higher in natural compounds) and number of aromatic moieties (higher in combinatorial chemistry libraries). Other chemical differences between these two groups include the nature of heteroatoms (O and N enriched in natural products, and S and halogen atoms more often present in synthetic compounds), as well as level of non-aromatic unsaturation (higher in natural products). As both structure rigidity and chirality are both well-established factors in medicinal chemistry known to enhance compounds specificity and efficacy as a drug, it has been suggested that natural products compare favourable to today's combinatorial chemistry libraries as potential lead molecules.

Natural Product Drug Discovery

Screening

Two main approaches exist for the finding of new bioactive chemical entities from natural sources.

The first is sometimes referred to as random collection and screening of material, but in fact the collection is often far from random in that biological (often botanical) knowledge is used about which families show promise, based on a number of factors, including past screening. This approach is based on the fact that only a small part of earth's biodiversity has ever been tested for pharmaceutical activity. It is also based on the fact that organisms living in a species-rich environment need to evolve defensive and competitive mechanisms to survive, mechanisms which might usefully be exploited in the development of drugs that can cure diseases affecting humans. A collection of plant, animal and microbial samples from rich ecosystems can potentially give rise to novel biological activities worth exploiting in the drug development process. One example of a successful use of this strategy is the screening for antitumour agents by the National Cancer Institute, started in the 1960s. Paclitaxel was identified from Pacific yew tree Taxus brevifolia. Paclitaxel showed anti-tumour activity by a previously undescribed mechanism (stabilization of microtubules) and is now approved for clinical use for the treatment of lung, breast and ovarian cancer, as well as for Kaposi's sarcoma. Early in the 21st century, Cabazitaxel (made by Sanofi, a French firm), another relative of taxol has been shown effective against prostate cancer, also because it works by preventing the formation of microtubules, which pull the chromosomes apart in dividing cells (such as cancer cells). Still another examples are: 1. Camptotheca (Camptothecin · Topotecan · Irinotecan · Rubitecan · Belotecan); 2. Podophyllum (Etoposide · Teniposide); 3a. Anthracyclines (Aclarubicin · Daunorubicin · Doxorubicin · Epirubicin · Idarubicin · Amrubicin · Pirarubicin · Valrubicin · Zorubicin); 3b. Anthracenediones (Mitoxantrone · Pixantrone).

Nor do all drugs developed in this manner come from plants. Professor Louise Rollins-Smith of Vanderbilt University's Medical Center, for example, has developed from the skin of frogs a compound which blocks AIDS. Professor Rollins-Smith is aware of declining amphibian populations and has said: "We need to protect these species long enough for us to understand their medicinal cabinet."

The second main approach involves Ethnobotany, the study of the general use of plants in society, and ethnopharmacology, an area inside ethnobotany, which is focused specifically on medicinal uses. Both of these two main approaches can be used in selecting starting materials for future drugs. Artemisinin, an antimalarial agent from sweet wormtree Artemisia annua, used in Chinese medicine since 200BC is one drug used as part of combination therapy for multiresistant Plasmodium falciparum.

Structural Elucidation

The elucidation of the chemical structure is critical to avoid the re-discovery of a chemical agent that is already known for its structure and

chemical activity. Mass spectrometry, often used to determine structure, is a method in which individual compounds are identified based on their mass/charge ratio, after ionization. Chemical compounds exist in nature as mixtures, so the combination of liquid chromatography and mass spectrometry (LC-MS) is often used to separate the individual chemicals. Databases of mass spectras for known compounds are available. Nuclear magnetic resonance spectroscopy is another important technique for determining chemical structures of natural products. NMR yields information about individual hydrogen and carbon atoms in the structure, allowing detailed reconstruction of the molecule's architecture.

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Noble Gas Elements in Chemistry

Dr. Sudhanshu Kumar Singh

Lecturer, Department of Chemistry,
D.B.S.D. College, Kadna, Rampur, Garkha, Saran

The noble gases make a group of chemical elements with similar properties: under standard conditions, they are all odorless, colorless, monatomic gases with very low chemical reactivity. The six noble gases that occur naturally are helium (He), neon (Ne), argon (Ar), krypton (Kr), xenon (Xe), and the radioactive radon (Rn).

For the first six periods of the periodic table, the noble gases are exactly the members of group 18 of the periodic table. It is possible that due to relativistic effects, the group 14 element flerovium exhibits some noble-gas-like properties, instead of the group 18 element ununoctium. Noble gases are typically highly unreactive except when under particular extreme conditions. The inertness of noble gases makes them very suitable in applications where reactions are not wanted. For example: argon is used in lightbulbs to prevent the hot tungsten filament from oxidizing; also, helium is breathed by deep-sea divers to prevent oxygen and nitrogen toxicity.

The properties of the noble gases can be well explained by modern theories of atomic structure: their outer shell of valence electrons is considered to be "full", giving them little tendency to participate in chemical reactions, and it has been possible to prepare only a few hundred noble gas compounds. The melting and boiling points for a given noble gas are close together, differing by less than 10 °C (18 °F); that is, they are liquids over only a small temperature range.

Neon, argon, krypton, and xenon are obtained from air in an air separation unit using the methods of liquefaction of gases and fractional distillation. Helium is sourced from natural gas fields which have high concentrations of helium in the natural gas, using cryogenic gas separation techniques, and radon is usually isolated from the radioactive decay of dissolved radium compounds. Noble gases have several important applications in industries such as lighting, welding, and space exploration. A helium-oxygen breathing gas is often used by deep-sea divers at depths of seawater over 55 m (180 ft) to keep the diver from experiencing oxygen toxemia, the lethal effect of high-pressure oxygen, and nitrogen narcosis, the distracting narcotic effect of the nitrogen in air beyond this partial-

pressure threshold. After the risks caused by the flammability of hydrogen became apparent, it was replaced with helium in blimps and balloons.

Physical and Atomic Properties

The noble gases have weak interatomic force, and consequently have very low melting and boiling points. They are all monatomic gases under standard conditions, including the elements with larger atomic masses than many normally solid elements. Helium has several unique qualities when compared with other elements: its boiling and melting points are lower than those of any other known substance; it is the only element known to exhibit superfluidity; it is the only element that cannot be solidified by cooling under standard conditions—a pressure of 25 standard atmospheres (2,500 kPa; 370 psi) must be applied at a temperature of 0.95 K (“272.200 °C; “457.960 °F) to convert it to a solid. The noble gases up to xenon have multiple stable isotopes. Radon has no stable isotopes; its longest-lived isotope, ^{222}Rn, has a half-life of 3.8 days and decays to form helium and polonium, which ultimately decays to lead.

The noble gas atoms, like atoms in most groups, increase steadily in atomic radius from one period to the next due to the increasing number of electrons. The size of the atom is related to several properties. For example, the ionization potential decreases with an increasing radius because the valence electrons in the larger noble gases are farther away from the nucleus and are therefore not held as tightly together by the atom. Noble gases have the largest ionization potential among the elements of each period, which reflects the stability of their electron configuration and is related to their relative lack of chemical reactivity. Some of the heavier noble gases, however, have ionization potentials small enough to be comparable to those of other elements and molecules. It was the insight that xenon has an ionization potential similar to that of the oxygen molecule that led Bartlett to attempt oxidizing xenon using platinum hexafluoride, an oxidizing agent known to be strong enough to react with oxygen. Noble gases cannot accept an electron to form stable anions; that is, they have a negative electron affinity.

The macroscopic physical properties of the noble gases are dominated by the weak van der Waals forces between the atoms. The attractive force increases with the size of the atom as a result of the increase in polarizability and the decrease in ionization potential. This results in systematic group trends: as one goes down group 18, the atomic radius, and with it the interatomic forces, increases, resulting in an increasing melting point, boiling point, enthalpy of vaporization, and solubility. The increase in density is due to the increase in atomic mass.

The noble gases are nearly ideal gases under standard conditions, but their deviations from the ideal gas law provided important clues for the

study of intermolecular interactions. The Lennard-Jones potential, often used to model intermolecular interactions, was deduced in 1924 by John Lennard-Jones from experimental data on argon before the development of quantum mechanics provided the tools for understanding intermolecular forces from first principles. The theoretical analysis of these interactions became tractable because the noble gases are monatomic and the atoms spherical, which means that the interaction between the atoms is independent of direction, or isotropic.

Chemical Properties

The noble gases are colorless, odorless, tasteless, and nonflammable under standard conditions. They were once labeled *group 0* in the periodic table because it was believed they had a valence of zero, meaning their atoms cannot combine with those of other elements to form compounds. However, it was later discovered some do indeed form compounds, causing this label to fall into disuse. The noble gases have full valence electron shells. Valence electrons are the outermost electrons of an atom and are normally the only electrons that participate in chemical bonding. Atoms with full valence electron shells are extremely stable and therefore do not tend to form chemical bonds and have little tendency to gain or lose electrons. However, heavier noble gases such as radon are held less firmly together by electromagnetic force than lighter noble gases such as helium, making it easier to remove outer electrons from heavy noble gases.

Noble Gas Notation

As a result of a full shell, the noble gases can be used in conjunction with the electron configuration notation to form the *noble gas notation*. To do this, the nearest noble gas that precedes the element in question is written first, and then the electron configuration is continued from that point forward. For example, the electron notation of carbon is $1s^2 2s^2 2p^2$, and the noble gas notation is $[He]2s^2 2p^2$. This notation makes it easier to identify elements, and is shorter than writing out the full notation of atomic orbitals.

Compounds

The noble gases show extremely low chemical reactivity; consequently, only a few hundred noble gas compounds have been formed. Neutral compounds in which helium and neon are involved in chemical bonds have not been formed (although there is some theoretical evidence for a few helium compounds), while xenon, krypton, and argon have shown only minor reactivity.

Occurrence and Production

The abundances of the noble gases in the universe decrease as their atomic numbers increase. Helium is the most common element in the

universe after hydrogen, with a mass fraction of about 24%. Most of the helium in the universe was formed during Big Bang nucleosynthesis, but the amount of helium is steadily increasing due to the fusion of hydrogen in stellar nucleosynthesis (and, to a very slight degree, the alpha decay of heavy elements). Abundances on Earth follow different trends; for example, helium is only the third most abundant noble gas in the atmosphere. The reason is that there is no primordial helium in the atmosphere; due to the small mass of the atom, helium cannot be retained by the Earth's gravitational field. Helium on Earth comes from the alpha decay of heavy elements such as uranium and thorium found in the Earth's crust, and tends to accumulate in natural gas deposits. The abundance of argon, on the other hand, is increased as a result of the beta decay of potassium-40, also found in the Earth's crust, to form argon-40, which is the most abundant isotope of argon on Earth despite being relatively rare in the Solar System. This process is the base for the potassium-argon dating method. Xenon has an unexpectedly low abundance in the atmosphere, in what has been called the *missing xenon problem*; one theory is that the missing xenon may be trapped in minerals inside the Earth's crust. After the discovery of xenon dioxide, a research showed that Xe can substitute for Si in the quartz. Radon is formed in the lithosphere as from the alpha decay of radium. It can seep into buildings through cracks in their foundation and accumulate in areas that are not well ventilated. Due to its high radioactivity, radon presents a significant health hazard; it is implicated in an estimated 21,000 lung cancer deaths per year in the United States alone.[

For large-scale use, helium is extracted by fractional distillation from natural gas, which can contain up to 7% helium.

Neon, argon, krypton, and xenon are obtained from air using the methods of liquefaction of gases, to convert elements to a liquid state, and fractional distillation, to separate mixtures into component parts. Helium is typically produced by separating it from natural gas, and radon is isolated from the radioactive decay of radium compounds.[

Applications

Noble gases have very low boiling and melting points, which makes them useful as cryogenic refrigerants. In particular, liquid helium, which boils at 4.2 K ("268.95 °C; "452.11 °F), is used for superconducting magnets, such as those needed in nuclear magnetic resonance imaging and nuclear magnetic resonance. Liquid neon, although it does not reach temperatures as low as liquid helium, also finds use in cryogenics because it has over 40 times more refrigerating capacity than liquid helium and over three times more than liquid hydrogen.

In many applications, the noble gases are used to provide an inert atmosphere. Argon is used in the synthesis of air-sensitive compounds that are sensitive to nitrogen. Solid argon is also used for the study of very unstable compounds, such as reactive intermediates, by trapping them in an inert matrix at very low temperatures. Helium is used as the carrier medium in gas chromatography, as a filler gas for thermometers, and in devices for measuring radiation, such as the Geiger counter and the bubble chamber. Helium and argon are both commonly used to shield welding arcs and the surrounding base metal from the atmosphere during welding and cutting, as well as in other metallurgical processes and in the production of silicon for the semiconductor industry.[

Noble gases are commonly used in lighting because of their lack of chemical reactivity. Argon, mixed with nitrogen, is used as a filler gas for incandescent light bulbs.

The noble gases are used in excimer lasers, which are based on short-lived electronically excited molecules known as excimers. The excimers used for lasers may be noble gas dimers such as Ar_2, Kr_2 or Xe_2, or more commonly, the noble gas is combined with a halogen in excimers such as ArF, KrF, XeF, or XeCl. These lasers produce ultraviolet light which, due to its short wavelength (193 nm for ArF and 248 nm for KrF), allows for high-precision imaging. Excimer lasers have many industrial, medical, and scientific applications. They are used for microlithography and microfabrication, which are essential for integrated circuit manufacture, and for laser surgery, including laser angioplasty and eye surgery.

Some noble gases have direct application in medicine. Helium is sometimes used to improve the ease of breathing of asthma sufferers. Xenon is used as an anesthetic because of its high solubility in lipids, which makes it more potent than the usual nitrous oxide, and because it is readily eliminated from the body, resulting in faster recovery.[

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Infectious Diseases of the Urinary System in Small Animals

Habibul Rahman

Research Scholar, Monad University, Hapur

Pyelonephritis

Kidney infection (pyelonephritis) is usually due to ascending bacteria, although hematogenous spread is possible. The organisms and predisposing causes are similar to those of bacterial cystitis. Renoliths and ureteroliths, which impede the normal flow of urine out of the renal pelvis, are a common cause. In young dogs or cats, congenital malformations (eg, ectopic ureters) are a predisposing cause. Animals at risk for pyelonephritis are the very young, the very old, the immuno-suppressed, or those with inadequate urine-concentrating ability. In many instances, an underlying cause is not identified.

Animals with acute pyelonephritis exhibit kidney or flank pain, fever, malaise, and sometimes vomiting, polyuria, and polydipsia. Urinalysis shows proteinuria, pyuria, bacteriuria, and/or hematuria. WBC casts are often present in fresh urine sediment. The urine culture is usually positive; the CBC may show leukocytosis with a left shift. The biochemical profile may be normal or show azotemia (prerenal or renal) and/or hyperglobulinemia. The animal may be in renal failure. Chronic pyelonephritis is more difficult to recognise because clinical signs may be subtle or absent. Polyuria and polydipsia are frequent. In many cases, the disease goes unrecognised until renal failure occurs. Although abnormalities in the urinalysis are present, they are often less dramatic than with acute kidney infection. A single urine culture can be negative if bacterial numbers are low. Other useful diagnostic tests include renal ultrasonography and IV pyelograms. Both studies may show dilation of one or both renal pelvises secondary to inflammation and partial obstruction. Asymmetric renal size and architectural changes with renal pelvic dilatation is highly suggestive of chronic pyelonephritis. In some cases, nephropyelocentesis via ultrasonographic guidance is useful to obtain a sample of urine from the dilated renal pelvis for analysis and culture.

Pyelonephritis should be treated aggressively with broad-spectrum antibiotics, based on urine culture and antimicrobial susceptibility testing, for 4-6 wk. The infection may respond to the same antibiotics

recommended for cystitis, but more frequent administration (eg, amoxicillin tid rather than bid) and/or higher dosages are indicated.

A fluoroquinolone or a combination of a fluoroquinolone with a α-lactam antibiotic is often effective. Dosages should be the same as for other soft-tissue infections. Animals that are febrile, anorectic, dehydrated, or azotemic should be hospitalized for IV antibiotics and fluid therapy.

Fluid therapy may prevent acute pyelonephritis from progressing to acute renal failure and will improve renal perfusion and uremic signs in animals already in renal failure. Animals with acute pyelonephritis may recover normal renal function, depending on the amount of damage that occurred prior to treatment. In selected cases of chronic pyelonephritis with a severely hydronephrotic, nonfunctional kidney, a nephrectomy may be the treatment of choice once the animal has been stabilized. This will remove the source of infection and hopefully save the opposite kidney. IV pyelography and/or renal scintigraphy are useful to assess the relative function of each kidney. If both kidneys are severely affected, medical management alone is the only alternative. Recovery to chronic, stable renal failure is possible in many cases.

The urine should be cultured after the first 5-7 days of therapy to assess antibiotic efficacy. A urinalysis and culture should be repeated 3-5 days following therapy, and then monthly for 3 consecutive months. If all of these cultures are negative, the interval between urine cultures may be gradually lengthened. Animals with pyelonephritis are at high risk for persistence or recurrence of kidney infection.

Interstitial Nephritis

Acute interstitial nephritis in dogs is caused most often by Leptospira *interrogans* (*Leptospirosis in Dogs*). Cats can develop leptospirosis, although signs are less severe than in dogs. Other infectious causes of interstitial nephritis in dogs include Leishmania *donovani* and Borrelia *burgdorferi*. Glomerulone-phritis is the predominant renal pathology, so these diseases are considered immune complex diseases rather than true infections. Infectious diseases that cause vasculitis in dogs or cats (eg, feline infectious peritonitis, Rocky Mountain spotted fever, ehrlichiosis) may also cause renal failure.

Capillaria Plica *Infection*

Capillaria plica may infect the urinary bladder, and occasionally the ureters and renal pelvises, of dogs and cats. Distribution is worldwide, and wild animals appear to be the primary hosts. A similar but less common organism, C *felis cati*, is also found in cats.

Dogs and cats become infected with C *plica* by eating earthworms that contain the first-stage larvae. The worms are threadlike, yellowish, and 13-60 mm long. The eggs are colourless, operculated, have a slightly pitted shell, and are 63-68 × 24-27 μm in size.

Most dogs and cats are asymptomatic. Some animals show signs of pollakiuria, urinary incontinence, and urinating in abnormal places. The eggs are shed in the urine and may be found in the urine sediment. Microscopic hematuria and increased numbers of epithelial cells may also be present. Reported treatments include levamisole, fenbendazole, albendazole, and ivermectin. The treatment of choice is unknown, but a single dose of ivermectin at 0.2 mg/kg, SC, is likely to be effective. It is not FDA approved for this use and is contraindicated in Collie breeds. The parasite may be self-limiting in the absence of reinfection.

Giant Kidney Worm Infection in Mink and Dogs

Mink are the most common definitive host for Dioctophyma renale, the largest known nematode, which has a worldwide distribution. Many other species, including dogs and humans, can become infected. The definitive host contracts the parasite by ingesting encysted larvae in raw fish (eg, pike, bullhead) or frogs, or by ingesting an infected annelid worm. The larvae penetrate the bowel wall and migrate first to the liver and later to the kidneys. In dogs, the parasite often fails to reach the kidneys and may be found free in the abdominal cavity. Kidney worms grow larger in dogs than in mink, reaching up to 103 cm. Female worms are larger than male worms, and both are blood red. Both male and female worms must be present in the same kidney to complete the life cycle. Barrel-shaped, yellow-brown eggs with a thick pitted shell measuring 71-84 × 45-52 μm are shed into the urine.

In the kidneys, the worm(s) cause obstruction, hydrone-phrosis, and destruction of the renal parenchyma. The right kidney is most commonly affected in both mink and dogs. Kidney failure can result if both kidneys are parasitized. Chronic peritonitis, adhesions, and liver disease are also possible, especially in dogs.

Clinical signs are hematuria, pollakiuria, weight loss, and renal or abdominal pain. Urinalysis may reveal proteinuria, hematuria, and pyuria. IV pyelography or ultrasonography shows the enlarged hydronephrotic kidney.

The diagnosis is made by finding the eggs in the urine sediment if both sexes of the nematode are present in the kidney and the ureter is patent. Alternatively, exploratory laparotomy may reveal the diagnosis.

Worms may be found in the peritoneal cavity, between the lobes of the liver, or within the affected kidney(s) via nephrotomy. Unilateral nephrectomy is the treatment of choice if the opposite kidney is unaffected. Preventing ingestion of raw fish or other aquatic organisms is recommended, especially in areas where the parasite is known to infect wild animals.

Bacterial Cystitis

Bacterial cystitis is infection and inflammation of the urinary bladder. Clinical signs are pollakiuria, hematuria, dysuria, and urinating in inappropriate places. Hematuria may be more noticeable at the end of the urine stream. An animal may exhibit pain on palpation of the caudal abdomen, and the bladder may feel thickened or irregular. Bacterial cystitis is occasionally diagnosed in an asymptomatic animal when a routine urinalysis is performed. Chronic glucocorticoid administration or hyperadrenocorticism are sometimes associated with asymptomatic urinary tract infections.

Urinalysis often shows increased protein and hemoglobin on the dipstick. The WBC part of the dipstick (ie, nitrate) is inaccurate in dogs and cats and should not be used. The urine pH may be alkaline (7.5-9.0) if the bacteria are urease positive (eg, Staphylococcus or Proteus). An alkaline urine pH by itself is not abnormal, however, as diet and other factors can affect urine pH. Urine sediment should be examined microscopically. Increased numbers of WBC, RBC, and/or bacteria are consistent with cystitis. Bacteria can be confused with stain precipitate; filtering the stain or evaluating the sediment without staining is advised. Lack of visible bacteria in the sediment does not rule out urinary tract infection.

If clinical signs and/or urinalysis are suggestive of infection, a urine culture and antimicrobial susceptibility should be performed. Cystocentesis is the preferred method for sample collection, followed by sterile urethral catheterization or a midstream free catch into a sterile collection cup. A quantitative culture is necessary to interpret the result, especially with samples not collected by cystocentesis. Ideally, the culture should be set up within 2 hr of collection. If the laboratory is off-site, the sample should be refrigerated and processed by the laboratory within 24 hr. If the specimen cannot be refrigerated, commercial collection kits that contain preservatives can be used to maintain a stable bacterial population at room temperature for 24 hr. Laboratories that can provide both quantitative culturing and a minimum inhibitory concentration-based method for antimicrobial susceptibility testing are preferred.

Simple bacterial cystitis is treated for 2 wk with a broad-spectrum antibiotic that achieves a high concentration in the urine. Appropriate initial choices include amoxicillin (10-20 mg/kg, PO, bid-tid), cefadroxil (22-30 mg/kg, PO, bid), or ometoprim-sulfadimethoxine (27 mg/kg, PO, day one, then 13.5 mg/kg, PO, sid). A repeat urine culture 3-5 days following therapy is recommended. If positive, another antibiotic based on the new susceptibility results is given for a longer treatment period (eg, 3-4 wk). Very resistant or recurrent infections should be treated for 4-6 wk. Every course of treatment should be followed by a urine culture, even if the signs have resolved. In animals that have a history of chronic or recurrent infections, a urine culture should be done every month for 3 mo following therapy. If all of these cultures are negative, then a urine culture every 2-4 mo for the next year is advisable. Because resistance to antibiotics can develop during therapy, antimicrobial susceptibility testing should be performed on every positive urine culture.

Animals with resistant or recurrent bacterial cystitis should be evaluated for an underlying cause. The history may reveal chronic glucocorticoid use. Survey abdominal radiographs are frequently diagnostic for cystic calculi. Negative survey films should be followed by double contrast cystourethrography, ultrasonography, and/or cystoscopy to rule out radiolucent urocystoliths, anatomic defects, and neoplasia. A serum biochemical profile and CBC are important to rule out predisposing systemic diseases. Other diagnostic considerations include feline immunodeficiency virus, feline leukemia virus, and hyperthyroidism in cats and hyperadrenocorticism in dogs.

In cases that respond to therapy but continue to have frequent bouts of cystitis without an identifiable cause, low-dose prophylactic antibiotics can be used to prevent ascending bacteria from establishing an infection according to the following protocol: 1) a therapeutic course of an antibiotic for the current infection is completed, 2) no antibiotics are given for 3 days, to allow collection of urine for a post-treatment culture, and 3) the prophylactic protocol is immediately started. Prophylaxis consists of using a broad-spectrum antibiotic (eg, amoxicillin, cefadroxil) at 1/3 of the total daily dose, given at bedtime, indefinitely. Every 6-8 wk, the antibiotic should be stopped for 3-5 days to obtain a sample for repeat urinalysis and culture. Every new infection should be treated with a therapeutic course of an antibiotic based on culture and susceptibility results. The treatment antibiotic will likely be different than the prophylactic antibiotic. The most valuable therapeutic antibiotics (eg, fluoroquinolones, second-generation cephalosporins) should be reserved for resistant infections. If

the recurrent infection is resistant to the prophylactic antibiotic, this antibiotic can still be used for future prophylaxis after the infection is eradicated. Encouraging frequent voiding during the daytime is helpful in preventing recurrent infections.

Noninfectious Diseases of the Urinary System in Small Animals

Renal Dysfunction

Failure of the filtration function of the kidneys leads to the development of azotemia (an excess of nitrogenous compounds in the blood), which may be classified as prerenal, renal, postrenal, or of mixed origin. Prerenal azotemia develops whenever mean systemic arterial blood pressure declines to values <60 mm Hg and/or when dehydration causes plasma protein concentration to increase. Conditions that may lead to the development of prerenal azotemia include dehydration, congestive heart failure, and shock. Prerenal azotemia generally resolves with appropriate treatment, because kidney structure has not been altered, which allows normal function to resume once renal perfusion has been restored. Renal azotemia refers to a reduction in glomerular filtration rate (GFR) of ~75% during acute or chronic primary renal (or intrarenal) diseases. Postrenal azotemia develops when the integrity of the urinary tract is disrupted (eg, bladder rupture) or urine outflow is obstructed (eg, urethral or bilateral ureteral obstruction). Once adequate urine flow is restored, postrenal azotemia will resolve.

Chronic Kidney Disease

This disease process involves a loss of functional renal tissue due to a prolonged (≥3 mo), usually progressive process. Dramatic changes in renal structure may be seen, although structural and functional changes in the kidney are only loosely correlated. Chronic kidney disease often smolders for many months or years before it becomes clinically apparent, and is invariably irreversible and progressive. Although congenital disease results in a transient increase in prevalence in animals <3 yr old, the prevalence increases with advancing age from 5-6 yr. In geriatric populations at referral institutions, chronic kidney disease affects up to 10% of dogs and 35% of cats. The prevalence in the general small animal population is likely to be lower. Several breeds of dogs and cats are afflicted with heritable chronic kidney disease. There is no apparent breed or sex predisposition for nonheritable chronic kidney disease in dogs or cats.

Chronic kidney disease is generally classified into various stages (*Table: International Renal Interest Society Classification of Stages of Kidney Disease*) based on laboratory tests and clinical signs. In Stage I, a process is

damaging the kidneys but azotemia and clinical signs have not developed. Unfortunately, renal disease is uncommonly detected at this stage. In Stage II, the disease has progressed, GFR has fallen to <25% of normal, and azotemia is present, but clinical signs are not yet observed. However, this stage may be associated with impaired urine-concentrating ability and increased urine volume. Stage III occurs when GFR has declined further and both azotemia and clinical signs are often present. Stage IV reflects further progression and severe azotemia, with clinical signs present.

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