TECHNICAL DEVELOPMENTS
JULY 1969 - JUNE 1970

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B. Wells

June 26, 1972

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FOREWORD

"As a public agency, the AEC has a responsibility to effect the prompt transfer to civilian industry, non-nuclear as well as nuclear, of all new information and technology developed as an integral part of AEC-supported programs," (AEC Manual, Chapter 3206).

To aid the AEC in fulfilling its responsibility for technology transfer, we have compiled this selection of unclassified LLL technical developments and concepts reported to the Commission during the period July 1969 through June 1970. From the total number of reported cases we have selected those which we believe will be of maximum interest to readers of this review. As in the past, our main purpose in compiling this report is to present a broad spectrum of novel solutions to difficult technical problems. Seeing how "the other guy" solved his problem is often an effective spur to solving one's own.

This report represents the work of many people. We thank our past and present Patent Engineers, Tom Schneck, Dave Newhouse, and Geert Drayer, for the comprehensive technical descriptions which are the basis of this summary. Our capable secretary, Beth Brewster, accurately translated cryptic signs and scribbles into clean typed copy. The drawings, which contribute immeasurably to a report of this type, were done by John Cary. We also thank the staff of LLL's Technical Information Department, especially Dick Crawford, for their expert help in the production of this report.

But our greatest debt is owed to those creative people for whom problem-solving is a way of life. They are the fountainhead of all progress in "science and the useful arts." It is to them, the scientists, engineers, and technicians, that we appreciatively dedicate this report.

Stanley S. Lerner
Berthold J. Weis

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Section 1. Charged Particle Beam Devices

Case No.: **IL-5209**

Title: **BEAM LENS APPARATUS FOR ENERGY CONVERSION**

Inventor: **Richard F. Post**

This apparatus is a device for converting the kinetic energy carried by the particles in a beam, such as the beam emerging from a controlled thermonuclear reaction device, into usable electrical energy. As shown in Fig. 1.1, a vacuum chamber contains a series of electrodes 11, 12, 13, 14 and 15, through which a beam 16 passes. To each of the electrodes, an increasing potential $V_1, V_2, V_3, V_4, \dots, V_n$, respectively, is applied. Usually the potentials will be on the order of tens of kilovolts. For example, V_1 might be equal to 10 kV, V_2 equal to 20 kV, V_3 equal to 30 kV, etc. A beam comprising mainly positive ions, such as ions emerging from a controlled thermonuclear reaction apparatus, tends to be decelerated by passing through such electrodes.

The electrodes are shaped so as to have a focusing effect on the beam. The convergence (reciprocal of the focal length of the beam lens) for a system of lenses will continue to focus an incoming beam along the beam axis as long as the beam particles have a sufficient energy. The mathematical condition for focusing of a beam system is represented by the equation:

$$CL < 4$$

where C is the convergence of an electrode, and L is the spacing between centers of adjacent electrodes. It is known that the value of the convergence per lens increases with decreasing particle energy, i. e., the beam focal length approaches zero with decreasing particle kinetic energy.

It will be noted that the electrode lengths L and the convergence for the lens system C can be adjusted so that particle paths diverge from the beam axis at the point where most of the kinetic energy has been extracted from the particle. This is illustrated by the particle path labeled 17, which shows a particle being focused while traversing electrodes 11 and 12. As the particle encounters a potential V_3 in electrode 13, it finally experiences defocusing, since the particle's kinetic energy has been decreased by the previous potentials V_1 and V_2 . The defocusing causes the particle to diverge sufficiently to land on electrode 13 and be carried away in the form of an electric current. By placing enough electrodes along the beam path, virtually all the beam particles can be trapped and extracted from the apparatus in the form of current at a discrete potential.

Fig. 1.1. Energy conversion beam lens.

Case No.: **IL-5276**

Title: **ENERGY CONVERSION APPARATUS**

Inventor: **Ralph W. Moir**

One of the interesting features of fusion reactors is that their output power may be converted directly into usable electrical energy without intervening heat cycles. The present device has been demonstrated as an effective means for the direct conversion of the mechanical energy of charged particles into electrical energy.

The apparatus will be described with reference to Fig. 1.2. A charged particle stream 11, parallel to an ambient magnetic field, B , is directed toward a first electrode 13 having a potential higher than the comparable kinetic energy of any particle in the stream so that all particles in the stream will be reflected from electrode 13.

Upon reflection, the particle is directed toward a second electrode 14 which is charged to a potential slightly less than the first electrode. If the potential of the second electrode is greater than that of the particle, the particle will again be reflected in the opposite direction. This process continues until the charged particle encounters an electrode which is charged to a potential slightly less than the kinetic energy of the particle, so that the particle is caught on that electrode. Hence, the charged particle is collected at the first electrode whose potential is less than the kinetic energy of the charged particle.

Once a particle is collected on a particular electrode, the electrical potential energy stored in the particle may be carried off by electrical transmission lines. The electrical power may either be used directly or converted to a common potential by an inverter-rectifier system.

The origin of the charged particle beam 11 is not important, although one such source having an axial magnetic field parallel to the beam is a thermonuclear reactor of the ALICE type. Since the magnetic field permeates a housing 12, and since an electrical field, E , is provided perpendicular to the magnetic field, B , an $E \times B$ drift is imposed on the charged particle trajectory, bringing the particles into contact with an appropriate electrode for collection.

An alternate embodiment of the apparatus is shown in Fig. 1.3. This is essentially the same as Fig. 1.2, except that the first electrode 13 reflects each particle only once, and each successive reflecting electrode has a lower potential.

The width of each electrode is designed to collect particles of a given energy. Since the particle trajectory is known and assumes the general shape of a triangle between adjacent electrodes, the width of an electrode is made slightly larger than the base of an adjoining triangular trajectory portion.

In principle, this recovery system can be 100% efficient, but due to space charge limitations, will probably be limited to approximately 30%.

Fig. 1.2. Reflective energy conversion electrode system, single reflector.

Fig. 1.3. Reflective energy conversion electrode system, multiple reflectors.

Case No.: **IL-5225**

Title: **ION SOURCE**

Inventors: **Arthur C. Bolson, Charles M. Henderson, and Paul G. Johnson**

An "isotope separator" is essentially a mass spectrometer designed for maximum recovery of separated isotopes. The isotopic sample is vaporized and ionized in an "ion source" and then introduced into the separating section of the spectrometer. Conventional ion sources consist of: (1) a furnace for vaporizing the sample; (2) a means for ionizing the vapor, such as an electron beam or an arc discharge; and (3) a system for extracting and focusing the resulting ions into a beam. In conventional sources, connections to electrical elements inside the furnace are made with insulating feed-throughs. The resulting "cold spots" on the furnace wall condense much of the sample vapor, limiting its contribution to the ion beam. The isotope yield of separators using such sources is therefore quite low, usually less than 10%. Some ion sources use a stream of carbon tetrachloride to improve sample vaporization by converting the metal into its volatile chloride. Although this technique does increase isotope yield, the chlorine gas produced in the metal/tetrachloride reaction tends to corrode the apparatus, and the increase in pressure is deleterious.

This apparatus is an improved ion source which eliminates the need for a separate vaporizing furnace, is free of cold spots which condense the sample vapor, and has virtually no paths for vapor loss except for the extraction hole. Also, optimum operation does not require close pressure control as in a conventional arc-discharge type source. Isotope yield for rare-earth elements is in the 10-60% range, a substantial improvement over conventional ion sources.

As shown in Fig. 1.4, the sample is placed in a cavity inside a cylindrical tungsten crucible. A cap seals the opening except for a small orifice in the extraction end. The crucible is heated by electron bombardment from a ring filament biased at several kilovolts negative with respect to the crucible. Because the extraction end of the crucible is closest to the filament, a temperature gradient of several hundred degrees Centigrade is produced, the crucible being hottest at the extraction end and coolest at the sample-containment portion. This temperature gradient allows a controlled volatilization of the sample while maintaining a temperature high enough for "surface ionization" of the sample vapor near the crucible orifice. Since the only opening is the orifice, there are no cold spots in the crucible to condense sample vapor. The use of electron bombardment, rather than resistance heating, to heat the crucible enables the sample to be heated to a temperature sufficient for surface ionization of the vapor, and also for complete vaporization of the sample without resorting to reaction with carbon tetrachloride.

The crucible and ring filament are surrounded by a shield cup held at the filament potential by a direct shield/filament weld. The electrons emitted by the filament are thereby directed onto the crucible rather than in random directions. In addition to heating the crucible, the bombardment electrons also contribute to ionization of the sample

vapor by colliding with vapor molecules as they leave the orifice. The major ionization mechanism is "surface ionization" resulting from contact of vapor molecules with the hot crucible surface. Still another mechanism is ionization by mutual collision between fast-moving vapor molecules, that is, "thermal ionization". After emerging from the crucible orifice, sample ions enter the separator through a small hole in the shield cup.

Typical operating conditions and apparatus dimensions are as follows:

Isotope yield	10-60% (for rare-earth elements)
Bombardment voltage	1-3 kilovolts
Beam current	10 microamps per isotope
Sample weight	1 milligram
Orifice diameter	1 millimeter
Filament-to-orifice distance	5 millimeters

Fig. 1.4. Electron bombardment ion source.

Case No.: **IL-5266**

Title: **ELECTRON LENS SYSTEM**

Inventor: **Ralph Kalibjian**

This apparatus is a lens system for transforming an electron beam having an initially circular cross section into a diverging sheet beam, i.e., a "fan-shaped" beam. Such a fan-shaped beam is useful for various scanning and readout devices.

Referring to Fig. 1.5, an electron gun 2 emits an electron beam 4 with circular cross section along the axis 6 toward a lens system 8. The lens system comprises a pair of trapezoidally-shaped metal plates 10 and 12 placed parallel to one another about the axis and equidistant therefrom, and a pair of triangularly-shaped metal U-sections 14 and 16 placed adjacent to, but not in contact with, plates 10 and 12 to form a rectangular duct or channel centered about the axis. Means (not shown) are provided for keeping the trapezoidal plates at a negative potential V_1 and the U-sections at a positive potential V_2 .

In operation, electrons in the circular electron beam, entering the lens system along the axis, experience a net force directed perpendicular to the axis and parallel to the trapezoidal plates due to the electric fields produced by the potentials V_1 and V_2 . The circular electron beam 4 is thus transformed into a diverging sheet beam 18 with an elongated rectangular cross section. Control of the transformation in the cross sectional shape of the circular electron beam 4 is achieved by varying the magnitudes of V_1 and V_2 .

In a typical embodiment of the present invention, the lens system dimensions were:

$$L = 2.6"; \quad W = 2"; \quad H = 1";$$

$$W_1 \approx W_2 \approx W_3 \sim 2/3"$$

Note that the lens system operates equally well for generating an output sheet beam in spatial quadrature to an input sheet beam by reversing the polarities of the potentials V_1 and V_2 .

Fig. 1.6 illustrates an exemplary use of the present lens system as a means for achieving bandwidth reduction in the scanning of a target by an electron beam. Filament 20, cathode 22, grid 24, accelerating electrodes 26, and deflecting plates 28 are standard elements in a conventional circular beam CRT device. The lens system 8 of Fig. 1.5 is placed between the accelerating electrodes 26 and the deflecting plates 28. A target 30 is positioned to receive the generated electron beam, and includes a matrix of elements 32 to be scanned. In operation with a circular electron beam, and without the lens system, the matrix elements would be scanned one at a time in the conventional "left-to-right and top-to-bottom" manner. In operation with the lens system, all matrix elements in the column are scanned simultaneously by deflecting the transformed

electron beam once across all rows from left to right. For equal horizontal beam deflection rates this system reduces the bandwidth in proportion to the increase in the number of channels required.

Fig. 1.5. Electron lens for forming a sheet beam.

Fig. 1.6. Scanning application of electron sheet beam.

Case No.: IL-5273

Title: X-RAY GENERATOR

Inventors: Jerry L. Gaines, Paul J. Ebert, Lloyd Mancebo, and Frederick D. Cook

In many research applications, sources of high intensity monoenergetic x-rays in the energy interval from 1 to 8 keV are required. There are very few radioactive sources with intense monoenergetic x-ray emission in this region. The present x-ray generator incorporates the following features: (a) the x-ray output is variable in intensity; (b) the x-ray beam is variable in discrete energy increments in the energy interval of 1.5 to 8 keV; (c) the effective source size is well-defined; and (d) the source strength as well as spectral purity can be accurately determined.

Figure 1.7 is a cutaway view of the present device. A hollow cylindrical anode 2 and a hollow cylindrical electrode 4 are disposed on opposite sides of a circular aperture 6, defined by an annular electrode 8, and about a common axis 10 through the center of the aperture. Electrode 4 is coaxially surrounded by a cylindrical grid 12 and a cylindrical electrode 14. The annular electrode 8 is supported by a conductive cylindrical structure 16 which is coaxially spaced from electrode 14 by means of conductive spacers 18. The cylindrical grid and electrode 12 and 14 respectively terminate in flared sections 20 and 22 with conforming curvatures which partially surround a ring-like cathode 24 coaxially disposed between the structures 14 and 16. The peripheral end section of the electrode structure 4 facing the aperture 6 is smoothly curved outwardly to form a rounded lip 26. The end section 28 of the anode includes a cone-shaped indentation 30 symmetrical about the axis. A thin layer of target material, such as aluminum or copper, is deposited on the walls of the indentation. Coolant circulates through the anode structure (inlet and outlet ports not shown). The surrounding evacuated envelope 32 includes suitable electrical lead-ins, supports for the various structures (not shown), and an x-ray transparent window 34 disposed about the axis near the lower portion of hollow electrode 4. Electrodes 4, 14, and 18 are connected to a reference ground potential, anode structure 2 is connected to a high voltage source (such as +30 kV), and the grid 12 is connected to a voltage source producing predetermined positive and negative voltage levels (such as +500 V, and -200 V).

With grid 12 at a negative potential, electrons emitted by the cathode fail to pass aperture 6 due to the opposing electrostatic field produced by the grounded electrodes 4, 14, and 18, and the negative potential on the grid. With the grid at a positive potential, the electrostatic field produced by the grid and electrodes accelerates electrons emitted by the cathode toward the target material on the walls of indentation 30. The x-rays generated on impact pass through the window via the hollow electrode.

Fig. 1.7. Coaxial x-ray tube.

Section 2. Electronics

Case No.: **IL-5201**

Title: **STEPPING MOTOR CONTROL CIRCUIT**

Inventor: **Howard P. Spracklen**

Stepping motors convert an input electrical pulse into a stepwise rotation of an output shaft. In general, stepping motors provide very accurate and positive control functions in electro-mechanical systems. Previously, stepping motors were controlled either by variable frequency generators which had a limited range, or by on-off controls which tended to cause overshooting or "hunting". Both the variable frequency generators and the on-off controls could be operated by analog inputs, but in general, the results were unsatisfactory for the reasons mentioned.

The present apparatus converts an analog error signal to a digital pulse train which precisely operates a stepping motor. Simultaneously, the digital pulse train is fed to an energy storage device in a feedback loop for nulling the original error signal after enough pulses have been accumulated.

As shown in Fig. 2.1, a sensor 11 detects a parameter to be measured, and converts this to an analog signal. A balanced amplifier 12, capable of detecting deviations from a preset null condition, receives this signal. When an imbalance occurs, the amplifier transmits a signal to the integrating capacitor 22 in the integrator network 13, tending to charge the capacitor. A comparator 14 senses when the voltage of the capacitor exceeds a preset level. When such a level is exceeded, a signal termed a "logic level" is generated on line 21 to the AND gate 16. Simultaneously, the AND gate receives a digital pulse train from a clock pulser 17. As long as the "logic level" exists at the AND gate, a digital pulse train segment is gated through on line 18, thereby causing the motor control 23 to step the motor 25 one increment.

The rotation direction of motor 25 is determined by the polarity of the comparator circuit 14 in evaluating the deviation of the charge on capacitor 22 from a preset condition. A pair of lines 15 carry forward and reverse signals to the motor control, and to a drive pulse detector 20.

The drive pulse detector functions to reset the apparatus by providing a signal through the resistor 24 which tends to discharge the capacitor. The drive pulse detector combines forward and reverse signals from the lines 15, together with signals gated through the AND gate as a basis for discharging the capacitor. The time delay required to null error signals represented by charge accumulation on the capacitor is set to correspond to the exact time required to initiate the action which will eliminate the source of the original error signal emanating from the sensor 11, e.g., opening a vent a necessary amount to cool a hot object to a desired temperature.

As the apparatus is utilized at the Lawrence Livermore Laboratory, the analog error signal is derived from a temperature sensor which detects the temperature deviations in a plutonium critical array. A stepping motor controls the flow of coolant toward the critical array by opening a vent regulating coolant flow. Use of the circuit

described herein precisely controls the coolant flow, so that the temperature of the critical array can be maintained within very close tolerances.

Fig. 2.1. Stepping motor control schematic circuit.

Case No.: IL-5230

Title: SENSING CIRCUIT FOR USE WITH SUPERCONDUCTING COILS

Inventor: Frank V. Harshbarger, Jr.

One of the problems in using superconducting coils is the tendency of the coil windings to return to a normal, i.e. nonsuperconducting, state. Since materials which are very efficient superconductors usually present high impedances at non-superconducting conditions, the return of a coil to a normal condition drastically changes the impedance state of the coil. A superconducting coil carrying thousands of amperes of current which suddenly becomes normal can erupt with explosive force if the coil is not deactivated or shunted. The present circuit senses a change from the superconducting to the normal state in any portion of a superconducting coil, and provides a warning signal.

In Fig. 2,2, the power supply 10 includes a resistance 12 across which a first signal, termed E_g , is derived. Since the power supply feeds the superconducting coil 11, the signal E_g is proportional to the total current going into the superconducting coil 11, and more particularly, to the voltage drops across the conductor splices in that coil. A signal E_t picked off from the potentiometer 13 is proportional to the total voltage across the coil. This signal includes a voltage drop due to the resistance of the conductor splices (E_g), a voltage drop due to the resistance of normal areas (E_n), an inductive or back EMF due to the charging current (E_i), and a voltage drop (E_j), caused by "jumps," sudden flux changes. E_t goes through the secondary transformer winding 16 which is wound in the same sense as the primary winding 11. The secondary winding also picks up the ac signal components, E_w , existing in the coil 11. Since the current passing through the winding 16 is already proportional to this signal, except in the opposite sense, the ac components are removed therefrom. Specifically, the signal derived from transformer winding 16, E_w , is proportional to E_j , the voltage due to flux jumps, and to E_i , the voltage due to the back EMF from charging current. Exact cancellation of the ac components in E_t occurs when the potentiometer setting exactly compensates for the turns ratio between the primary and secondary coils. The remaining signal which goes to the isolation amplifier 15 is proportional only to E_g and E_n . A signal proportional to E_g already exists in amplifier 14, however. The two signals are combined in the difference amplifier 16, which subtracts the voltage due to splices from the combined signal, leaving a signal which is proportional only to the voltage due to normal areas. When this signal exceeds the preset threshold, the signal may be used to activate alarms or otherwise deactivate the coil 11, thereby preventing damage to the coil or to surrounding personnel.

Fig. 2.2. Damage preventing sensing circuit for superconducting coils.

Case No.: IL-5317

Title: DRIFT COMPENSATING CIRCUIT FOR OPERATIONAL AMPLIFIERS

Inventor: Jasper J. Morrison

One of the principal sources of error attending the use of operational amplifiers is their tendency to drift. The causes for amplifier drift include the output resistance of devices connected to the amplifier input, temperature coefficient and stability of the amplifier, and variations in the amplifier power supplies. In ordinary applications, amplifier drift may be compensated for by periodic manual adjustments. High precision applications, however, may not be able to tolerate any drift between adjustments. This circuit continuously detects and compensates for amplifier drift.

Referring to Fig. 2.3, a differential input amplifier 2 is supplied with capacitive feedback 4 to integrate the output of a device 6, such as an accelerometer, connected to the inverting amplifier input 8. The drift compensating circuit of the present invention, shown in dashed block 10, is connected between the amplifier output 12 and the noninverting amplifier input 14. The time constant of the RC network comprising resistor 16 and capacitor 18 determines the frequency response of the drift compensating circuit. The time constant is made long so that the drift compensating circuit reacts to the low frequency drift voltage variations in the amplifier output 12, but ignores the high frequency variations of the desired integrated signal voltage. The output of the RC network is connected to the gate electrode 20 of a field effect transistor (FET) 22. The drain electrode 24 of the FET 22 is connected to the output of a voltage divider 26 to allow initial adjustment of the current flow through the FET. The source electrode 28 of the FET 22 is connected, via a resistor 30 and a resistor 32, to the noninverting amplifier input 14. In addition, the source electrode 28 is connected, via resistor 30, to the output of a voltage divider 34 to allow initial adjustment of the effective resistance and the pinch-off voltage of the FET 22.

In operation, the RC network produces an output voltage signal representative of the amplitude of the drift voltage of the amplifier 2. Responsive to the RC network output signal, gate electrode 20 changes the effective resistance of the FET, thereby changing the current to the amplifier input 14. This change in current is such as to reduce the amplifier drift voltage.

Fig. 2.3. Operational amplifier drift compensating circuit.

Case No.: **IL-5222**

Title: **COMPUTER-TO-TV DISPLAY BUFFER**

Inventors: **Jame L. Littlepage, Stuart P. Stone, Bruce R. Clegg, and
Wayne McMoran**

This device is a data buffer that enables computer data to be displayed on a television monitor. The apparatus includes a magnetic disk memory unit and disk control electronics. The disk memory unit serves a dual function: (1) It provides a time buffer which permits the computer output data rate to be effectively increased to satisfy television display requirements; and (2) the disk memory unit provides the large data storage capacity necessary for television display. The disk control electronics generates all synchronizing and control signals required by the computer, the disk memory, and the television system.

The actual translation of computer data into a television signal is achieved in the following manner. Since the television picture matrix is displayed as a sequence of visible horizontal lines and invisible retraces, the time between any given picture element in a line and the equivalent point in the next line is exactly equal to the sum of the visible and retraced periods of one line. If the computer data is arranged corresponding to the vertical columns of the television picture matrix, then the necessary computer data flow rate can be reduced by a factor equal to the number of lines in the picture matrix. This principle is used by the disk control electronics to control the writing of computer data onto the disk. It follows that the writing of the entire picture matrix will require as many revolutions of the disk as there are picture elements in one line. Because four picture elements are written in parallel, the actual time is reduced to 128 disk revolutions to write an entire picture. The selection of four picture elements which are written in parallel enables the picture elements to be coded as one of 16 possible shades of gray on the television display screen.

The television signal is generated by reading the data in a line-by-line manner. Each picture element is represented by four bits of computer data arranged as the four parallel disk memory tracks. If the picture elements were to be read one at a time, in series, the resulting data flow would be insufficient for the television data requirements. Therefore, four picture elements at a time are obtained by reading 16 tracks in parallel. High speed circuitry in the disk control electronics separates the four picture elements and converts their digital data into a continuous analog signal for television video.

Case No.: **IL-5305**

Title: **PULSE RATE DISCRIMINATOR**

Inventors: Donald D. McCoy, Bruce H. Shue, and Ralph G. Gutmacher

This digital system detects pulsed input signals falling within a preselected range of pulse rates. While completely general in application, the system was specifically designed to improve the signal-to-noise ratio (contrast) of electron microprobe display images. Figure 2.4 shows the present pulse rate discriminator as part of the circuitry in a conventional electron microprobe instrument. Figure 2.5 is a detailed block diagram of the pulse rate discriminator itself.

Referring to Fig. 2.4, the direction of electron beam 2 produced by an electron gun 4 is controlled, via deflection coils 6, by a horizontal and vertical sweep generator 8 to permit left-to-right and top-to-bottom scanning of a sample 10 under test. The sweep generator is also connected to the horizontal and vertical deflection plates of a CRT display device 12 to permit generation of a sweep pattern in synchrony with and similar to the scanning pattern of the sample. Sample 10 typically contains a number of homogeneous areas or phases 14a, 14b, ..., 14n, the dimensions and relative locations of which are to be selectively displayed by the CRT device. When the electron beam strikes the sample, it produces x rays which have an energy characteristic of the particular phase 14 receiving the electron beam. An x-ray detector 16 generates a pulsed signal whose amplitude and pulse rate are representative of the energy and intensity, respectively, of the x rays. In a conventional electron microprobe system, the pulsed signal is selectively amplitude-discriminated by analog means (not shown), such as a multi-channel analyzer and upper and lower amplitude level controls, and applied to the Z-axis input (intensity control input) of the CRT device to generate the images of the associated phases of the sample. The images of the phases so obtained exhibit poor contrast, however, due to the low signal-to-noise ratio (where noise is defined to include pulses due to background and other phases of the sample 10) of the pulsed signals associated with particular phases 14 to be displayed. The present pulse rate discriminator device, generally indicated by block 18, is connected between the output of the x-ray detector 16 and the Z-axis input of the CRT device 12 to achieve improved signal-to-noise ratio of the pulsed signals and, hence, improved contrast of the generated images.

Referring to Fig. 2.5, the main components of the present pulse rate discriminator include a scaler circuit 20 for selectively controlling the lower limit of a predetermined range of pulse rates associated with a particular phase 14 to be displayed, a timing circuit 22 for continuously recycling the scaler circuit and for selectively controlling the upper limit of the predetermined range of pulse rates, and a gating circuit 24 for controlling the passage of input pulses between discriminator input and output in response to the outputs of the scaler circuit and the timing circuit.

Timing circuit 22 comprises a clock 26 adjusted to generate clock pulses at preselected time intervals, and a time delay 28 adjusted to delay each clock pulse by a

fractional value of the time interval between clock pulses and representative of the upper limit of the predetermined range of pulse rates. Scaler circuit 20 comprises a counter 30 for accumulating input pulses between successive clock pulses, a switch register 32 for manually inserting a digital number representative of the lower limit of the predetermined range of pulse rates, and a comparator circuit 34 for producing an output when the accumulated count in counter 30 equals or exceeds the contents of the digital number in register 32. Gating circuit 24 comprises a pair of bistable multi-vibrators 36 and 38 having their reset input connected (not shown) to the output of clock 26, and an AND gate 40 for passing discriminator input pulses when the pulse rate thereof falls within the predetermined range of pulse rates.

In operation, and when the input pulse rate exceeds the predetermined range of pulse rates, comparator 34 produces an output before the delayed clock pulse from time delay 28, thereby disabling AND gate 40, via the set state of bistable 38. When the input pulse rate is less than the predetermined range of pulse rates, comparator 34 produces no output between clock pulses, thereby disabling AND gate 40, via AND gate 44. When the input pulse rate falls within the predetermined range of pulse rates, AND gate 40 is enabled, via the set state of bistable 36 and AND gate 44, and input pulses exceeding the lower range limit set by the contents of register 32 pass AND gate 40 to activate the Z-axis input of the CRT device 12 in Fig. 2.4, thereby generating an image of the associated phase 1.

Fig. 2.4. Electron microprobe display circuit.

Fig. 2.5. Pulse rate discriminator schematic circuit.

Case No.: **IL-5285**

Title: **CHANNEL-COUPLED OSCILLATOR**

Inventor: **William N. Carver**

In general, the frequency of an oscillator is affected by an external load connected to the oscillator. In a typical electrical oscillator circuit, an increase in the external oscillator load causes a decrease in the oscillator output frequency, and, inversely, a decrease in the external oscillator load causes an increase in the oscillator output frequency. If the external load is constant and known, the oscillator circuit may be adjusted to produce a desired output frequency. Maintaining a constant oscillator output frequency despite external load variation, however, requires continuous adjustment of the oscillator circuit. Such compensating adjustments may become difficult or impossible. The general object of this circuit is to stabilize the output frequency of an oscillator under varying external load.

Figure 2.6 is a diagram of the present oscillator circuit. An oscillatory signal is generated in resonant circuit 2 connected across a first gate electrode 4 and the source electrode 6 of a conventional dual gate field effect transistor 8 (MOS FET) via a signal amplitude stabilizing circuit 3. A voltage divider, comprising resistors 10 and 12 connected between a positive potential source 14 and ground reference potential, produces a positive voltage level at a second gate electrode 16 to enable passage of the oscillatory signal to the field effect transistor drain electrode 18 via the transistor channel. The drain electrode is connected, via a coupling capacitor 20 and an output terminal 22, to a variable load 24. An rf coil 26 provides zero dc voltage drop between the potential source and the drain electrode. Variable resistor 28 provides means for adjusting the amplitude of the oscillatory signal at the drain electrode.

Referring to Fig. 2.7, curve 30 graphically illustrates the output frequency of the present oscillator circuit operating at 2 MHz as a function of the magnitude R of the variable load 24. Of particular importance is the asymptotic approach of curve 30 to vertical line 32 at increasing oscillator load values.

Figure 2.8 is a schematic of a related vacuum tube circuit known as the "electron-coupled oscillator" (also called "Dow oscillator"). This type of oscillator employs a multi-grid vacuum tube 40, such as a tetrode tube, with the two grids operating as an oscillator in any conventional manner. In the present "channel-coupled" oscillator, a dual-gate field effect transistor replaces vacuum tube 40 to achieve a nearly constant output frequency under varying load conditions.

Fig. 2.6. Transistorized channel-coupled oscillator.

Fig. 2.7. Frequency characteristic under variable load.

Fig. 2.8. Related vacuum tube oscillator.

Case No.: **IL-5288**

Title: **MAGNETICALLY COUPLED BAND-PASS FILTER**

Inventor: **Rudolph W. Buntentbach**

It is often necessary to pass a band of frequencies between an upper cutoff frequency and a lower cutoff frequency. The device most commonly used to achieve this is the conventional constant-K band-pass filter shown in Fig. 2,9. The inductors 2, 4 and 6 can be air core type inductors, or can be constructed on high permeability metal or ferrite cores. In certain frequency ranges, however, the metal or ferrite core inductors exhibit extremely high loss characteristics (low Q), and filters utilizing such inductors have poor discrimination capabilities. Air core inductors exhibit more favorable loss characteristics (high Q), but filters utilizing such inductors are subject to cross-talk between the individual inductors. Cross-talk can, of course, be minimized by appropriate shielding, but this tends to introduce losses.

Figure 2.10 is a schematic diagram of an improved band-pass filter using magnetic coupling. A capacitor C_1 and a movable air core inductor L_1 are series-connected across a pair of input terminals 8 and 10. Similarly, a movable air core inductor L_2 and a capacitor C_2 are series-connected across a pair of output terminals 12 and 14. A third movable air core inductor L_3 is connected across a capacitor C_3 and is arranged between the inductors L_1 and L_2 to provide magnetic coupling therebetween. The degree of magnetic coupling, which determines the pass band of the filter, is controlled by adjusting the distances D_1 and D_2 between the movable inductors L_1 , L_2 , and L_3 .

Operation is best described with reference to Fig. 2.11 which illustrates a magnetic equivalent circuit of the filter circuit shown in Fig. 2.10. At the resonant frequency (center of the pass band), P_3 and $N_3^2 C_3$ constitute an essentially zero magnetic impedance connecting the filter input circuit, comprising inductor L_1 and capacitor C_1 , and the filter output circuit, comprising inductor L_2 and capacitor C_2 . At frequencies lower than the resonant frequency, this impedance becomes large due to P_3 . At frequencies higher than the resonant frequency, this impedance becomes large due to $N_3^2 C_3$. Thus, the inductors L_1 and L_2 are magnetically connected only within a band about the resonant frequency (center of the pass band).

The present magnetically coupled band-pass filter has the same impedance and attenuation properties as the conventional constant-K band-pass filter shown in Fig. 2,9. In fact, the present filter must be used in the same manner as the conventional constant-K band-pass filters, i.e., it must be inserted between specific generator and load impedances. The present filter readily lends itself to impedance transformations. By making $L_1 \neq L_2$, $C_1 \neq C_2$, and $L_1 C_1 = L_2 C_2$, the input and output impedance may be made different.

When inductors with magnetic material cores are employed, the embodiment schematically shown in Fig. 2.12 further offers economy in components, i.e., only two cores are required, whereas three cores would be required in the conventional filter

circuit shown in Fig. 2.9. Again, the magnetically coupled filter circuit of Fig. 2.12 is adaptable to impedance transformations.

Fig. 2.9. Conventional constant-K band-pass filter.

Fig. 2.10. Magnetically coupled band-pass filter.

Fig. 2.11. Magnetic equivalent of filter circuit Fig. 2.10.

Fig. 2.12. Band-pass filter using magnetic cores.

Case No.: **IL-5308**

Title: **CONTINUITY CHECKING DEVICE**

Inventor: **Robert P. Rumble**

Conventional continuity checking devices for tracing a desired wire in a bundle of unmarked identical wires typically comprise a battery and a lamp or buzzer series-connected between a pair of test leads. The use of such devices is time consuming and requires removal of wire insulation at both ends of the bundle to obtain direct electrical connection to the test leads.

The present continuity checking device, schematically shown in Fig. 2.13, comprises an oscillator, a phase discriminating detector, and a signal probe. The oscillator 2 produces in- and out-of-phase output signals on lines 4 and 6, respectively. The probe 8 houses a lamp 10 and a signal-sensing line 12. The phase-discriminating detector 14 has a pair of inputs connected to line 4 and sensing line 12 for producing an output to lamp 10 proportional to the component of the signal on line 12 which is in phase with the signal on line 4.

The in-phase oscillator output signal is directly applied, via line 4, to one end of a wire to be traced in a bundle of wires 16 under test. The out-of-phase oscillator output is capacitively coupled to each of the wires of the bundle 16 by means of a flexible conductor 18, such as a strip of foil, connected to line 6 and wrapped around the bundle.

In operation, when the probe comes near (about 0.5 inch or less) to the wire to be traced, line 12 is capacitively coupled to the in-phase oscillator output signal and the light output of lamp 10 increases, since both inputs to detector 14 are in phase. When the probe 8 is brought near the other wires in the bundle 16, the light output of lamp 10 remains small, since the two inputs to detector 14 are out-of-phase.

An important feature of the present device is the application of the out-of-phase signals to the bundle and the wire to be traced. This prevents capacitive coupling of the signal on the wire to be traced to the other wires in the bundle.

Fig. 2.13. Application of continuity checking device.

Case No.: **IL-5334**

Title: **STEPPING SWITCH**

Inventor: **Stanley R. Bishop**

This development is a stepping switch for selectively connecting a number of input signals to a single output line. The present switch design eliminates electromechanical components which are major sources of electrical noise in conventional stepping switches.

Fig. 2.14 illustrates a conventional electromechanical stepping switch 3 in combination with an n -stage binary counter 4 and a clock 6 for automating the operation of a rotatable sample changer. Each of the n switch input terminals is connected to the set output of one of the n counter stages. The switch output terminal 12 is connected to a control circuit 14 which has two output lines 16 and 18 for controlling the stepping switch and the sample holder 8, respectively. With the counter in the reset condition, the first sample on sample holder placed below the analyzer 10 and the stepping switch positioned to connect the set output of the first counter state (2^0) to the input of the control circuit 14, clock 6 and analyzer 10 are activated to start operation. When counter 4 stores the first pulse from clock 6, the set output of the first counter stage (2^0) provides an input to control circuit 14, which, in response thereto, rotates the sample holder 8 such that the second sample is positioned below analyzer 10 and advances the switch 3, via an amplifier 20 and switch coil 22, to connect the movable switch arm with the set output of the second counter state (2^1). The second pulse from clock 6, when added to the counter contents, produces a signal at the set output of the second counter stage (2^1) which causes rotation of the sample holder 8 to position the third sample below the analyzer 10 and advances the switch 3 to connect the movable switch arm with the set output of the third counter stage (2^2). Similarly, the set outputs of the third and higher order counter stages produce signals to rotate sample holder 8 and to advance the switch 3. Thus, in general, the m -th sample is analyzed after a time interval equal to 2^{m-1} times the period of clock 6.

With the counter 4 and control circuit 14 constructed of transistorized or integrated circuits, which typically exhibit low noise rejection capabilities, the use of the conventional stepping switch poses major problems due to the large noise pulses generated by the switch coil 22 and contact bounce of the switch arm.

The stepping switch of the present design, shown in dashed block 24 of Fig. 2.15, provides a low noise switching means for improved operation with transistorized or integrated circuits, such as the counter and the control circuit of Fig. 2.14, connected to its input and output terminals. The set output of each stage of the counter 4 is connected to a first or set input of one of n bistable multivibrators comprising a pair of cross-coupled NAND gates G. A second or reset input of each bistable is connected to ground potential via a manual switch 26. The set output of each bistable is connected, via a coupling capacitor C, to the input of an OR gate 28 which has its output

connected to the input of the control circuit 14. Before operation, the switch 26 is momentarily closed to reset the counter 4 and each of the bistables.

In operation, the first counter state (2^0) when set by a pulse from clock 6, triggers the associated bistable to the set state. The set output of the bistable activates the control circuit 14, via capacitor C and OR gate 28, to rotate the sample holder 8 as described. Similarly, the second and higher order counter stages set the associated bistables to cause rotation of the sample holder 8. Since each bistable remains in the set state when triggered by the associated counter stage, no additional pulses can be transmitted to the control circuit 14 by bistables in the set state. The n bistables and OR gate 28 thus function as a stepping switch between counter 4 and control circuit 14.

Fig. 2.14. Conventional electromechanical stepping switch.

Fig. 2.15. Electronic stepping switch.

Section 3. Lasers

Case No.: IL-5233

Title: DISPLAY DEVICE FOR INFRARED LASER BEAMS

Inventors: George A. Condas and Don W. Brown

One of the problems in using an infrared laser is that its beam is invisible. A crude method of "seeing" the beam is to place a wooden block in its path; the resulting charred spot on the block is an image of the beam. A more satisfactory display method is to place a phosphor-coated metal plate in the beam and irradiate the phosphor with ultraviolet light. The beam image is displayed as a dark spot on a fluorescent background. The mechanism by which the heat from the beam locally darkens the phosphor is called "thermal quenching".

Conventional fluorescent display devices operate over a limited range of incident laser power. If a phosphor having a high quenching temperature, T_q , is selected, then a low power laser beam may not supply sufficient heat to darken the phosphor. Conversely, if a phosphor having a low T_q is chosen, then high power lasers may overheat the phosphor. The dark spot is smeared out, thus decreasing image resolution.

This apparatus is a fluorescent display device for use over a wide range of incident laser power. The phosphor-coated plate is electrically heated to a temperature $T_p = T_a - T_b$, where T_b is the temperature increase contributed by the impinging laser beam. By adjusting T_p to correspond to the variations in T_b (variations in incident laser power), this equality can always be maintained. For example, in the case of a high-power laser, T_p will be relatively low since most of the temperature increase will be contributed by T_b . Conversely, T_p will be relatively high in the case of a low-power laser. For a given laser power, maximum image resolution occurs at a particular phosphor temperature. Thus, image resolution can be maximized simply by changing the temperature of the phosphor.

The device is shown schematically in Fig. 3.1. A thermocouple (not shown) is fixed to the metal plate to monitor temperature. Using a carbon dioxide laser (10.6 microns wavelength), and a 1-mil layer of a commercial P-2 phosphor ($ZnS: Cu$), a power density of 1 W/cm^2 produced maximum image resolution at a plate temperature of 250°C . A power density of 20 W/cm^2 produced maximum image resolution at 175°C .

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Fig. 3.1. Infrared laser beam display device.

Case No.: **IL-5282**

Title: **DUAL-BEAM HOLOGRAPH CAMERA**

Inventor: **Henry H. M. Chau**

Holography produces a three-dimensional image by recording a coherent light interference pattern on film. Two coherent light wave fronts of the same frequency are required, i.e., a wave front scattered from the object and an unscattered wave front. We call the scattered light wave front the information beam, and the unscattered wave front the reference beam. Only a laser generates coherent light strong enough for recording holograms.

Conventional methods for simultaneously recording more than one hologram of an object use a single highly diverged laser beam illuminating both the object and the recording film. The film is located to intercept information light reflected from the object and reference light directly from the diverged beam. However, a single beam can illuminate only one-half of the object. Accordingly, holograms taken with this technique record only a 2π steradian view of the object.

It is possible to use mirrors to reflect unscattered light to illuminate the other half of the object, thereby enabling perspective views of the object greater than 2π steradians to be recorded. This scheme is unsatisfactory, however. Because the beam is diverging, the object is not uniformly illuminated. Moreover, the beam is directional, hence irregularities on the surface of the object create very dark and sharp shadows. Also, light reflected from a mirror surface is essentially scattered. Irregularities in the mirror surface are also included in the hologram, significantly degrading its resolution.

This device is a holograph camera for simultaneously recording a full 4π steradian view of an object. Two highly diverged coherent light beams from a laser illuminate opposite sides of the object. Several film plates located around the object intercept scattered (information) light from one beam, and unscattered (reference) light from the other beam. Fast, high resolution holograms are possible because the dual beams illuminate the object uniformly.

In Fig. 3.2, the beam from a laser 11 is split equally into two beams 12 and 13. Mirrors 14 direct each beam into high power (60X) microscope objective lenses 16 which diverge the beams. Pinholes 15 ($5\ \mu$ in diameter) are placed at the focal points of the lenses 16 to "clean" the beams. The diverged beams illuminate opposite sides of the object 17. Film plates 18 are placed around the object such that each intercepts unscattered light from one beam yet can receive only scattered light from the other beam. For example, the film plate 18a intercepts unscattered light from diverged beam 13, but does not intercept any unscattered light from beam 12. On the other hand, light scattered by the object from beam 12 will be incident on the film plate 18a. Although the drawing shows one film plate per $\frac{\pi}{2}$ steradians of the viewing sphere, obviously more film plates can be used.

Fig. 3.2. Dual-beam holograph camera.

Case No.: **IL-5315**

Title: **METHOD FOR WIDE BAND LASER MODULATION**

Inventor: **Oved S. Zucker**

Using the inherently large channel capacity of laser beams in communications requires a wide band laser modulation method. The present modulation method is a frequency-division multiplexing process; many messages are simultaneously transmitted on a common laser beam by using a different frequency band for each signal.

Referring to Fig. 3.3, the laser output beam B of a continuous laser 2 is divided into m laser beams and applied to the input of m identical frequency-division multiplexers $X_1, X_2, \dots, X_m$. At each frequency-division multiplexer, the input laser beam is further subdivided into n component laser beams $B_1, B_2, \dots, B_n$. Message signals $S_1, S_2, \dots, S_n$ each independently modulate one of the component laser beams by means of modulators $M_1, M_2, \dots, M_n$ to produce modulated component beams, each having a bandwidth f_b centered about the laser beam frequency f_0 . Modulated laser beam B_1 is applied to a frequency shifter F_1 , to be described in more detail hereinafter, where the laser beam is shifted in frequency by f_b to produce a frequency-shifted laser beam B_1' . Modulated laser beam B_2 is added to modulated laser beam B_1' , and the resulting laser beam $(B_1' + B_2)$ is applied to frequency-shifter F_2 to produce a frequency-shifted laser beam $(B_1' + B_2')$. Similarly, each modulated beam $B_3, \dots, B_n$ is added to the summed modulated laser beams of the preceding stages and shifted by f_b to produce a frequency multiplexed output beam $(B_1' + B_2' + \dots + B_n)$ with a bandwidth of nf_b . The output of the multiplexer X_1 is shifted by nf_b and added to the output of multiplexer X_2 . Similarly, the output of each multiplexer $X_3, \dots, X_m$ is added to the summed output of the preceding multiplexers to produce a final output laser beam with a bandwidth mnf_b . Both m and n are selected such that the bandwidth mnf_b of the final output covers the available bandwidth of the original laser beam B.

In a second embodiment (Fig. 3.4) of the multiplexers $X_1, X_2, \dots, X_m$, the outputs of the odd-numbered modulators $M_1, M_3, \dots$ are initially shifted by f_b , and the outputs of the even-numbered modulators $M_2, M_4, \dots$ are initially shifted by a multiple of f_b . Inspection of Fig. 3.4 shows that for $n \geq 4$, the total number of frequency shifters along the path of a particular modulated laser beam in a multiplexer is reduced as compared to that of the corresponding modulated laser beam of Fig. 3.3. For example, in a multiplexer with $n = 8$, the output of modulator M_1 of Fig. 3.4 is applied to only 3 frequency shifters (F_1, F_2 , and F_4) as compared to 6 frequency shifters for the output of modulator M_1 of Fig. 3.3. Such reduction in the number of frequency shifters along the path of a particular laser beam between the multiplexer input and output is advantageous, since imperfections in a shifter introduce certain errors in the frequency-shifted output which are independent of the total frequency shift produced by the shifter. Hence, the cumulative effect of these errors is reduced by reducing the number of shifters.

Figure 3.5 illustrates a first embodiment of the frequency shifters in the multiplexers of Fig. 3.3 and Fig. 3.4. The output 4 of an acoustic transducer 5, incident upon a crystal 6, produces within the crystal a traveling acoustic wave which, in turn, produces an index of refraction wave traveling with the acoustic wave. A modulated (or unmodulated) laser beam 8, to be shifted in frequency by f_b or a multiple thereof and applied to the crystal 6 in a direction opposite to that of the acoustic transducer output 4, is reflected off the crests of the index of refraction wave. By adjusting the frequency of the acoustic output 4 such that the wavelength of the index of refraction wave equals a multiple of wavelength of the laser beam 8, constructive interference takes place in the reflection of the laser beam 8. Since the laser beam 8 is reflected off the crests of the traveling index of refraction wave (analogous to reflection off a Fabry-Perot pattern traveling at the speed of the acoustic wave), a Doppler frequency shift occurs which is a function of the velocity of the acoustic wave in the crystal and the wavelength of the laser beam. Proper selection of the acoustic wave velocity in the crystal results in the desired frequency shift f_b or a multiple thereof. The maximum obtainable frequency shift with the shifter device of Fig. 3.5 is approximately 1 GHz.

Fig. 3.6 illustrates a second embodiment of the frequency shifters. Two laser beams 10 and 12 of different frequencies ω_1 and ω_2 are superimposed in space to generate an intensity pattern having a speed and direction dependent on the angle θ between the two laser beams and the laser frequencies ω_1 and ω_2 . A crystal 14, whose index of refraction can be changed in response to a varying electromagnetic field intensity, is placed in the generated intensity pattern to produce within the crystal a traveling index of refraction wave. A modulated (or unmodulated) laser beam 16, to be shifted in frequency by f_b or a multiple thereof and applied to the crystal 14 in a direction normal to the index of refraction wave, is reflected off the index of refraction wave. By adjusting θ , ω_1 , and ω_2 such that the wavelength of the index of refraction wave equals a multiple of the wavelength of the laser beam 16 and the velocity of the index of refraction wave equals a selected value, constructive interference and the desired Doppler frequency shift of f_b or a multiple thereof are obtained in the reflection of the laser beam 16. Frequency shifts as high as 10 GHz can be obtained with the shifter device of Fig. 3.6.

Fig. 3.3, Laser frequency-division multiplexer.

Fig. 3.5. Laser frequency shifter, acoustic.

Fig. 3.6. Laser frequency shifter, refractive index wave.

Case No.: **IL-5325**

Title: **SYNCHRONOUS COMMUTATOR SYSTEM**

Inventor: *Oved S. Zucker*

This device is a commutator for synchronizing the transmission and reception of information between wideband data systems. As shown in Fig. 3.7, a data source 2, such as a digital computer, produces a 1-Gigahertz (GHz) pulse code modulated data signal which is to be transmitted, via a communication laser channel 4, to a distant data storage system 6, such as another digital computer. Although the laser channel 4 has a bandwidth far exceeding the required bandwidth for transmitting the 1-GHz data signal, presently existing modulators for placing signals on a laser channel have a maximum bandwidth of only about 0.5 GHz. The present synchronous commutator overcomes this modulator bandwidth limitation by separation and subsequent recombination of the original data signal as explained below.

A commutating switch 8, operating at 1 GHz under control of an adjustable clock 10, divides the 1-GHz data signal applied thereto via a buffer 3 into 0.5 GHz signals. Each 0.5 GHz signal is applied to a 0.5 GHz modulator of a frequency-division multiplex transmitter system 12, such as the wide band laser modulation system disclosed in IL-5315 (pg. 3-5). The resulting multiplexed output of system 12 is transmitted, via laser channel 4, to a distant multiplex receiver system 14 where the two 0.5 GHz signals are recovered. A second 1-GHz commutation switch 16, operating under control of an adjustable clock 18, combines the two 0.5 GHz signals into the original 1-GHz signal which is stored in data system 3.

Combining the two 0.5 GHz signals at the receiving end of the transmission channel to produce the original data signal requires synchronization of the two commutators 8 and 16 to better than 1 nanosecond, since both commutators run at a rate of 1 GHz. The required synchronization is achieved by synchronizing both clocks 10 and 18 to the pulsed output of a mode-locked laser 20. In particular, mode-locked laser 20 produces subnanosecond pulses at a rate of about 100 megahertz. A beam splitter 22 diverts part of the mode-locked laser pulsed output to synchronize clock 10, via a laser pulse detector 24, and transmits the remainder of the pulsed output to the receiving end of the data transmission channel to synchronize clock 18, via a laser pulse detector 26. Commutators 8 and 16, controlled by clocks 10 and 18, respectively, are now synchronized to better than 1 nanosecond. Buffer 3 has a sufficiently large data storage capacity to prevent buffer overflow or depletion while it slowly adjusts the output rate of mode-locked laser 20, via line 28, in response to variations in the 1 GHz output rate of data source 2.

Fig. 3.7. Synchronous commutator system.

Case No.: **IL-5212**

Title: **TUNABLE DYE LASER**

Inventors: **David W. Gregg and Scott J. Thomas**

Lasers would make excellent light sources for various analytical instruments except that they emit light only in a very narrow frequency band. A laser that can be "tuned" over a wide frequency range is needed. An operator could then select any desired frequency within the range.

This apparatus is a tunable laser system in which different organic dye solutions are selectively circulated through a single optical cavity. The dye solutions lase in overlapping frequency bands to provide a wide overall frequency range. Fine frequency selection within a particular frequency band is done with a grating mounted within the cavity.

Figure 3.8 shows a tube 13 and two end pieces 14 enclosed in a chamber 11. The end pieces 14 include ports 15, through which the dye solution enters and leaves the chamber. Each end piece also has a window 16 axially aligned with the tube axis. The windows are optically flat, and are secured to the end pieces by a triangular screw mounting 17. By adjusting the screws on the mounting, the windows can be adjusted to be optically parallel. O-ring seals 19 are placed between the tube and end pieces, and between end pieces and windows. Flashlamps 20 are placed around the tube in a reflector housing 21. Ideally, the flashlamps include mercury cathodes. Vaporized mercury will not plate onto the envelope, and will dissolve away any materials vaporized from the anode. Thus, the usable lifetime of the flashlamps is greatly enhanced. Both the tube and the flashlamp envelopes are composed of a material which absorbs ultraviolet light, thereby preventing this cause of dye solution deterioration.

Referring to Fig. 3.9, a solenoid-driven sparkgap 23 switches charged capacitors 24 for driving the flashlamps. The rate at which the sparkgap is driven is variable, and is adjusted by a sparkgap driver 25. A charging circuit charges the capacitors. Containers 30 serve as reservoirs for the various different dye solutions. A pump 32 circulates the dye solutions through the chamber. The particular dye solution is selected by manual valves 33. A mirror 27 and a grating 28 are placed at opposite ends outside the chamber to define the optical lasing cavity. The grating is mounted on a rotating table.

To operate the system, an organic dye solution is circulated into and through the chamber. The dye solution generates a laser light pulse responsive to pumping flashes from the flashlamps. The pulse repetition rate is determined by the rate at which the sparkgap is driven.

Each organic dye solution lases within a characteristic spectral region. The spectral width of this region is normally on the order of 100-200 Å. However, the grating provides a frequency-dependent optical feedback to the dye solution in the cavity which narrows the spectral width of the generated laser pulse to 1-2 Å without loss of power. The grating is simply rotated to select the desired wavelengths (frequency)

within the characteristic spectral lasing region of the dye solution. Specifically, the grating provides a continuous fine-tuning means across the relatively broad characteristic spectral lasing region of the dye.

To change the frequency or spectral region, the original dye solution is returned to its container, emptying the chamber. A solvent for that particular dye is then circulated through the chamber, rinsing it thoroughly. Another solution which lases in the desired spectral frequency region is then circulated into the chamber, and the grating is adjusted or replaced if it is not effective in the spectral or frequency region of the second dye solution.

All dye solutions which lase when optically pumped by flashlamps can be used in the described laser system. The following table lists a possible combination of organic solutions, their respective lasing bands, and their respective solvents.

<u>Organic Compound</u>	<u>Lasing Band-Å</u>	<u>Solvent</u>
4-methyl umbelliferone	4490-4590	Water
9-aminoacridine hydrochloride	4570-4600	Ethyl alcohol water (acid)
7-diethylamino-4-methyl coumarin	4550-4650	Ethyl alcohol
Benzyl β methyl-umbelliferone	4635-4680	Ethyl alcohol water (base)
Fluorescein	5450-5550	Water
Rhodamine S	5780-5945	Ethyl alcohol
Acridine Red	5965-6075	Ethyl alcohol
Rhodamine B	6050-6500	Ethyl alcohol
Pyronine B ^a	6150-6320	Ethyl alcohol
Kiton Red S	6100-6500	Ethyl alcohol

^a Reference to a company or product name does not imply approval or recommendation of the product by the University of California or the U. S. Atomic Energy Commission to the exclusion of others that may be suitable.

Fig. 3.8. Tunable dye laser construction.

Fig. 3.9. Tunable dye laser circuit.

Case No.: **IL-5270**

Title: **GAS LASER PUMPED BY ELECTRON BEAM**

Inventors: **Barton Krawetz, David W. Gregg, and Ray E. Kidder**

Gas lasers are ideal for high energy applications because a gas lasing medium is not damaged by the intense light beam. It is difficult, however, to obtain and maintain a large population inversion in a lasing energy state of a gaseous medium. For example, optical pumps such as flashlamps simply cannot transfer energy quickly enough to create a large population inversion in a gas.

Presently, electrical discharge excitation is the most efficient technique for pumping a gaseous lasing medium. In an electrical discharge, the lasing gas is both directly excited by electron collision and excited by resonant energy transfer from a second gas excited by electron collision. However, uniform, stable electrical discharges exist only at low (vacuum) pressures. Hence, the energy density of a gas laser system is relatively low.

This apparatus is a high power, high pressure gas laser system wherein an energetic charged particle pulse from an accelerating device is directed through a gaseous medium to initiate a uniform electrical discharge between two large plate electrodes. Due to the very short time-width and uniformity of the charged particle pulse and the triggered discharge, extremely large quantities of energy can be uniformly delivered to the lasing medium with very small net free-electron displacement. Laser light pulses having an average energy content on the order of 20,000 J can be generated in this manner. Total pulse duration is on the order of 0.5 μ sec, giving a pulse power rating on the order of 10^4 MW.

Alternatively, the short, high energy charged particle pulse could uniformly ignite a high pressure gaseous mixture placed in an optical cavity. The energy released in the conflagration creates a large population inversion in the lasing energy states of excited chemical species existing during the explosion.

Referring to Fig. 3.10, the high energy gas laser system includes a pulsed electron accelerator 12, a pressure vessel 13 formed of two large plate electrodes 14, two opposite insulating sidewalls 15 transparent to light, and two opposite sidewalls 16 transparent to high energy electrons. The vessel contains a pressurized, gaseous lasing medium 11. Mirrors 17 which can be placed within the vessel, integral with the vessel walls, or outside the vessel, define an optical cavity between the plate electrodes 14. The vessel 13 is placed in the electron beam path 18. Conventional means 19 disperse the electron beam such that it has the same cross sectional area as the pressure vessel. A capacitor bank 20 is electrically connected to the plate electrodes. Conventional electronic means 21 charges the capacitor bank. An exemplary gaseous lasing medium consists of 40 parts CO_2 , 50 parts N_2 , and 10 parts He.

In operation, the capacitor bank is charged until the voltage impressed across the vessel is slightly less than the breakdown voltage of the gaseous lasing medium. A short, high energy (several MeV) electron pulse is generated by the accelerator 12.

The beam passes through the gaseous lasing medium with approximately the speed of light, ionizing it slightly by electron collision, thereby triggering a large, uniform electrical discharge between the plate electrodes. Both the electron pulse and the triggered electrical discharge excite the lasing medium, creating a large population inversion. Specifically, for the CO_2 - N_2 -He lasing gas mixture, the beam electrons and discharge electrons excite the CO_2 , N_2 and He to high energy states by collision. The excited N_2 resonantly transfers energy to unexcited CO_2 . Accordingly, a population inversion is generated in the lasing energy state of the CO_2 molecules. Spontaneous emission from the excited CO_2 initiates lasing.

The alternative, explosive configuration of the laser is shown in Fig. 3.11. Here the electron beam accelerator 30 directs a high energy electron beam pulse through a contained, high pressure, explosive gaseous mixture 33 in an optical cavity. The mirrors 34 define the cavity. The electron beam pulse uniformly ignites the explosive gas mixture. Excited chemical species which have a lasing capability are created in the explosion. The short timewidth of the initiating electron pulse, and the uniformity of initiation, creates a population inversion in the energy states of the lasing free radical chemical species. Alternatively, an inert gas such as carbon dioxide can be included in the explosive gas mixture. The excited species existing during the explosion transfer energy to the inert gas, thereby generating a population inversion suitable for lasing.

Fig. 3.10. Electrical discharge gas laser pumped by electron beam.

Fig. 3.11. Explosive gas laser pumped by electron beam.

Case No.: **IL-5318**

Title: **NITROGEN-FLUORINE-HYDROGEN CHEMICAL GAS LASER**

Inventors: **David W. Gregg and Richard K. Pearson**

Energy released in an exothermic chemical reaction can be used to "pump" a lasing medium which is a chemical species in the reaction. For example, the reaction of molecular hydrogen (H_2) and atomic fluorine (F) generates excited hydrogen fluoride (HF) which will lase (Aviation Week and Space Technology, 1/12/70). However, high power pulsed chemical lasers have not been considered practical for a variety of reasons which are discussed below.

First, the chemical reaction rate inherently determines the pulse rise time. Specifically, the rise time of the laser pulse is dependent on the degree of population inversion per unit volume in the lasing medium and the photon density existing when lasing initiates. The spontaneous emission from the lasing energy state determines the initial photon density. During lasing, this photon density (energy) grows exponentially in time, since the rate of photon production is directly proportional to the photon density already present. However, as photon density increases, the degree of population inversion decreases. Hence, as the degree of inversion drops, so does the photon density. Since the degree of population inversion is a function of the differential rates of energy gained due to "pumping" and energy loss due to lasing and spontaneous emission, the "rate of pumping" ultimately determines the rise time of a laser pulse. In the chemical laser, the rate of pumping is a chemical reaction rate. Accordingly, for a fast rise time laser pulse, there must be an exceedingly fast chemical pumping reaction.

Also, the peak power per laser pulse and the time width of the generated pulse are dependent on the chemical reaction rate. Specifically, spontaneous emission from the lasing energy state of a lasing species occurs at a rate proportional to the degree of population inversion. To obtain the population inversion, the "pump rate" into the lasing energy state must initially exceed the spontaneous emission rate from that energy state. Subsequently, an equilibrium status is reached at which the spontaneous emission rate equals the pumping rate. Since lasing depletes the inversion, it also decreases the spontaneous emission rate. Accordingly, the rate at which energy can be tapped by lasing theoretically equals the difference between the spontaneous emission rate during lasing and the pump rate from and into the lasing energy state, respectively. Accordingly, to obtain a very high power laser pulse, the chemical reaction rate must greatly exceed the spontaneous emission rate from the lasing energy state of the lasing species. The time width of the generated pulse is determined by the amount of energy which can be tapped by lasing and the rate at which it can be tapped.

Another problem with high power, pulsed chemical lasers relates to the non-uniformity of fast exothermic chemical reactions. Where such reactions are initiated at a particular point, the reaction spreads spatially from that point, generating shock

waves. Accordingly, at any instant of time during an exothermic chemical reaction, there are at least three different volume regions: (1) the region of unreacted chemical species; (2) the region of reacting chemical species; and (3) the region of reaction products. Each of these regions has a different composition, temperature, and density. Hence, each region must be considered a different medium, with different light-transmitting properties. The generated shock waves introduce still another variable in the light-transmitting properties across an exothermic chemical reaction. However, to obtain a high efficiency light output from a lasing cavity, the medium within the cavity must be fairly uniform. Accordingly, it is practically impossible to obtain a lasing output across an exothermic chemical reaction initiated in the typical fashion.

One technique for providing uniformity necessary in a lasing cavity of a chemical laser is to uniformly initiate the chemical reaction throughout the lasing cavity volume. This may be accomplished by flash-photolysis, using a very intense light source, or by an electron beam (see Invention Case IL-5270 pg. 3-16). Another technique would be to locate the lasing cavity across only one of the volume regions of the chemical reaction (see Aviation Week and Space Technology, Jan. 12, 1970, p. 17).

In summary, only an exothermic chemical reaction which has a lasing reaction species or product, which releases energy suitable for pumping that lasing species at a sufficient rate to generate an extremely large population inversion, and which can be uniformly and controllably initiated, is suitable for high power pulsed lasing systems. Additional desirable properties of a laser pumping chemical reaction are a low initial energy investment to obtain lasing, gaseous reaction products, and complete reaction of reactants.

A chemical reaction which meets the above conditions has been found. Specifically, an explosive, gaseous mixture of tetrafluorohydrazine (N_2F_4) and molecular hydrogen (H_2) can be controllably and uniformly ignited by flash-photolysis, or by a high energy electron beam pulse. The reaction generates hydrogen fluoride, a lasing species, in an excited state. The reaction rate is extremely fast (a branching chain explosive reaction is hypothesized). Laser pulses with rise times less than 20 nsec have been generated. Also, the total energy investment beyond that released in the reaction to obtain lasing is very low. The reaction, once initiated, continues until the reactants are completely combusted. Moreover, the reaction products are all gases. Finally, the reaction of H_2 and N_2F_4 has an energy capacity on the order of megajoules per kilogram. The pumping reaction will also work with H_2 plus other nitrogen-fluorine compounds such as NF_3 and N_2F_2 .

It is hypothesized that the rise time of the generated laser pulse is inversely related to the pressure of the gaseous mixture, on the theory that a chemical reaction rate is proportional to the collision rate of the reactant particles. The collision rate increases with pressure. Accordingly, the reaction rate must also increase. Recalling that the reaction rate determines the rise time of the generated laser pulse, it has been shown in experiments that, for a flash-photolysis initiated reaction, increasing the initial gas pressure from 5 Torr to 10 Torr significantly decreased the rise

time of the generated laser pulse (from 200 nsec to 100 nsec). Accordingly, at least initial evidence indicates the above hypothesis is correct.

The $N_2F_4-H_2$ mixture can be ignited by flash-photolysis. An array of flashlamps disposed around a lasing cavity containing N_2F_4 and H_2 are simultaneously fired to uniformly initiate the chemical reaction. This technique of initiation has inherent disadvantages in that light penetration into the $N_2F_4-H_2$ mixture is dependent on density. Accordingly, the technique introduces inherent limits on the density of the explosive mixture, and on the dimensions of the lasing cavity. The rise time of the laser pulses generated by this technique are on the order of 100 nsec.

Another technique involves initiating the explosive gas mixture by means of a high energy electron beam pulse. For an outline of this technique, see Case No. IL-5270, pg. 3-16. Using this technique, laser pulses with rise times less than 20 nsec were obtained.

Section 4. Materials, Fabrication

Case No.: IL-5281

Title: Be_4B ARMOR

Inventors: Carl F. Cline and Mark L. Wilkins

"Composite armor" consists of a ceramic facing bonded to a supporting backing. The hard ceramic blunts and shatters the projectile on impact; the backing supports the ceramic against the resulting stresses. For ease in fabrication, it would be desirable if the ceramic portion of the armor could be made directly by casting molten ceramic into a suitable mold. However, the high melting temperature ($> 2000^\circ\text{C}$) of most ceramics virtually rules out casting as a feasible fabrication method. Consequently, composite armor is currently made by hot-pressing or sintering a ceramic powder and then bonding the resulting ceramic compact to a backing material such as fiberglass.

This development is a lightweight personnel armor based on the beryllium-boron system. It has been found that hot-pressed Be_4B bonded to a fiberglass backing provides a composite armor which is 15% more effective than conventional B_4C /fiberglass armor. Even more important, however, is the realization that the existence of a particular eutectic in the Be/B system permits the direct casting of a two-phase material consisting of Be_4B crystals dispersed in a Be matrix. These two aspects of the development (hot-pressed Be_4B and cast $\text{Be}_4\text{B}/\text{Be}$) are discussed separately in the following paragraphs.

Be_4B powder (prepared by heating a mixture of Be and B powders) was vibrated into a graphite die and heated under an argon atmosphere for 15 minutes at 1040°C and 5600 psi. The resulting Be_4B compact, 0.317 inch thick, was bonded with polyurethane adhesive to a 0.125 inch thick fiberglass backing. The areal density of the $\text{Be}_4\text{B}/\text{fiberglass}$ composite was 4.38 lb/ft^2 . Ballistic tests of the composite showed it to be 15% more effective (in terms of projectile velocity required for penetration) than conventional $\text{B}_4\text{C}/\text{fiberglass}$ armor of equal areal density.

Figure 4.1 is a simplified and abbreviated phase diagram of the beryllium-boron system. The existence of a eutectic $\underline{e}$ (m. p. = 1110°C) at 10 atomic % (at. %) boron raises the possibility of making a two-phase armor material by direct casting from the liquid state. To see how this could be done, consider the line of constant composition $\underline{abcd}$ drawn in at 15 at. % B. At $\underline{a}$, the system is entirely liquid, that is, a liquid solution of Be containing 15 at. % B. As the system is cooled, solid Be_4B starts coming out of the solution at $\underline{b}$. As the Be_4B continues crystallizing out with decreasing temperature (the system is now in the region labeled $\text{L} + \text{Be}_4\text{B}$), the saturated liquid becomes richer in Be. The liquid composition thus moves along the liquidus line $\underline{be}$. At $\underline{c}$, the solution has the eutectic composition $\underline{e}$ and is saturated with respect to Be also, so solid Be begins to precipitate. As heat is withdrawn from the system, the amount of eutectic liquid (composition $\underline{e}$) diminishes and the amounts of solid Be_4B and solid Be increase. Even though heat is withdrawn from the system, the temperature remains constant at 1110°C as long as there is still some eutectic liquid remaining. Finally, the last drop of liquid solidifies, and the temperature of the mixed solids

decreases along the line cd. The result is a two-phase solid consisting of Be_4B crystals dispersed in a Be matrix, the overall composition of the solid being 85 at. % Be and 15 at. % B.

For purposes of illustration, the constant composition line abcd is shown at 15 at. % B in the attached Be/B phase diagram. However, abcd could be fixed anywhere between 10 at. % B (the eutectic composition) and 20 at. % B (the Be_4B composition). Thus, the final two-phase material can be made to have any desired composition between 10 and 20 at. % B simply by cooling a Be/B solution of the same desired composition.

One of the features of the two-phase Be_4B /Be material is that the Be_4B crystals can be uniformly aligned within the Be matrix by establishing a thermal gradient along the Be/B solution as it cools. As the Be_4B crystals come out of solution, they are uniformly aligned in the direction of the thermal gradient. This method is called "directional solidification" and has been used to produce materials having enhanced properties in a given direction. For applications to armor, the structure would be oriented to provide maximum resistance to penetration. It has been previously established that a balance between high hardness and ductility will result in the most efficient armor material (UCRL-50694). Directional solidification offers the possibility of orienting the high hardness of the ceramic phase and the ductility of the metal phase to provide the optimum material properties to defeat a ballistic impact.

In summary, two distinct armor materials have been developed based on the beryllium/boron system. The first material is hot-pressed Be_4B bonded to a fiberglass backing. This material has been tested ballistically and shown to be 15% more effective than conventional B_4C /fiberglass armor. The second material is a two-phase dispersion of Be_4B crystals in a Be matrix. The chief advantage of this material over other ceramics is that it can be cast directly from the liquid state. This permits large, complex armor shapes to be made cheaply and easily. In addition, the composition of the Be_4B /Be material can be varied between 10 and 20 at. % B as desired. Application of a thermal gradient to the cooling Be/B solution uniformly aligns the Be_4B crystals to provide increased penetration resistance in a given direction.

Fig. 4.1. Phase diagram, beryllium-boron system.

Case No.: IL-5297

Title: HONEYCOMB ARMOR

Inventors: Mark L. Wilkins and Carl F. Cline

When a metal projectile strikes a ceramic plate, two competing mechanisms are set into operation. First, the projectile impact causes cracks to propagate through the ceramic. The cracks eventually puncture the ceramic making it unable to resist penetration by the projectile. Simultaneously, the projectile is blunted and abraded by the ceramic, thereby decreasing its penetration effectiveness. Of course, the projectile can only be abraded if the ceramic remains essentially intact. It is thus apparent that the cracking and abrasion mechanisms are antagonistic. Consequently, a major concern in designing ceramic armor is to minimize or delay the cracking of the ceramic so that maximum abrasion of the projectile will occur.

In this design, cracks in the ceramic are delayed and deflected by distributing numerous thin metal surfaces throughout the ceramic. This is done by filling the voids in a metal honeycomb with ceramic powder, and then hot-pressing the ceramic-filled honeycomb into a unitary structure. Figures 4.2 and 4.3 show the action of the honeycomb in delaying propagation of cracks in the ceramic. Figure 4.2 shows a projectile striking a ceramic plate without honeycomb reinforcement. Cracks originate at the projectile tip and propagate through the ceramic in the so-called "fracture conoid". Figure 4.3 schematically shows the effect of the metal honeycomb on crack propagation. Some cracks are terminated at the metal interface; others are delayed in passage through the metal; and others change direction. The net effect of the honeycomb is to increase the time that the ceramic remains essentially intact, and hence to increase projectile abrasion. In addition to the crack delay effect, the metal honeycomb also adds ductility to the structure, thus increasing the ceramic's ability to resist the bending load of the projectile.

The essential feature is the distribution of numerous thin metal surfaces throughout the ceramic. These surfaces need not necessarily be in the form of a honeycomb. Strips of corrugated metal foil would serve the same purpose, that is, to blunt the cracks and delay the fracture propagation. With the exception of increasing the composite's ductility, the metal foil serves no structural purpose. The thickness of the foil is fairly critical; too much metal distributed throughout the ceramic could increase the material's compressibility. This would decrease the penetration resistance, and hence is undesirable.

Suitable materials for making the honeycomb armor are: Be honeycomb filled with Be_3B ; and Ti honeycomb filled with either BeO or TiBe_{12} . The Be and Ti honeycombs are commercially available.

Fig. 4.2. Cracks propagating in solid ceramic.

Fig. 4.3. Cracks propagating in honeycomb reinforced ceramic.

Case No.: IL-5298

Title: SCREEN-BONDED ARMOR

Inventors: Mark L. Wilkins and Carl F. Cline

"Composite armor" consists of a ceramic facing bonded to a backup plate. The projectile impact causes the ceramic to crack in a conical pattern called the "fracture conoid." The backup plate supports the fracture conoid after projectile impact so as to delay the breaking up of the ceramic. Eventually the conoid is broken completely, and the load on the backup plate is transmitted over smaller and smaller areas. The backup plate is now required to deflect in order to avoid high shear stresses caused by application of the load over an area smaller than the conoid base.

The broken ceramic is still effective in defeating the projectile when the ceramic is held in place by the backup plate. This is because the broken ceramic is stiff to compression loads, and friction makes it difficult for the projectile to move material aside. Obviously, it would always be better to have a rigid backup plate to hold the broken ceramic in place—except that shear failure becomes the weakest link. Therefore, the optimum backup plate design is a plate just stiff enough to avoid shear failure by the applied load. The energy of the impact is absorbed by deflection of the backup plate.

This design provides a backup plate structure that has the required stiffness during the period of ceramic breakup, yet is capable of subsequent enhanced deflection to absorb the projectile energy and avoid shear failure. As shown in Fig. 4.4, the backup consists of two metal plates bonded together by a metal screen or grid (the screen dimensions are greatly exaggerated for clarity). After the fracture conoid is fully formed (Fig. 4.5), the backup plate supports the broken ceramic with a stiffness proportional to h^3 , where h is the total thickness of the plate/screen/plate sandwich. Figure 4.6 shows the backup plate deflecting under the load transmitted by the broken ceramic. The deflection gives rise to a shear stress in the screen bond; so long as the plates are coupled together by the screen, however, they act as a unitary structure and support the broken ceramic with a stiffness proportional to h^3 . In Fig. 4.7, the shear stress has exceeded the strength of the screen bond to the metal plates. The plates are "decoupled", and they deflect independently. If the thickness of each plate is $h/2$, then the stiffness of each plate is only $1/8$ its former value, $(1/2)^3$. The increased flexibility of the backup at this stage of the impact absorbs projectile energy in much the same way that the "web" in a fielder's glove absorbs the energy of a baseball. The breaking shear stress of the screen bond depends on the screen material and the manner in which it is bonded to the metal plates. In a typical application to composite armor, an aluminum alloy screen would be bonded to titanium plates.

Fig. 4.4. Screen-bonded armor, time of impact.

Fig. 4.5. Screen-bonded armor, ceramic fractured.

Fig. 4.6. Screen-bonded armor, back-up plate deflecting.

Fig. 4.7. Screen-bonded armor, back-up plate decoupled.

Case No.: **IL-5228**

Title: **EPOXY VACUUM CHAMBER**

Inventors: George E. Vogtlin, Carl J. Anderson, Angus L. Hunt, and John J. Murphy

Vacuum chambers in plasma compression apparatus are usually made of glass, plastic, or some other electrically nonconducting material. This is because the rapidly changing magnetic field would induce large eddy currents in a metal vacuum chamber, thus causing power loss and distortion of the field. Although electrically nonconducting, glass chambers are mechanically fragile, and plastic chambers are difficult to maintain at the required vacuum pressures (less than 10^{-8} Torr).

This improvement is an epoxy/fiberglass vacuum chamber having a polyimide film bonded to its inner wall. The epoxy/fiberglass provides mechanical strength and the polyimide film acts as a diffusion barrier permitting the chamber to be maintained at pressures in the 10^{-9} Torr range. The vacuum chamber and its method of construction are described in the following example.

EXAMPLE

A collapsible wooden mandrel was constructed, 13 feet long and 17.5 inches in diameter. A 0.125-inch thick layer of Mylar* tape was wrapped around the mandrel to smooth out irregularities on the surface. A 5-mil Kapton 500 H* film was spirally wrapped over the Mylar. The overlap seam in the Kapton layer was glued with Furane 202* epoxy containing 10% diethylenetriamine (DTA) as curing agent. More Furane-DTA* was brushed on the Kapton and one layer of 9-ounce fiberglass "boat tape" was laid on in a spiral wrap. This was followed with another coating of Furane-DTA and an additional layer of fiberglass. The assembly was then cured in air for 3 - 4 hr at 80°F. The process was repeated until a final epoxy/fiberglass thickness of 0.5 inch was built up. The cylindrical wooden mandrel was then collapsed and removed to give an epoxy/fiberglass cylinder 13 feet long, 17.5 inches inside diameter, 0.5 inch wall thickness, with a 5-mil Kapton film bonded to its inner surface. The cylinder was installed in the "2X" plasma machine, pumped down to 10^{-5} Torr, and baked out at 180°F for 4 hours. The bakeout was repeated twice more. A titanium gettering pump was used to reduce the pressure to a final value of 1×10^{-9} Torr. The rate of pressure rise and concentration of residual gases was comparable to that in an all-glass vacuum chamber.

Reference to a company or product name does not imply approval or recommendation of the product by the University of California or the U.S. Atomic Energy Commission to the exclusion of others that may be suitable.

Case No.: IL-5267

Title: FIBER IMPREGNATION PROCESS

Inventors: Tsu T. Chiao and Richard L. Moore

Glass fiber impregnated with epoxy resin is often used to make "filament-wound" objects. Low resin viscosity is desirable to promote good wetting of the fiber; however, high-viscosity resins are sometimes needed to produce parts having specific desired properties such as flexibility and high impact strength. These two conflicting requirements (low vs high viscosity) are met in practice by taking an initially high-viscosity resin and heating it to reduce its viscosity before application to the fiber. Unfortunately, heating the resin introduces a new problem—the resin pot life is shortened drastically. Short pot life is so inconvenient that high-viscosity resins are often not considered for certain uses despite their otherwise desirable properties.

This development is an improved glass fiber/epoxy impregnation process which allows high-viscosity resin to be used, yet features good fiber wetting and normal resin pot life. As shown in Fig. 4.8, heated epoxy resin and curing agent are mixed on a spinning disk supported in a plastic funnel. This method of continuous mixing minimizes heat-generating shear forces which would shorten the pot life of the mixed resin. Also, the mass of mixed resin in the funnel at any time is fairly small. Epoxy/curing agent mixing causes an exothermic reaction; minimizing the mass improves heat dissipation and hence helps retain normal pot life.

The mixed resin is dripped onto the glass fiber as the latter passes over a heated, grooved wheel (like a pulley sheave). Both the resin and the fiber are heated at the point of resin application. This promotes good wetting of the fiber by locally reducing the viscosity of the resin. Another way of applying the resin to the fiber is shown in Fig. 4.9. The glass fiber passes over a heated, grooved wheel which just dips into a pan of mixed resin (high viscosity). The resin is heated locally at the point where the heated wheel dips into the bath. A small amount of heated resin (lowered viscosity) is held in the wheel groove and applied to the glass fiber. Because of the small contact area, heat transfer from the wheel to the bulk of the resin is very small, thereby minimizing any shortening of resin pot life.

As shown in Fig. 4.8, the epoxy-impregnated fiber passes around several heated pins contained within a vacuum chamber. The pins serve several functions: (1) they improve epoxy wetting of the fiber by mechanically separating the fiber strands; (2) they help remove air bubbles in the impregnated fiber; (3) resin content of the fiber can be controlled by varying the temperature of the pins; and (4) they apply an adjustable tension to the fiber. After emerging from the chamber through a vacuum seal, the impregnated fiber is wound directly onto a winding mandrel (not shown).

The present fiber impregnation process has two essential features which distinguish it from conventional fiber impregnation methods. First, the epoxy resin is mixed only as used, rather than in a batch. Second, the mixed resin is heated only at the point where it is applied to the fiber. It is the combination of these two features which

enables high-viscosity resin to be used while still providing good wetting of the fiber and normal resin pot life.

Fig. 4.8. Fiber impregnation process.

Fig. 4.9. Alternative way of applying heated resin.

Case No.: **IL-5258**
Title: **BURNISHING TOOL**
Inventor: **Carroll M. Gordon**

Machining tools typically produce a "scalloped" surface finish on the workpiece, i.e., a finish with regular ridges and valleys. This scalloped finish can be smoothed by grinding, lapping, or burnishing; however, each of these smoothing operations has its own limitations. Grinding removes material rapidly; the process is hard to control, especially on contoured surfaces. Lapping requires a congruent surface, severely limiting its application. Burnishing with conventional tools wipes and tears the surface, leaving it non-uniform and highly strained.

In this improvement the burnishing element is a free-rolling ball pneumatically pressed against the workpiece surface. The ball rolls or "irons" out the surface irregularities left by the machine tool. The free-rolling action eliminates wiping and tearing effects; the resulting surface finish is much smoother and more uniform than the original machined finish.

In its simplest form, the tool consists of a finishing ball held in a socket by air pressure. Fig. 4.10 shows an improved device in which a larger intermediate ball increases the total force applied to the finishing ball (the force is increased by the ratio of the ball surface areas). In addition to forcing the finishing ball against the workpiece, the pressurized air also serves as a lubricant for the balls in the same manner as an air bearing. Vents are provided in the ball socket to permit air to escape.

In operation, the burnishing tool is mounted in a lathe compound and brought to bear against the rotating workpiece. The finishing ball rolls across the workpiece parallel to the ridges left by the prior machining operation. In addition to the rolling motion, the ball also moves slightly in-and-out to match the changing contour of the rotating workpiece. This combination of rolling and translation motions allows the finishing ball to smooth the workpiece surface without the wiping and tearing actions of conventional burnishing tools.

Figure 4.11 shows a cross section of a machined surface before and after burnishing. Notice that the burnishing tool does not remove material from the workpiece. It merely redistributes material between the high and low spots to produce a smoother average finish. The degree of finish improvement depends on the roughness of the original machined surface, the workpiece material, and the air pressure behind the finishing ball. For example, a 40-AA machined aluminum finish was improved to 4-AA finish at 60 psi air pressure. An 80-AA machined steel finish was reduced to a 14-AA finish at 60 psi, and a 32-AA machined brass finish was improved to a 12-AA finish at 20 psi air pressure.

Fig. 4.10. Burnishing tool, pneumatic.

Fig. 4.11. Work surface profiles.

Case No.: **IL-5261**

Title: **FLUID CONTROL SYSTEM FOR MACHINING OPERATIONS**

Inventor: **Emmett W. Wrenn**

In some machining and spinning operations the workpiece is shaped by short, repeated passes of a tool over the work surface. This type of work is usually done by hand since automatic feed mechanisms are not suitable for such operations.

This fluid control system is designed to make short, repeated tool passes over a work surface. As shown in Fig. 4.12, tool motion is controlled by a piston 11 moving in a cylinder 12; the tool holder is fastened to rod 14. On the work stroke (left-to-right), compressed air is admitted to the left end of cylinder 12 via ports A and D of a 4-port switching valve 18. The right side of cylinder 12 is filled with a fluid which flows through a control valve 15 into a reservoir 16. The velocity of piston 11 to the right on the work stroke is determined by the fluid damping which is, in turn, controlled by the setting of valve 15. During the work stroke, air trapped in reservoir 16 is vented via ports B and C of switching valve 18. At the end of the work stroke, valve 18 is automatically switched so that compressed air is admitted to fluid reservoir 16 via ports B and D. At the same time, ports A and C are connected so that air in the left side of cylinder 12 is vented to atmosphere. The combination of high fluid pressure on the right side of the piston and atmospheric air pressure on the left side rapidly returns the piston to the left to complete the work cycle.

Figure 4.13 shows a detail view of fluid control valve 15. Flow rate is controlled by the area of an annular orifice 23 between ball 19 and valve seat 21. The position of the ball is varied by a needle 22 carried on the end of a threaded stem 25. The ball is retained in the valve body by a threaded insert 26 having a flow orifice of fixed size.

A wide variety of linear tool motions can be obtained by varying the input air pressure, the setting of valve 15, and the timing cycle of switching valve 18.

Fig. 4.12. Fluid control system layout.

Fig. 4.13. Fluid control metering valve.

Case No.: **IL-5312**

Title: **THORIUM SURFACE TREATMENT**

Inventors: Carlos A. Colmenares, Donald R. McKenzie, and Nino Pompilio

The corrosion of thorium by water vapor is summarized by the reaction: $\text{Th} + 2\text{H}_2\text{O} \rightarrow \text{ThO}_2 + 2\text{H}_2$. This technique is a surface treatment method for decreasing the corrosion rate. The cleaned thorium part is immersed for 5-10 minutes in a stirred solution (room temperature) of 50/50 vol. % $\text{HNO}_3/\text{H}_2\text{O}$ containing 1/2 to 1% HF. The part is then rinsed with distilled water for 10-15 minutes, followed by an alcohol rinse and drying in air (or argon).

Figure 4.14 shows the effect of this surface treatment on the water vapor corrosion rate of thorium. Treated and untreated thorium specimens were exposed to an atmosphere of 100 Torr H_2O and 660 Torr N_2 at 140°C . For comparison purposes, the effect of a 100% HF surface treatment is also shown. The graph indicates that the corrosion rate of thorium treated with 50% HNO_3 /1/2% HF is about 1/25 that of an untreated sample, and about 1/8 that of thorium treated with 100% HF.

Fig. 4.14. Corrosion of thorium exposed to H_2O vapor.

Case No.: **IL-5255**

Title: **METHOD FOR MAKING THIN BERYLLIUM SHAPES**

Inventor: Robert S. Juntz

When metals are vapor-deposited, the substrate is usually positioned perpendicular to the vapor beam. This maximizes the amount of metal deposited onto the substrate per unit area per unit time. If the substrate surface is inclined to the vapor beam, the deposit becomes thinner and more porous. The limiting case is reached when the substrate is parallel to the vapor beam; in this position, the thickness of the metal deposit is essentially zero.

The above effect can be used to make thin beryllium shapes. As shown in Fig. 4.15, an outline of the desired shape is scored onto the surface of a copper substrate. Beryllium is then vapor-deposited onto the scored substrate. Since the sides of the scored groove are at an angle (other than 90°) to the vapor beam, the beryllium which deposits in the groove is thinner and more porous than the deposit on the un-scored surface of the substrate. This effect is shown in the enlarged view of the score groove in Fig. 4.16. The copper substrate is then etched away in nitric acid to leave the beryllium deposit shown in Fig. 4.17. The desired beryllium shape can then be easily broken out along the scored groove where the deposit is weak and porous.

Beryllium shapes ranging in thickness from 0.001 to 0.007 inches have been produced using standard vapor-plating equipment and methods. Virtually any shape can be made. The method is particularly adapted to mass production of thin beryllium shapes, that is, a fairly large copper surface can be multiple-scored in cookie-cutter fashion. In addition, although the method has been described with reference to depositing beryllium on copper, the process is completely general and can be easily adapted to other metals.

Fig. 4.15. Beryllium vapor-deposition substrate, plan and enlarged cross section.

Fig. 4.16. Beryllium vapor-deposited plating.

Fig. 4.17. Vapor-deposited beryllium shape, cross section.

Case No.: **IL-5309**

Title: **METHOD FOR MAKING TITANIUM FOIL**

Inventors: **Edward D. Albrecht and Robert S. Juntz**

Titanium foil is usually made by rolling commercial titanium to the desired thickness. The repeated rolling operations contribute significantly to the high cost of the finished foil. In addition, commercial titanium is only 99.5% pure; the 0.5% impurity level in the finished foil is much too high for many research purposes.

This procedure is a method for making high-purity titanium foil simply and inexpensively. Titanium is vapor-plated to the desired thickness onto a beryllium substrate maintained at 900°C. As the plated substrate is cooled to room temperature, the titanium undergoes a phase transformation at 884°C. The accompanying density change breaks the Ti/Be bond and the titanium can be lifted as a foil from the substrate.

This method has been used with conventional electron-beam coating apparatus to make 0.8 mil titanium foil in 3-inch squares. Much larger foils could be made by plating titanium on a revolving beryllium drum or, possibly, a beryllium-plated steel drum. Typical titanium deposition rate is about 1 mil per minute.

A significant advantage of this method is that titanium foils of extremely high purity can be made. X-ray diffraction and electron microprobe analyses have shown no detectable contamination of either the titanium or beryllium surfaces. This probably means that the Ti/Be cross-diffusion and compound-formation rates are very low at the 900°C plating temperature. The method could possibly be extended to other metal/substrate pairs provided two conditions are met: (1) the plated metal goes through a phase transformation involving a large density change; and (2) there is minimal cross-diffusion or compound formation between metal and substrate at temperatures just above the phase transformation temperature.

Case No.: IL-5310

Title: METHOD FOR BONDING BeO TO TITANIUM

Inventors: Edward D. Albrecht and Robert S. Juntz

BeO has been investigated as a ceramic facing for composite armor in conjunction with a Ti backup plate. One of the problems encountered has been the lack of an adequate method for bonding BeO to Ti. Conventional brazing methods are not satisfactory because braze metals do not adhere to the BeO.

In this procedure Ti is vapor-plated onto a BeO substrate heated to 1300°C. The plated Ti reacts rapidly with the hot BeO to form a tightly adherent layer of mixed intermetallic compounds (TiBe_{12} , TiBe_2 , TiO_2 , etc.). The BeO is then slowly cooled to 500°C while still vapor-plating Ti. As the temperature is gradually reduced, the Ti reacts less rapidly with the BeO to form intermetallic compounds. This results in the formation of a graded transition layer consisting of intermetallics/Ti in a continuously decreasing ratio. Finally, at 500°C, only pure Ti is deposited. The coated BeO is then cooled to room temperature and the outer Ti layer is brazed to a Ti backup plate with an aluminum brazing foil. Alternatively, both Ti surfaces (the outer layer and the backup plate) can be "flushed" with copper and then brazed at a fairly low temperature. This reduces thermally-induced stress in the BeO.

The coating procedure described above can also be used to bond BeO to itself; an aluminum foil is used to braze together the Ti outer layers of two coated BeO pieces. A related method is useful for bonding TiBe_{12} (an excellent ceramic facing for armor) to a Ti backup plate. Ti is vapor-plated onto a TiBe_{12} substrate heated to 800°C. After a 1-mil coating of Ti is deposited (about 20 minutes), the plating shutter is closed and the coated TiBe_{12} is held at 800°C for an additional 40 minutes. A portion of the Ti coating diffuses into the TiBe_{12} to create a strong TiBe_{12} /Ti bond. The Ti "outer layer" of the coated TiBe_{12} is then brazed to a Ti backup plate with an aluminum brazing foil. It should be noted that the bonding mechanism is diffusion in this case, rather than the formation of intermetallic compounds as in the BeO/Ti bonding method described above.

The process for coating BeO in a conventional "electron beam coater" follows:

1. Load BeO into vacuum chamber (typical BeO piece is 3" x 3" x 1/4").
2. Pump chamber down to 10^{-6} Torr.
3. Heat BeO to 1300°C.
4. Heat Ti to 1800°C.
5. Open plating shutter to uncover Ti source.
6. Wait 5 minutes (this forms intermetallic layer on BeO substrate).
7. Cool BeO linearly from 1300°C to 500°C over 20 min period.
8. Close plating shutter and turn off electron beam.
9. Cool BeO to room temperature, remove from vacuum chamber.

Section 5. Measurement

Case No.: **IL-5256**

Title: **METHOD FOR MEASURING TRANSIENT CURRENTS**

Inventor: **Robert Parker**

This is a method for accurately measuring the total energy of a transient current pulse. The measurement can be made at a point remote from the source of the current pulse, and the method is responsive to all pulse energies above zero. As shown in the schematic drawing (Fig. 5.1), the current pulse is passed through a resistive sensor such as a platinum wire, thermistor, etc. The pulse source and the sensor are isolated from the recording system by a "zero-time isolator" which may consist of a fast-acting mechanical switch, an LC network, or any equivalent device which provides electrical isolation between the incoming pulse and the recording system. The passage of the current pulse heats the sensor very rapidly (several microseconds), thereby increasing its electrical resistance. After the current pulse and its associated electrical noise have ended (several milliseconds), the "isolator" connects the heated sensor with a constant current source. The recording system measures the voltage drop across the sensor caused by the passage of the constant current. This voltage drop is proportional to the total energy of the original current pulse.

For purposes of illustration, assume the sensor is a platinum wire 0.001 inch in diameter and 0.100 inch long. If a pulse energy of 0.5 mJ is "dumped" into this wire, its resistance increases from an initial value of 0.5 ohm to a final value of 0.77 ohm. If the heated wire is "interrogated" with a constant current of 45 mA, the increase in voltage drop (caused by the increase in resistance) is 12 mV. This corresponds to an output signal of 24 mV per mJ of input energy to a wire. One of the advantages of the method is that the wire can be interrogated several milliseconds after the end of the input pulse. The pulse and its associated electrical noise occur within microseconds. Thus, the measurement of wire resistance several milliseconds after the pulse essentially eliminates noise interference. A delay of several milliseconds in measuring the wire's resistance does not introduce any significant error due to cooling. This is established by calculations which show that, at 6.25 msec after the pulse, the center of the wire is still a 95% of its peak temperature. Calculations also show that heating of the wire caused by the constant interrogation current is negligible compared to heating caused by the input pulse.

Fig. 5.1. Transient current measurement schematic.

Case No.: **IT-5232**

Title: **METHOD FOR MEASURING OPTICAL WEDGE ANGLE**

Inventors: **Gilbert W. Leppelmeier and Donald J. Mullenhoff**

An "optical flat" is a transparent disk having two plane parallel surfaces. The deviation from parallelism is called the "wedge angle." It is desirable to have a quick and inexpensive method of measuring wedge angle so that the relative quality of different optical flats can be determined. Figure 5.2 shows a simple optical setup for measuring wedge angle. The components are first arranged as shown, but without the negative lens. The flat, preferably mounted in a gimbal, is manipulated until the light from the laser is reflected back on itself, thus forming a spot of light on the screen. The position of the spot is marked. The negative lens is then inserted between the laser and the beam splitter (the alignment of the lens is not critical as long as the screen is illuminated by light reflected from the flat). The insertion of the negative lens into the system produces an approximately circular interference pattern on the screen as shown in Fig. 5.3 (the interference pattern is caused by rays reflected from the front and back surfaces of the flat). The center of this circular pattern is displaced from the previously marked light spot by a distance, Δx , which is proportional to the wedge angle of the flat (if the faces of the flat were perfectly parallel, the center of the interference pattern would coincide with the marked spot). The value of the wedge angle, α , is calculated from the following relation:

$$\alpha = \frac{h \Delta x}{n^2 S_1 (S_1 + S_2 + S_3)}$$

where:

α = wedge angle (radians)

h = thickness of optical flat (cm) in Fig. 5.4

Δx = displacement of pattern from marked spot (cm)

n = index of refraction of flat

S_1, S_2, S_3 = various distances marked on Fig. 5.2 (cm)

The accuracy of the method is about ± 0.15 arc second of wedge angle. The concentricity of the circular interference pattern is a qualitative measure of the relative flatness of the two reflecting surfaces. The method can also be used to align resonant reflectors.

Fig. 5.2. Layout for measuring optical flat's wedge angle.

Fig. 5.3. Fringe shift proportional to wedge angle.

Fig. 5.4. Defining wedge angle.

Case No.: IL-5293

Title: INTERFEROMETRIC METHOD FOR MEASURING
PARALLELISM OF LASER ROD FACES

Inventor: E. Douglas Baird

Laser rod faces must be parallel to very close tolerances. Conventional optical measuring methods are not sufficiently precise where the required angular tolerance between the two faces is less than 10 seconds of arc.

This procedure is an interferometric method for directly measuring a small angle between any two spaced-apart surfaces, where the wavelength of the light serves as the determining reference. Measurements to an uncertainty of one fringe (less than 1 second of arc) are possible.

To measure the parallelism of laser rod faces, a laser, two mirrors M_1 and M_2 , an aperture, an interference screen or camera, and a beam splitter are set up as shown in Fig. 5.5. The mirrors M_1 and M_2 are independently adjusted such that the reflected image of each face C and D of the laser rod are centered in the aperture. One of the mirrors M_1 or M_2 is then adjusted to give a zero order light interference pattern on the screen. The screen should appear either all white or all black, without any interference fringes. The beams reflected from the laser rod faces are thus collinear, and the optical path length of each beam is an integral number of half-wavelengths. A mirror M_3 is then inserted in front of face D of the laser rod and adjusted to give a zero order interference pattern on the screen. The mirror M_3 is thus parallel to the plane of the face D of the laser rod. Then, without disturbing the mirrors, the laser rod is rotated end-for-end and adjusted to give a zero order interference pattern at the screen. The face D of the laser rod is then parallel to the plane which face C of the laser rod was in prior to the rotation of the rod. Care must be taken in this step to rotate the laser rod only about the axis through its center perpendicular to the plane of the figure. The mirror M_3 is then removed and the number N of fringes resulting are counted for a length L.

The number N of fringes over the length L is directly related to the angular relationship α between faces C and D of the laser rod by the following equation:

$$\alpha = \frac{N\lambda}{2L}$$

where λ is the laser wavelength. The above relationship is an approximation, and applies only where the angular relationship α between the two surfaces is very small. The relationship is derived by assuming that light is reflected from two surfaces of a very thin wedge. The wedge angle in this case would equal twice the angular relationship α between the two rod faces, due to the manner in which the rod is rotated.

Fig. 5.5. Layout for measuring laser rod face parallelism.

Case No.: IL-5342

Title: IMPROVED REFLECTOMETER

Inventors: Ton; Huen, Geoffrey B. Irani, and Frederick O. Wooten

Reflectometers measure the proportion of an incident light beam reflected from the surface of a sample. In one application the measured ratio is used to compute the complex dielectric function $\epsilon(\omega)$, useful in studying the sample's electronic structure.

Conventional reflectometers typically employ two light beams; one reflects off the sample, the other is a reference. Although such reflectometers operate quite satisfactorily at long wavelengths, in the vacuum ultraviolet spectral region they pose major problems. It is extremely difficult to match the optical components in the two light beam paths, reflections from mirrors cause large losses in beam intensity and it is hard to maintain an atomically clean sample surface in an ultrahigh vacuum environment.

Figure 5.6 illustrates the basic components of the present improved reflectometer. An ultrahigh vacuum chamber 2 houses a rotatable ellipsoidal mirror 4, a sample 6, and a photomultiplier detector 8. A monochromator 10 produces a diverging ultraviolet output beam 12, which is directed, via a suitable window 14, toward the mirror 4 about an axis coinciding with the rotational axis of the mirror. When facing the sample 6, the mirror 4 focuses the diverging beam 12 to produce a nearly normal incident beam with small cross section at the sample 6. Upon reflecting from the sample 6 at a small angle (less than 3°), the beam 12 travels to the detector 8, which is symmetrically disposed on the opposite side of the rotational axis of the mirror 4. When facing the detector 8, the mirror 4 focuses the beam 12 directly on the detector. Symmetrical placement of the sample 6 and the detector 8 on either side of the center line through the rotational axis of the mirror 4 and at distances therefrom equal to one-half of the ellipsoidal mirror focal length allows illumination of the same photocathode area of the photomultiplier detector 8 for the two positions of the mirror at which reflection of the beam to the sample and detector occurs. This has the effect of reducing errors caused by spatial variations of photocathode sensitivity.

In operation, continuous rotation of the mirror 4 produces a sequence of pulses at the detector 8 output representative of the intensity, ΔI , of the beam reflected by sample 6 and the intensity, I , of the beam directly focused on the detector. A ratio-meter circuit 16 receiving the output of detector 8 produces an output representative of the ratio $\frac{\Delta I}{I}$, which is displayed as a function of the beam wavelength λ by a conventional X-Y recorder 18.

Fig. 5.6. Reflectometer schematic diagram.

Case No.: **IL-5271**

Title: **LOW-PRESSURE PLANE-WAVE EXPLOSIVE LENS**

Inventor: **Paul B. Archibald**

The effects of shock waves on material properties can be conveniently studied if the incident shock has a planar configuration. In general, there are two methods for producing such planar shocks. One method employs the planar impact of a flat-faced projectile on a flat-faced specimen. This method is quite versatile since it can generate shock pressures (stress amplitudes in the specimen) over the wide range of a few to several hundred kilobars (kbar). However, complex equipment and precise control are required to produce a truly planar shock wave. In the second method, a flat-faced specimen is placed in contact with an explosive plane-wave generator, i.e., an explosive "lens" designed to produce a plane detonation front. This method has the advantage of simplicity, but it is limited to the production of shock pressures greater than about 100 kbar. Shock pressures less than 100 kbar can be obtained, but this requires the use of acoustically mis-matched elements to attenuate the shock, thereby greatly increasing the amount of explosive required to load the lens.

This device is an explosive lens which generates planar shocks in the 1-10 kbar range. These relatively low-pressure shocks are obtained by loading the lens with pressed PETN as the "fast" explosive component and a low-density powdered mixture of PETN and Cab-O-Sil (colloidal silica) as the "slow" explosive component. Instead of the conventional conical geometry, the pressed PETN in the present lens is shaped like a watch crystal, i.e., a spherical segment thickest in the middle and tapering out toward the edge. It was found that this shape for the fast explosive produced a more planar shock wave over a larger area than the standard conical shape.

As shown in Fig. 5.7, the present lens consists of a thin-walled container 11 having an interior spherical partition 12. A pressed spherical segment 13 of fast explosive (PETN) is held in place against partition 12 by a thin insert 14. The thickness of explosive segment 13 is not uniform, but is greatest at the pole and decreases out toward the edge. The interior of container 11 is filled with a powdered slow explosive 15 (Cab-O-Sil/PETN mixture) which is held in place by a cover plate 16. A detonator 17 is centrally mounted in a back plate 18 and the lens face is protected by a face plate 19.

Figure 5.8 shows the lens being assembled. Container 11 is inverted and the pressed segment of fast explosive 13 is positioned between partition 12 and insert 14. The slow explosive powder 15 is poured into the container and cover plate 16 is pressed into place with the aid of assembly fixtures 20 and 21. After placing the detonator 17 and attaching the face plate 19, the lens is ready for use.

Fig. 5.7. Low pressure plane wave lens.

Fig. 5.8. Low pressure plane wave lens assembly fixture.

Case No.: IL-5207

Title: AN MHD SHOCK DETECTOR

Inventor: Elmer White

Measurement of very large pressure shocks is difficult because the shocks usually destroy the measuring means before a meaningful signal is generated.

This apparatus is a magnetohydrodynamic (MHD) device for measuring large pressure shocks. The device consists of a longitudinal chamber 11, shown in Fig. 5.9. The front face 12 of the chamber is composed of a material having the same shock impedance as the medium in which the device is emplaced. The chamber is filled with a low density, easily ionized gas 14. Two permanent magnets 15, near the back end of the chamber, create a magnetic field perpendicular to the axis of chamber 11 for separating and directing particles of opposite polarity onto two charge-collecting electrode plates 16 which are insulated from the chamber. Electronic sensing means 18 measure the electrical potential difference across the plates 16. The device is oriented such that the pressure shock impinges perpendicularly on the front face 12 of the chamber.

In operation, the impinging pressure shock is transmitted through the front face of the chamber with little or no attenuation. The energy of the transmitted pressure shock creates in the gas an ionized shock front which propagates down the chamber. Due to the lower density of the gas, the ionized shock front propagates much faster than the pressure shock in the medium. Accordingly, the ionized shock front will reach the back end of the chamber substantially before the pressure shock. The magnetic field separates the ionized gas molecules and electrons, directing them to the charge-collecting electrodes 16. The electrodes 16 collect the separated charged particles of opposite polarity, generating a potential difference (voltage) signal which is measured by the electronic sensing means 18. Appropriate choices of chamber length and gas density make it possible to record the entire voltage signal before the pressure shock reaches the back end of the chamber and destroys the device.

The magnitude of the voltage pulse generated by the MHD shock detector can be related to the magnitude of the impinging shock. Specifically, the amount of energy transferred from the pressure shock to the gas 14 can be theoretically calculated. The magnitude of the pressure shock in the gas is calculated from the degree of ionization which is directly measured by the MHD device.

Fig. 5.9. MHD shock detector.

Case No.: **IL-5217**

Title: **SHOCK RECORDER**

Inventor: **Jasper J. Morrison**

The AEC uses various types of shipping containers for transporting ultra-hazardous materials including explosives, isotopes and the like. In transporting such materials it is often desirable to ascertain, for purposes of safety and security, the various forces which the container has experienced during transportation. The present development is an inexpensive apparatus for recording the acceleration experienced by such a container integrated over a long period of time.

As shown in the drawing, Fig. 5, 10, an accelerometer, such as a piezoelectric crystal, provides an accelerometer input at the point 11 corresponding to a shock experienced by the system. An operational amplifier 12 with capacitive feedback 13 operates as a so-called charge amplifier to put out a signal having combined voltage and current levels proportional to the accelerometer input. A series of diodes, 14a, 14b, 14c, 14d and a series of constant current diodes 15a, 15b, 15c, 15d operate to provide signal threshold levels, from low to high, which correspond to the level of the accelerometer input. For example, a relatively low level signal exceeding a first threshold would be gated through only the diodes 14a, 15a; a second higher level signal would be gated through both 14a, 15a and 14b, 15b. A third, still higher, intermediate signal would be gated through the diodes 14a, 14b, 14c and 15a, 15b, 15c, and so on. Each time a signal is gated through a corresponding pair of diodes, a current flows in one of the corresponding cells 16a, 16b, 16c, 16d, causing material to be plated from one side of the cell to the other. The amount of material plated on each cell is representative of the time current sum, and hence a resolution of the shock spectrum experienced by the accelerometer is available.

To determine the cumulative shock level to which the apparatus was exposed, a reverse current is applied to each of the cells 16a, 16b, 16c, 16d, at the level of each constant current diode 15a, 15b, 15c, 15d, and the time is measured to complete the reverse depositing process. In this manner, the shock integral at each threshold level is available. Once each cell 16a, 16b, 16c, 16d is read, the cells are available for re-use.

Fig. 5.10. Multi-threshold integrating shock recorder.

Case No.: IL-5219

Title: PEAK ACCELERATION SHOCK RECORDER

Inventor: Jasper J. Morrison

Previously, shock amplitudes in the range of 5 to 20 kilo g's have been measured by shock wave and schlieren techniques. Recently, telemetry techniques have been used in high acceleration projectiles. These prior techniques are cumbersome and difficult at best.

This device is a peak acceleration shock recorder which may be packaged in a miniaturized container occupying but a few cubic inches. The recording system is mounted on an object which is to undergo a kilo g acceleration, and is retrieved afterward. The present system obviates the bulky telemetry and schlieren techniques of the prior art.

As shown in the drawing, Fig. 5.11, an accelerometer input 11, e.g., from a piezoelectric crystal, feeds an operational amplifier 12 with a capacitive feedback loop 13 arranged in the amplifier configuration known as a charge amplifier. The current and voltage output from the charge amplifier are proportional to the amplitude of the shock pulse received from the accelerometer input. This signal is gated through the field effect transistor 15 and is sent to the comparator 17. The comparator sends a positive pulse to the monostable multivibrator 18, which sets the multivibrator into a high state, permitting a current to be transmitted to the electrolytic cell 19, thereby causing material to be plated from one electrode to the other. As the high side of the multivibrator transmits current to the cell, the low side of the multivibrator drains current from the capacitor 20. This in turn partially discharges the capacitor 14. Since the capacity of the capacitor 14 is approximately a thousand times greater than the capacitor 20, the amount of charge removed from the capacitor 14 is relatively small. The multivibrator 18 remains in the high state until the RC circuit 21 causes the multivibrator 18 to lapse back to its quiescent condition. At this point, if a sufficient charge on capacitor 14 still exists, the comparator 17 will continue to conduct, once again urging the multivibrator 18 to assume the high state, and once again resume plating the cell 19. With a typical kilo g shock pulse, the multivibrator 18 will oscillate between high and quiescent states dozens of times, until the discharging of the capacitor 20 depletes sufficient charge from the capacitor 14 to close the gate 16.

Since the current output of the multivibrator 18 into the cell 19 is at a constant level, the shock amplitude is relatively easy to read by reversing the plating which occurs in the cell 19. To determine the peak shock amplitude, a reverse current is run through cell 19 until the cell is returned to a neutral configuration. The time taken to return the cell to the neutral configuration, at the same current level as put out by the multivibrator 18, is directly proportional to the accelerometer shock signal.

Fig. 5.11. Peak acceleration shock recorder.

Case No.: **IL-5262**

Title: **SPARK-GAP FOIL SWITCHES**

Inventors: **Gary W. Carter and Kenneth D. Poor**

Spark-gap switches are widely used for measuring the detonation velocity of explosives. The most common type is the so-called "pin switch" which consists of two coaxial conductors formed into a cylindrical stub about 1/8 inch in diameter and 1 inch long. To measure the detonation velocity of an explosive, several of these pin switches are fixed to the surface of an explosive charge at precisely measured distances from the center of detonation. When the detonation front passes a given switch, the ionized products of detonation create a momentary short circuit between the two conductors. The occurrence time (counting from the instant of detonation) of the resulting current pulse is recorded. Since the distance between successive switches is known, the detonation velocity can be calculated. Radial symmetry of the detonation front is measured by placing several pin switches around a circle centered on the explosive detonator. The simultaneity of the switch pulses is a measure of the detonation front's radial symmetry.

The pin switch method has the disadvantage that each switch must be placed individually at a precisely-measured distance from a reference point (usually the explosive detonator). Since the overall accuracy of the method depends on the individual accuracy of the distance measurements, placement of the pin switches is a very time-consuming procedure.

This method employs a spark-gap "foil" switch that is made from copper-clad polyester film by a photographic etching process similar to that used in making printed circuits. The great advantage of the foil switch over conventional pin switches is that a complex pattern of several switches can be made as a single unit with precisely determined distances between the individual spark gaps. Thus, once the first spark gap is positioned at a known distance from a reference point, the distances of the other spark gaps from the reference point are automatically and accurately fixed.

Foil switches in various configurations have been made from 1.4 mil copper foil laminated to a 1.5 mil Mylar⁺ film. The copper is first coated with a photosensitive resist, and the desired switch pattern (drawn on an art master) is photographed onto the coated copper by exposure to ultraviolet light. The exposed portion of the copper foil (not masked by the switch pattern) is subsequently etched away. This method of making spark-gap foil switches is essentially identical with standard methods for making printed circuits.

A multiple switch consisting of 4 separate spark gaps is shown in Fig. 5.12 (all dimensions are in inches). The copper conductor on the far right is connected to a

⁺Reference to a company or product name does not imply approval or recommendation of the product by the University of California or the U. S. Atomic Energy Commission to the exclusion of others that may be suitable.

-150 volt DC supply. The other four conductors are grounded. A detonation front proceeding down the long axis of the switch will trigger each of the four spark gaps in turn. The four resulting signals are recorded by external electronics (not shown). As indicated in Fig. 5.12, the distances between the individual spark gaps are known to an accuracy of ± 0.001 inch. Therefore, it is only necessary to position the first spark gap (relative to the center of detonation) with an accuracy of ± 0.001 inch to automatically insure the same position accuracy for the remaining three spark gaps.

A radial switch pattern having two individual spark gaps is shown in Fig. 5.13. As mentioned previously, this switch configuration is used to measure the radial symmetry of a detonation front. Figure 5.13 shows the two spark gaps positioned at a distance of 0.750 inch from the detonation center. Radial symmetry measurements made with switches of this type have yielded a signal simultaneity of $\pm 1 \times 10^{-8}$ second. For comparison purposes, a typical simultaneity figure for conventional pin switches is $\pm 25 \times 10^{-8}$ second. The improved performance of the foil switches is due solely to the increased accuracy of spark gap placement.

Various commercial adhesives can be used to bond the foil switches to the surface of an explosive charge. Adiprene L-100 has been found particularly suitable for this purpose. The adhesive is placed onto the copper surface of the switch, filling the void between the spark gap and the parallel copper conductors. The excess adhesive is removed from the raised portion of the copper by gently scraping with plastic shim stock material. The adhesive bonding layer between the spark-gap tip and the explosive material should be of minimum thickness in order to place the switch as close as possible to the explosive surface. A piece of transparent tape is then attached to the polyester side of the switch and the unit is placed copper-side down upon the surface of the explosive. The tape retards movement of the switch until the adhesive is cured (12 to 24 hours). In many cases, the tape alone (no adhesive) may be sufficient to fix the switch to the surface of the explosive.

Fig. 5.12. Spark-gap foil switches, linear array.

Fig. 5. 13. Spark-gap foil switches, radial array.

Case No.: **IL-530**
 Title: **PRESSURE SENSITIVE SWITCH**
 Inventor: **James W. Lyle**

This device is a pressure-sensitive switch having a closure time of less than 1 nanosecond. This extremely fast response time makes the switch useful for the measurement of rapid phenomena such as the passage of shock waves through a test specimen.

Referring to Fig. 5.14, the switch comprises a thin strip 2 of vacuum-evaporated selenium bridging a pair of parallel leads 4 and 6 such as thin strips of gold. The selenium strip 2 and part of the leads 4 and 6 are encapsulated in a dielectric, such as mica, to form a thin flexible wafer 8. For a typical switch the width W is approximately 1 mm; the length L is approximately 5 mm; and the thickness of the selenium strip and the leads within the dielectric encapsulation is approximately 5000 Angstroms.

Operation of the switch is based on the abrupt (less than 1 nanosecond) change of the electrical resistance of selenium when it is subjected to an increasing pressure as graphically shown in Fig. 5.15. The abrupt resistance change is believed to be due to a lowering of the energy bandgap separating valence electrons and conduction electrons at and above a particular pressure characteristic of selenium (approximately 120 kilobars).

The switch is particularly useful in measuring travel time of shock waves through a medium. Referring to Fig. 5.16, two switches are placed against opposite surfaces of a test medium 10 subjected to the output of a shock wave generator 12. Each switch closes when the magnitude of the arriving shock wave equals or exceeds 120 kilobars. The difference in time between closure of the two switches is then equal to the shock wave travel time through the medium, and may be measured by a suitable circuit 14, such as a voltage source and oscilloscope, connected to the leads of the two switches.

Fig. 5.14. Pressure sensitive vacuum-evaporated selenium switch.

Fig. 5.15. Resistance characteristic of selenium vs pressure.

Fig. 5.16. Shock wave velocity measurement.

Case No.: **IL-5291**

Title: **PYROELECTRIC X-RAY DETECTOR**

Inventor: **Arthur Toor**

Certain crystals lack a center of symmetry and have a single unique polar axis. Barium titanate and lead zirconate-lead titanate are typical examples. Such crystals, when simultaneously subjected to an internal electric field and cooled through their Curie point, exhibit a permanent polarization along their polar axis. Crystals with such permanent polarization have so-called pyroelectric properties, i.e., changes in crystal temperature cause changes in the polarization charges located in certain regions of the crystal. These changes in the polarization charges produce electrical signals which are uniquely related to the associated temperature changes. The present development is an x-ray detector which employs the pyroelectric effect.

Figure 5.17 illustrates the principal components of the present device. On a disk-shaped pyroelectric crystal 2, thin conductive electrodes 4' and 6 are formed by vacuum deposition on the two flat crystal surfaces. The permanent polarization of the pyroelectric crystal is directed along the axis of symmetry 8 and is represented by the polarization vector 10. A conventional pulse-measuring device 12 is connected across the electrodes via leads 14 and 16, respectively. One of the two flat crystal surfaces faces a beam of x-rays 18 to be measured.

In operation, impinging x-rays deposit their energy in the crystal, raising the crystal temperature. The change in temperature causes a change in the polarization charges in certain regions of the crystal. The change in polarization charges, in turn, produces across the electrodes a time-varying voltage pulse which is measured by the pulse-measuring device. The measured voltage pulse across the electrodes during a given time interval is proportional to the total energy deposited by the incident x-rays during the time interval, provided that the time characterizing the energy deposition is shorter than the thermal diffusion time through the crystal. If the x-rays have a narrow and known band of energies, the total x-ray flux can be determined from the measured voltage pulse. If the energies of the x-rays occupy a wide band, suitable filters may be employed to select x-rays having energies within a desired narrow band for input to the detector.

In comparison with conventional x-ray detectors, such as thermocouples, the pyroelectric detector offers: (1) Simplicity and low cost; (2) radiation hardness; (3) fast response time; (4) large dynamic range; and (5) extreme sensitivity.

Fig. 5,17. Pyroelectric x-ray detector.

Case No.: **IL-5331**

Title: **X-RAY CALORIMETER**

Inventor: **Robert Parker**

The fluence (energy per unit area) of pulsed x rays is usually measured by placing a thin metal foil in the beam. The resulting temperature "spike" is recorded by a thermocouple attached to the foil. A disadvantage of this method is that the heat capacity of the thermocouple is comparable to that of the thin foil. As a result, a large portion of the incident x-ray energy goes into heating the thermocouple instead of the foil. The thermocouple output is therefore not an accurate measure of the x-ray fluence. This problem is partially solved in practice by applying correction factors to the thermocouple output. Since the correction factor is different for each foil/thermocouple pair, the foil method is tedious to apply and accuracy is poor.

This device is a passive x-ray calorimeter which does not require an attached thermocouple for readout purposes. The device consists of a metal foil coated with a temperature-sensitive paint (such paints are commercially available under the trade-name "Tempilaq").* In use, the coated foil is shielded so that the x-ray pulse irradiates only the central part of the foil. The resulting temperature "spike" causes the paint to melt outward from the center. The distance of the melt line from the foil center is a measure of the x-ray fluence.

Figure 5.18 shows the coated foil being irradiated by a source of pulsed x rays. The foil is shielded so that only the central portion (width = $2a$) is exposed to the x rays. Now suppose that at time t_0 an x-ray pulse deposits its energy into the exposed part of the foil. A plot of temperature vs distance at time t_0 is shown in Fig. 5.19. The temperature is given in terms of a dimensionless ratio T/T_0 , where T is the temperature at any point on the foil, and T_0 (do not confuse with t_0) is the temperature at the centerline of the foil. Distance is given in terms of a dimensionless ratio x/a , where x is the distance of any point on the foil from the centerline, and " a " is the half-width of the exposed portion (see Fig. 5.18). Figures 5.19-5.22 show the temperature profile of only the right half of the foil; the left profile is symmetrical about the centerline at $x = 0$. At time t_0 , just after energy deposition by the x-ray pulse, the temperature throughout the exposed part of the foil is uniform and equal to T_0 . This is because x-ray absorption by the foil acts as a uniformly-distributed heat source. At the boundary of the exposed foil ($x/a = 1$), the temperature drops abruptly to a relatively low value (room temperature) so that T/T_0 is essentially zero. The "step function" profile in Fig. 5.19 indicates that at time t_0 , energy has just been "dumped" into the exposed part of the foil, and heat flow by conduction into the unexposed part has not yet occurred.

*Reference to a company or product name does not imply approval or recommendation of the product by the University of California or the U. S. Atomic Energy Commission to the exclusion of others that may be suitable.

Figure 5.20 shows the temperature profile of the foil at a time t_1 which is later than t_0 . Sufficient time has elapsed for heat to flow by conduction into the unexposed part of the foil. Therefore, the original step-function profile is now broadened into a smooth curve. Note that the temperature of the foil's center is still essentially at T_0 ($T/T_0 \approx 1$). Figure 5.21 shows the foil temperature at still a later time t_2 . Here the temperature at the foil's center has dropped considerably, and points further out on the foil have been heated by conduction. Finally, Figure 5.22 shows the foil temperature at a time t_3 when the profile has flattened out almost to a horizontal line. This means that all points on the foil are at essentially the same temperature, heat flow along the foil having now reached a steady state.

Figure 5.23 is a composite of Figs. 5.19-22, with intermediate profile lines added to show more clearly how the temperature equalizes along the foil with time. The heavy line in Fig. 5.23, separating the shaded and non-shaded areas, is the envelope of the individual temperature profiles. This line represents the maximum temperature of any point on the foil, independent of time. Thus, the maximum temperature at $x/a = 1.5$ is $T = 0.33T_0$ ($T/T_0 = 0.33$). If either T or T_0 is known, then the other can be determined.

A specific example will clarify the use of Fig. 5.23 in calculating T_0 from an observed T . A rectangular nickel foil, 3 mils thick, was coated with a Tempilaq paint melting sharply at 50°C . A 6.5-inch zone in the center of the coated foil ($2a = 0.5$ inch) was then exposed to a radiation pulse. The Tempilaq melted out to a distance of 0.375 inch on either side of the foil's centerline, corresponding to $x/a = 0.375/0.25 = 1.5$.

Fig. 5.18. X-ray calorimeter.

Fig. 5.19. Foil temperature distribution at t_0 .

Fig. 5.20. Foil temperature distribution at $t_1 > t_0$.

Fig. 5.21. Foil temperature distribution at $t_2 > t_1$.

Fig. 5.22. Foil temperature distribution at $t_3 > t_2$.

Fig. 5.23. Composite temperature distribution showing temperature envelope.

From Fig. 5.23, $T/T_0 = 0.33$ for $x/a = 1.5$. Since $T = 50^\circ\text{C}$ (the known melting point of the Tempilaq paint), $T_0 = 50^\circ\text{C}/0.33 = 150^\circ\text{C}$. This calculated value for T_0 was confirmed by an independent measurement.

It should be noted that the curves in Figs. 5.19-5.23 are not just calibration curves valid for a particular foil. The curves are applicable to any foil, provided only that the geometry is rectangular. A different set of curves would be needed for a circular foil. Although such a circular foil could be used to measure x-ray fluence, it has been shown that a rectangular foil provides better sensitivity (the ability to measure small fluences).

Section 6. Plowshare Program

Case No.: **IL-5283**

Title: **REFRIGERATION SYSTEM FOR NUCLEAR DEVICE CANISTERS**

Inventor: **William M. Wells**

The emplacement of nuclear test devices underground subjects various heat-sensitive device components to temperatures up to several hundred °F. Some kind of refrigeration system must be provided to insure that these components will function properly when the device is fired. Heretofore, conventional closed-cycle refrigeration systems have been used to cool the device canister. Such closed-cycle systems require either an independent down-hole power supply or else power supplied by cable from ground surface. In addition, a fairly complex control mechanism is needed to maintain the canister temperature within desired limits.

This development is a simple absorber-type refrigeration system which operates without a power supply and does not require external control. The refrigerant is water vaporizing at 1 psi. The water vapor is absorbed by a hygroscopic compound. The system will operate in a high temperature environment (up to 500°F) to provide cooling to 100°F. Suitable absorber compounds include NaOH, CaO, and BaO.

Referring to Fig. 6.1, the absorber refrigeration system comprises large containers of H₂O 12 and dry NaOH 13 placed above the device in an instrument canister 14. A portion of the device canister serves as an evaporator 15. A simple liquid level control valve 16 regulates the water supply from the water container 12. A pressure regulating valve 17 on the line between the device canister and the sodium hydroxide container maintains the vapor pressure in the canister at 1 psi. The latent heat of vaporization of water at 100°F, 1 psi, is approximately 550 kilocalories per liter (1,000 BTU per cu ft). It is estimated that cooling can be provided for approximately 60 days for a vacuum-insulated device canister emplaced in an environment at 400°F.

Fig. 6.1. Refrigeration system for nuclear device canisters.

Case No.: **IL-5234**

Title: **LIGHT-TRIGGERED CAPACITOR DISCHARGE DEVICE**

Inventors: **Walter B. Lindquist and Clifford W. Walker**

Conventional blasting caps are usually fired by a low-current pulse furnished by a hand-cranked generator. However, the multiple detonator arrays used in LLL weapons experiments require a well-defined high-current pulse for successful firing. The firing pulse is provided by a capacitor discharge circuit which "dumps" its energy into the detonator array upon receiving an input trigger signal over a direct cable connection from the control point. A major problem with this system is that the input cable connection to the discharge circuit is vulnerable to electrical noise pickup from various sources. This can lead to spurious trigger signals to the discharge circuit with resulting accidental firing of the detonator array.

The present improved capacitor discharge circuit responds to a trigger signal in the form of a light pulse transmitted from the control point over a light pipe. The circuit is therefore immune to electrical noise pickup, with resulting security against accidental detonator firing. The light pipe itself will transmit only that light which enters the end of the pipe, i.e., the trigger light pulse. Environmental light is reflected from the surface of the pipe and is not transmitted.

Referring to Fig. 6.2, the principal components of the present device include a light source 2, a photodetector switch circuit 4, a light pipe 6 placed between the output of the light source 2 and a control input of the photodetector switch circuit 4, and a capacitor charge-discharge circuit 8 connected across the photodetector circuit 4. An external load 10, such as a detonator for activating an explosive device, is connected to the output of the capacitor charge/discharge circuit 8. In operation and with no output of light source 2, the capacitor charge/discharge circuit 8 is charged to a predetermined voltage. When the light source 2 is turned on, the output light beam is transmitted, via the light pipe 6, to the control input of the photodetector switch circuit 4. This activates the switch circuit 4 which, in turn, causes capacitor circuit 8 to discharge through the external load.

Figure 6.3 is a detailed circuit diagram of the capacitor charge/discharge circuit and the photodetector switch circuit. Before operation, terminals 12 and 14 are connected across the output of a suitable positive voltage source (not shown) to charge capacitor 16 in the capacitor charge/discharge circuit, via diode 18 and resistor 20, and capacitor 22 in the photodetector switch circuit, via voltage divider 24 and diode 26. After charging of the capacitors 16 and 22, the voltage source is disconnected. In operation, when a light beam through the light pipe impinges upon the photoresistor 28, the resistance of the photoresistor changes from a high value to a low value and the capacitor 22 discharges via resistor 30. The change in resistance of photoresistor 28 increases the potential on the grid 32 of a gas-filled switch tube 34, such as a krytron, to cause discharge of the capacitor 16 through the tube 34 and the external load 10, via the anode 36 and the cathode 38.

Fig. 6.2. Light-triggered capacitor discharge device.

Fig. 6.3. Capacitor charge/discharge circuit.

Case No.: **IL-5292**

Title: **REMOTE SWITCH POSITION SENSING DEVICE**

Inventor: **Rudolph W. Buntentbach**

There are many instances in which it would be desirable to continuously monitor the open-closed state of remote or inaccessible switches. A specific example is in underground testing of nuclear devices, where personnel at the ground surface must monitor the open-closed state of a multitude of switches emplaced several thousand feet underground. The present design is an oscillator circuit which converts the closure of a switch into a frequency that is uniquely identified with that particular switch. By observing the output frequency of the circuit, one can readily determine the open-closed state of a multiple switch array.

Referring to Fig. 6.4, a tuned-circuit oscillator is formed by a transistor 2, a pair of spaced inductors 4 and 6, a magnetic coupling circuit 8, and a load 10. The magnetic coupling circuit 8 includes an inductor 12, positioned between inductors 4 and 6, and n circuits connected in parallel across inductor 12, each comprising a capacitor C and a switch S . The capacitors in the n circuits have different capacitance values $C_1, C_2, \dots, C_n$. A conventional frequency detector 14 is connected across the load 10 to sense the oscillator frequency which, as shown below, is uniquely related to the closure of particular switches.

In operation, and with all switches open, transmission between inductors 4 and 6 by the magnetic coupling circuit 8 is weak and the feedback in the oscillator loop is too low to sustain oscillations. When one of the switches is closed, the associated capacitor is inserted in parallel with the inductor 12 and the magnetic coupling circuit establishes sufficient transmission between inductors 4 and 6 to sustain oscillations with a frequency dependent upon the capacitance of the inserted capacitor and the inductance of the inductor 12. Since the capacitors C have different capacitance values, the closure of each switch produces a unique oscillator frequency which is measured by frequency detector 14. Similarly, combinations of particular switch closures produce oscillator frequencies uniquely related thereto.

Figure 6.5 illustrates a second form of the present circuit. An electric motor 20 rotates a device 22 including a switch S , the open or closed position of which is to be detected. The shaft 24 of motor 20 is extended by means of a hollow nonmagnetic shaft 26. A winding or rotor coil 28 is placed on shaft 26 with leads 30 and 32 thereof extended through the hollow shaft and connected, via a capacitor C , across the switch S in the device 22. First and second static windings 34 and 36 are placed concentric with the shaft 26 on either side of the rotor coil 28 and are connected to a transistor 38 and a load 40 to form therewith a tuned-circuit oscillator similar to the oscillator circuit of Fig. 6.4. Operation is identical to that of the Fig. 6.4 circuit.

It should be noted that the open and closed positions of switch S in Fig. 6.5 may also be detected by connecting a signal generator across inductor 34 and by connecting a signal detector across inductor 36. When switch S is open, no transmission takes

place between inductors 34 and 36. When switch S is closed, the magnetic coupling circuit comprising inductor 28 establishes sufficient transmission between inductors 34 and 36 to couple the signal generator output to the signal detector.

The capability of the Fig. 6,5 configuration to transmit information as to switch positions in a rotating device to a static detecting device without the conventional use of slip rings is an important advantage, since, in many applications, the use of slip rings constitutes a safety hazard due to sparking.

Fig. 6.4. Remote switch position sensing circuit.

Fig. 6.5. Brushless position sensing on rotating switch.

Case No.: **IL-5204**

Title: **METHOD FOR IN SITU PRESSURE LEACHING**

Inventor: **Arthur E. Lewis**

Copper can be recovered from sulfide ores by placing the crushed ore in a pressure vessel, filling the vessel with water, and percolating oxygen at 500 psi through the ore/water mixture. Oxygen dissolves in the water and oxidizes FeS_2 (pyrite) in the ore to sulfuric acid and iron oxides. The resulting solution of dilute sulfuric acid plus dissolved oxygen leaches copper from the ore. The dissolved copper is then recovered electrolytically from the leach solution. The process is usually carried out at 100°C or higher to increase reaction rates.

This process is an ore leaching method which utilizes a nuclear chimney, created in an ore deposit located well below the water table, as an in situ pressure vessel for oxidizing the ore. Specifically, a nuclear device is emplaced and detonated in an ore deposit situated well below the water table. This creates a nuclear chimney, i.e., a region of shattered permeable ore, lying completely below the water table. At least two access holes are drilled to the chimney from the surface, one hole communicating with the top of the chimney, the other hole communicating with the chimney bottom (see Fig. 6.6). After letting the chimney fill with water, an oxidizing gas is introduced into the chimney at high pressure via the bottom access hole. The gas dissolves in the water and percolates through the chimney, thereby oxidizing the ore compounds. The desired element, e.g., Ag, Cu, Au, or Mo, is dissolved in the water. The leach solution is subsequently pumped to the surface and the desired element recovered by conventional methods.

It is essential to the present leaching method that the ore deposit is situated below the water table. Ore-bearing formations below the water table are saturated with water, thus creating a hydrostatic pressure head in the formation. Accordingly, a gas can be introduced into the chimney at very high pressure (more than 500 psi) and dissolved in the water in high enough concentration to oxidize the ore compounds.

In the case of copper ores, the oxidation reactions are sufficiently exothermic to raise the chimney temperature well above that of the surrounding formation temperature. This, of course, is desirable since it speeds up the oxidation of the ore. Heat can also be added to the chimney by preheating the oxidizing gas, or preheating the shattered ore with steam. Slower reaction rates may also be economically feasible due to the large amount of ore which can be processed in the chimney at one time. Thus the reaction can be carried out at the in situ formation temperature without additional process heat.

For ores of copper and similar elements, oxygen or air are suitable oxidizing gases. Chlorine is a suitable oxidizing gas for gold and molybdenum ores.

CHEMICAL MINING FLOODED PRIMARY SULFIDE ORE

Fig. 6.6. Chemical mining of flooded primary sulfide ore.

Case No.: **IL-5275**
Title: **STEMMING METHOD**
Inventor: **James L. Cramer**

Many PLOWSHARE applications require access to the postshot nuclear chimney. Reentry to the chimney is usually done by drilling out the hardened grout stemming material in the emplacement hole. Occasionally, an entirely new access hole is drilled to the chimney. Reentry by either method is time-consuming and expensive.

This is an improved method for stemming an emplacement hole so that it can be used for post-shot chimney access without drilling. As shown in Fig. 6.7, the annulus between emplacement casing and the hole wall is stemmed as usual. The bore of the emplacement casing, however, is stemmed only from the explosive canister to a point just below the minimum predicted chimney height. A series of valves or rupture disks are installed in the emplacement casing above the maximum predicted chimney height. The stemming in the emplacement casing shields the valves or rupture disks from the initial blast and, after collapse of the chimney, they are subjected only to the formation pressure. The valves are then opened electrically from the surface, or the rupture disks are broken by pressurizing the emplacement casing from the surface. Since, as shown in the drawing, the emplacement casing will be broken in the chimney region, opening of the valves or rupture disks establishes direct postshot communication between the chimney and the surface. Additional protection against venting can be provided by stemming the emplacement casing bore above or between the valves with an easily removed stemming fluid such as drilling mud. A high-pressure well-head assembly is installed preshot on the surface to guarantee containment.

Fig. 6.7. Quick-access emplacement stemming.

Section 7. Miscellaneous

Case No : **IL-5268**

Title: **SHAPED CHARGE WITH OXIDIZING LINER**

Inventor: **Paul B. Archibald**

A shaped charge is a block of explosive having a conical metal-lined cavity in one end. When the charge is detonated, the liner is collapsed inward and ejected as a high-speed jet of molten metal. The speed of the molten jet is higher than the sonic velocity in the target material. Consequently, target penetration is caused by hydrodynamic shock waves induced in the target by the jet. It is this mechanism that enables a charge-liner made of a chemically nonreactive, low-melting metal such as aluminum (mp = 660°C) to penetrate a steel target (mp = 1430°C).

As shown in Fig. 7.1, this device is a shaped charge having a liner made of an oxidizing material such as potassium perchlorate (KClO_4). When the charge is detonated, the resulting high-speed molten jet of KClO_4 reacts chemically (oxidizes) with the target material, thus providing a second, independent mechanism for target penetration in addition to shock wave penetration. To test this concept, the aluminum liner in a standard shaped charge (100 grams C-4 explosive; aluminum liner thickness = 0.0625 inch; cone angle = 90°) was replaced with a KClO_4 liner. Several charges having composite aluminum/ KClO_4 liners were also prepared on the theory that the Al-KClO_4 reaction is more energetic than the Fe(steel)-KClO_4 reaction. The test charges were fired against steel targets and the results compared with the standard aluminum-lined shaped charge. Tests results are given in Table 7.1. The data show that the volume of the target penetration hole is about 50% greater with the KClO_4 liner than with the standard aluminum liner. In the case of the 0.125-inch thick KClO_4 liner, a 25% increase in target penetration (hole depth) was obtained when compared to the aluminum liner. The composite aluminum/ KClO_4 liners were approximately equivalent to the standard aluminum liner in target penetration. Evidently, the aluminum and KClO_4 did not react with one another as expected. In addition to forming a larger hole than the standard aluminum-lined charge, the KClO_4 -lined charges did not form a metal slug to plug up the hole. The hole edges appeared more melted than eroded; this is consistent with the oxidizing effect of the KClO_4 liner on the target material.

Table 7.1. Penetration of steel targets with shaped charges having KClO_4 cones.

Control	Penetration (in.)	Volume (cc)
0.0625 Al cone	2-7/8	9
	2-5/8	10
	3-1/8	8
^a 0.067 KClO_4 cone	2-7/8	14.5
	3-3/8	15
	2-6/8	15
	2-6/8	13
	2-7/8	15
0.125 KClO_4 cone	3-1/2	13.5
	3-7/8	15
	3-1/8	12
0.250 KClO_4	2-5/16	13
	2-3/8	13
	2-1/4	13
0.125 KClO_4 + 0.065 Al	2-1/2	9.5
	2-3/8	9.5
	2-5/8	9.5
0.025 KClO_4 + 0.065 Al	2.0	9.1
	1-7/8	8.2
	2-1/8	10.2

Site 300 shaped charge with 100 gm C-4 explosive, modified by replacing 0.065 Al cone with KClO_4 cone. Cone angle: 90° standoff 7-3/8".

^aAt 0.067" thickness KClO_4 has the same mass as 0.0625 Al.

Fig. 7.1. Shaped charge with oxidizing liner.

Case No.: **IL-5277**

Title: **IMPROVED RANKINE CYCLE ENGINE**

Inventor: **Carl D. Henning**

The regenerative Rankine cycle has long been known, but seldom used because of its impracticality with conventional turbines. The heat transfer limitations and high moisture content in the vapor leaving the turbine are serious disadvantages.

The present device is a screw expander consisting of a pair of intermeshing helical screws having means for greatly improving the efficiency of Rankine cycle engines. As a preliminary, the conventional Rankine cycle is illustrated in Fig. 7.2. The elements include a boiler or burner receiving liquid from a pump and producing vapor transmitted to a prime mover. The prime mover includes a central motive element, and an expander to expand the vapor adiabatically. A condenser receives the exhaust vapor from the prime mover and converts it into a liquid. A pump between the condenser and the boiler moves the fluid back to the boiler at the boiler pressure. A temperature-versus-entropy graph for the system is shown in Fig. 7.3.

In the present expander, condensed vapor is directed through the hollow inside of the helical screws after having passed on the outside of the screw. This has the effect of cooling the screws and preheating the liquid entering the boiler in the path shown in Fig. 7.4. In Fig. 7.5, the line 4-5 represents the vapor flowing through the turbine, and it parallels the line 1-2-3 for the pumping process and counterflow in the expander. The areas 2-3-b-a-2 and 5-4-d-c-5 are equal, and represent the heat transferred to the liquid and from the vapor, respectively. The remaining heat addition process line 3-4 and heat rejection process line 5-1 occur at constant temperatures. Thus, the regenerative cycle has the same efficiency as the Carnot cycle which is the highest possible efficiency. An estimate of this efficiency is approximately 68% based on a working fluid which can operate at 1600°F. Preliminary calculations indicate that the screw wall temperature will be approximately 1200°F, a value which should not present metallurgical problems.

Figure 7.6 illustrates the manner in which helical screws intermesh in expanding or compressing vapor.

Fig. 7.2. Classic Rankine cycle flow diagram.

Fig. 7.3. Classic Rankine cycle entropy diagram.

Fig. 7.4. Screw expander flow diagram.

Fig. 7.5. Screw expander entropy diagram.

Fig. 7.6. Screw expander impellers.

Case No.: **IL-5303**

Title: **X-RAY MACHINE POWER MONITOR**

Inventors: **Melton H. Chew, Robert A. Lockhart, and Fiorello I. Danielli**

In many x-ray machine applications, such as x-ray diffraction experiments, adjustment of the x-ray target by a human operator is required. The present monitoring device was developed to give visual and audible warning signals to the operator when ac power is applied to an x-ray machine.

Referring to Fig. 7.7, an x-ray machine 2 produces a beam of x rays, directed toward a target 4 under test, when ac "on-off" switch 6 is turned "on." A section of flexible tubing 8 is connected at one end to a base 10. The other end of the tubing 8 houses a de-rated incandescent lamp 12, and is swung in the vicinity of the target 4. On the base 10 are mounted a buzzer 14 and a control circuit 16 having an input connected to receive ac power when the switch 6 is turned "on", and having a set of outputs connected to the lamp 12 and the buzzer 14.

Figure 7.8 schematically illustrates the control circuit 16. When ac power is applied to the x-ray machine 2, an ac-to-dc converter 18 activates a timing circuit 20. When activated, timing circuit 20 produces a steady stream of pulses which are selectively applied, via a manual switch 22, to lamp 12 or buzzer 14. The pulses applied to lamp 12 and buzzer 14, respectively, produce an "on-off" visual signal and an "on-off" audio signal, thereby alerting the operator to remove ac power to the x-ray machine 2 before making adjustments of target 4.

Further included in the control circuit 16 are a failsafe circuit (not shown) for automatically overriding switch 22 when the selected lamp 12 or buzzer 14 fails, and means (not shown) for manually checking overall operation of the device.

Fig. 7.7. X-ray machine power monitor.

Fig. 7.8. Power monitor control circuit.

Case No.: **IL-5304**

Title: **SPECIMEN CARRIER**

Inventor: **Charles E. Gilliam**

Electron microscope specimens (in the form of thin films) are transported in special holders furnished by the microscope manufacturer. These holders are open at both ends so the electron beam can pass through the specimen. The specimen is thus exposed to the atmosphere through the openings in the holder. This exposure is not a problem in most cases. However, in the case of specimens which react with oxygen or water vapor, an enclosure must be provided for transporting the specimens under an inert atmosphere or vacuum. There then remains the problem of transferring the specimen from the enclosure into the microscope without exposing it to the air.

This device is an enclosed carrier for transporting electron microscope specimens under an inert atmosphere. The carrier couples directly to the inlet column of an electron microscope, thereby avoiding exposure of the specimen to air when it is transferred into the microscope. As shown in Figs. 7,9 and 7,10, several specimen holders 12 are positioned on a turntable 14 enclosed within a gas-tight housing 16. A transfer tube 18 communicates between the housing interior and the microscope inlet via a transfer port 20. In the transportation mode, port 20 is closed by a sealing disk 22 on the end of a threaded stem 24. Turntable 14 is rotated by a yoke 26 which engages stem 24. The yoke is actuated by a rod 28 and an attached knob 30 (yoke 26 is not shown in Fig. 7,10 for clarity). Specimen holders are manipulated by a collet 32 on the end of a tubular rod 34 which is extendable through transfer tube 18. Collet 32 is actuated by a rod 36 inside tubular rod 34 (rod 36 is not visible in Fig. 7,9). The lower end of rod 36 is an Allen head which engages a corresponding Allen socket in the top of threaded stem 24. The collet is extended and retracted by means of a knob 38 fastened to the upper end of rod 36 and a spring 40 which normally retracts the collet into its sleeve. The lower end of spring 40 bears against a knob 42 which is fastened to the upper end of tubular rod 34. Rods 28 and 34 extend inside the housing through gas-tight seals 44 and 46, respectively. The gas pressure within housing 16 is indicated on a gauge 48.

The specimen holders are loaded into the carrier in an argon-filled glove box. The housing cover plate is removed and specimen holders 12 are press-fitted into the peripheral holes in turntable 14. The cover plate is then replaced and the turntable is rotated (by turning knob 30) until disk 22 is positioned above port 20. The socket in the top of stem 24 is engaged by the Allen head in the bottom of rod 36, and the stem is screwed down until disk 22 seats firmly against port 20. This seals off the interior of housing 16. The loaded specimen carrier is then removed from the glove box.

To transfer a specimen holder from the carrier to the microscope, transfer tube 18 is first coupled to the microscope's inlet column. The inlet column is flushed several times with argon to remove any water vapor which could react with the specimens. Stem 24 is then engaged by rod 36 and unscrewed slowly until gauge 48 indicates

a pressure rise (from the positive argon pressure in the microscope inlet column). The stem is then backed off completely until disk 22 contacts the recessed undersurface of turntable 14. This opens port 20 so that the interior of housing 16 communicates with the microscope inlet through transfer tube 18. After disengaging and withdrawing rod 36, turntable 14 is rotated until the desired specimen holder 12 is aligned with port 20. Collet 32 is extended by compressing spring 40 between knobs 38 and 42, and the extended collet is slipped over the specimen holder by advancing rod 34. Knob 38 is then released so that spring 40 retracts rod 36 within tubular rod 34. This retracts collet 32 partially into its sleeve and thereby firmly grips holder 12 in the collet. The specimen holder, now grasped in the collet, is passed through transfer tube 18 and into the microscope inlet (Fig. 7.9). When the specimen holder is properly positioned within the microscope, the collet is released and rod 34 is withdrawn. Turntable 14 is rotated to align disk 22 over port 20 and the housing interior is sealed off by screwing down stem 24. After the specimen has been examined in the microscope, the holder is removed from the microscope inlet by reversing the transfer procedure. A different specimen holder is then inserted into the microscope until all the specimens have been examined.

Fig. 7.9. Electron microscope specimen carrier.

Fig. 7.10. Specimen carrier details.

Case No.: **NY-5313**

Title: **MEMBRANE FILTER**

Inventors: **Carlos A. Colmenares and Earl R. McCann**

Frequently a vessel or enclosure containing friable radioactive material, such as PuO_2 , must be pumped down and maintained at high vacuum. A filter must be inserted in the line between the vessel and the vacuum system to avoid contaminating the vacuum pump with radioactive particulate matter. Conventional corrugated paper filters can't be used because the water held in the fibers would outgas into the vacuum system. Fritted ceramic filters are also unsuitable because their pore openings are too large, generally about 5 microns.

This device is a radioactive particulate filter in which the filter medium is a porous silver membrane. Such membranes are commercially available with a maximum pore diameter of only 0.45 micron. As shown in Figs. 7.11 and 7.12, the porous membrane is sandwiched between two fritted disks (stainless steel) for support against the pressure drop of the flow stream. The disk/membrane assembly is held in a retainer positioned within a housing.

The membrane filter has several advantages. First, and most important, the filter completely stops particles down to submicron size. In tests with PuO_2 particles in an air stream, there was no detectable alpha activity on the downstream side of the filter, whereas the upstream side was highly radioactive. Because of its all-metal construction, the filter can be baked out at temperatures up to 700°C . In addition, the filter can be designed for vacuum use or for high-pressure work up to 10,000 psi. When the membrane pores become filled with particles, the filter can be cleaned simply by filling the upstream side with detergent and then backflushing from the downstream side. The cleaning process is completed by a bakeout to remove water and any remaining detergent.

Fig. 7.11. High vacuum-compatible membrane filter.

Fig. 7.12. Membrane filter details.

Case No.: IL-5345

Title: TRANSPORTATION SYSTEM

Inventor: Paul B. Archibald

The idea of propelling a vehicle by compressed air has a long history; a pneumatically-propelled train was operated in Ireland as far back as 1840. All the pneumatic propulsion systems thus far proposed operate on the principle of expanding compressed air behind a piston in the same way that the expanding powder gases drive a bullet down a rifle barrel. The propelling force is provided by the static gas pressure acting on the cross-sectional area of the piston.

This design is a pneumatic propulsion system that operates on a fundamentally different principle than previous systems. The piston is propelled by a moving air stream rather than by expanding air volume. The moving stream generates a "velocity pressure" equal to $\rho V^2/2$, where ρ = air density, and V = air stream velocity. It is this velocity pressure acting on the piston which provides the propulsive force.

As shown in Fig. 7.13, a road vehicle is coupled to a strut which extends downward through a slot in the road surface. The lower end of the strut terminates in a magnetic clutch which nearly encircles the piston pipe. The piston itself can be either a magnet or else made of a ferromagnetic material. The piston pipe is constructed of plastic to permit magnetic coupling between the clutch and piston without formation of eddy currents. Air velocity in the piston pipe is maintained by a series of tangential jets which admit high-velocity air from a high-pressure air line connected to the output of a compressor (not shown). The lower pressure air in front of the piston is returned to the compressor intake via a low-pressure return line.

Figure 7.14 is an end view of the piston pipe and associated structures. Although the piston itself could be made as a solid cylinder, a hollow piston has the advantage of allowing the vehicle driver to "decouple" the vehicle from the air stream. Figure 7.14 shows a hollow piston having a butterfly valve mounted in one end. In normal vehicle operation the valve is closed, thus presenting the full-cross-sectional area of the piston to the air stream. The vehicle is fully coupled to the stream and hence moves at stream velocity. However, if the driver desires to slow down, he can partially open the valve to decrease the area presented to the air stream. Part of the stream is, in effect, bypassed, thus decreasing the net force acting on the piston. The driver can come to a gradual stop by fully opening the valve so as to completely decouple the vehicle from the air stream. The mechanism for controlling the valve is not shown; however, it could be done magnetically by a structure mounted on the clutch and actuated via a remote control in the vehicle.

The clutch is constructed in the form of hinged "clamshell" halves to permit the vehicle driver to select any of various auxiliary routes branching off the main piston pipe. Figure 7.14 shows a dotted outline of the right clutch half lifted away from the piston pipe. Thus the net magnetic force acting on the piston is directed toward the left, and the vehicle would enter a left-branching path at this point. The raising and

lowering of the clutch sections is controlled by the driver by a simple hydraulic system indicated schematically in Figs. 7.13 and 7.14. Auxiliary power for the hydraulic system is obtained from a pump geared to the vehicle wheels.

The piston and clutch ride on air bearings to minimize friction with the inner and outer walls of the piston pipe. Air for the piston bearing is obtained from the main stream via small passages which open out onto the rear piston face. The other ends of these passages open out into a shallow annular cutaway portion in the middle of the piston, or else into separate annular grooves spaced along the length of the piston. The high pressure air at the rear piston face enters the passages and emerges into the annular space between the piston and the inner pipe wall, thereby providing a film of high pressure air which prevents physical contact between the piston and the pipe. Similarly, the inner surface of the clutch is held away from the outer wall of the piston pipe by an air bearing. The high pressure air for the clutch bearing is supplied by a small air compressor located in the vehicle and driven off the vehicle wheels. Air is conveyed from the compressor to the clutch by a pressure hose mounted on the strut. The air bearing details have been omitted from the drawings for the sake of clarity.

As mentioned previously, the air velocity in the piston pipe is maintained by momentum transfer from the high-velocity air jets. In an operating system, these jets would be equipped with solenoid valves which would remain closed in the nonenergized state. As the vehicle passed by a particular jet, an electrical contact on the clutch would open the valve and admit a high-velocity air jet into the main air stream. This would accelerate the vehicle down to the next jet location, where the injection of high-velocity air would be repeated. The triggering of the jets could be made automatic or else under the control of the driver. In sections of the route where the vehicle was climbing a grade, the jets would be spaced closer together than in level sections.

Performance calculations for a transportation system of the type described indicate that vehicle acceleration equal to that of today's compact cars could easily be obtained. The cost of such a system (including cost of obtaining road right-of-way) comes to three cents per vehicle mile. The components required, such as plastic piping, compressors, etc., are all currently available. In addition, the system has the great advantages of being completely smog-free and essentially noiseless.

Fig. 7.13. Pneumatic drive for passenger vehicle.

Fig. 7.14. End view of piston and clutch assembly.