Electrical Engineering

Quals Questions

1997

Scores depended not so much on how much students already knew about a topic but rather on how well they were able to reason about the topic. Students I had to "lead" the whole way or who couldn't respond even with a lot of hints didn't do as well as other students. CS

The questions:

1) Consider two hosts on the Internet. They are workstations running an OS and some applications. One of the applications is a file transfer program. Let's say one host wants to transfer a file reliably to the other host. What hardware and software components participate in this transfer? What reliability problems could occur with each of these components, and how could you address them?

Answer: at the lower levels, packets could be corrupted or perhaps lost. At the routing level they could be lost or delivered out of order or perhaps even corrupted. You'd like a network protocol that uses checksums for detecting corruption, retransmissions for handling lost or corrupted packets, and that has sequence numbers for proper ordering. Basically any kind of handshake that accomplishes this was considered okay. At higher levels, packets must be delivered to the correct application. The OS participates with its handling of network buffers and file cache. If there's a bug in the OS, such as a pointer error, then the file could be corrupted even if the packets were successfully delivered. Thus, the ultimate way to make sure the file is reliably transfered is an end-to-end check. The application could send a length and checksum as well as the file, so the receiver can compare these.

2) [This one included a picture I drew.] Say there is a file server in a distributed file system. The file server is intended to be very robust, so it has two CPUs. The two CPUs have a fast interconnect, and they are both connected to the same dual-ported disks. One CPU is the "primary" CPU. It handles client requests to read/write files. The other CPU is idle. Its sole purpose is to take over for the primary CPU if the primary fails.

Assumptions you can make: 1) The primary fails by halting. 2) Don't worry about how client requests get to the correct CPU. (You can imagine that client requests get delivered to a queue that either CPU could draw from. Don't worry about coordination or synchronization with this queue.) 3) The primary conveniently only fails between idempotent file system requests, so you don't have to worry about any state in its memory when it fails.

Given these assumptions, what mechanism could the secondary CPU use to determine whether the primary has failed so it knows to take over? Given this mechanism, what are the performance and reliability concerns you can think of?

Answer: The interconnect between the CPUs allows the primary to send "heart beat" messages to the secondary. The secondary could also poll the primary. Students who tried to use the disks for this mechanism didn't do as well, since that would be so incredibly slow. Given the heart beat at intervals (which I then asked them to consider, even if they came up with the polling mechanism), one performance consideration is the interval of the heart beat. Too frequent a heartbeat could be a bit of a burden on the primary. Too large an interval between beats means that client requests wouldn't be handled for a long time if the primary failed just after a successful heart beat. The heart beat should be based on an interval and not on the granularity of client requests, as some suggested, because the secondary doesn't know whether there are just no requests or whether the primary failed. One reliability consideration is that heartbeats could be lost if there's a problem with the interconnect or such. In this case the secondary would become active even though the primary was still active. This could cause various kinds of harm.

OFFICE MEMORANDUM ◊ STAR LABORATORY

February 6, 1997

To: Diane Shankle

From: Tony Fraser-Smith

Electromagnetics

Subject: Ph.D. Quals Question, 1997

Question: Explain how you might send an electromagnetic signal down to the center of the earth from the surface. Ignore the possibility that the inner core of the earth consists of molten iron and assume the earth is wholly a conducting material of conductivity $\sigma = 0.01$ S/m. Other possibly useful information: The radius of the earth is 6370 km, and at 1 Hz the skin depth (δ) for an electromagnetic wave propagating in a medium with a conductivity of 0.01 S/m is 5 km.

Answer: There are many different ways to answer the above question. An ideal answer would include most or all of the following: (1) A brief discussion of good conductors and poor conductors, leading to the conclusion that the above problem must treat the earth as a good conductor ($\sigma/\omega\epsilon \gg 1$, where ω is the angular frequency and ϵ is the permittivity). (2) Some technical discussion in which the student must either remember or derive an expression for the attenuation of electromagnetic fields in a good conductor in terms of the attenuation constant (α) or the skin depth (δ), where $\delta = 1/\alpha = \boxed{2/(\omega\mu\sigma)}^{1/2}$, and where μ is the permeability. (3) The student should demonstrate some knowledge of how the wave is exponentially attenuated (with attenuation constant $\alpha = 1/\delta$).

Good conduct is prerequi to use skin depth

Using the above information, the ideal answer would then include (4) a scaling of the skin depth information given at 1 Hz to derive the frequency corresponding to a skin depth of around 6370 km. The answer obtained is typically $\boxed{\text{around } 10^{-6} \text{ Hz,}}$ which is essentially dc as far as most electrical engineering students are concerned. The students are then asked if there are any electric or magnetic fields EMERGING from great depths in the earth, which would be subject to the same attenuation. This should lead into a discussion of (5) the earth's magnetic field, which is very nearly steady, but not quite, and (6) how there can be some very long term variations, which are consistent with the above frequency estimate. Finally, the students are told that there are some shorter term variations with periods around 1 Hz, and asked to explain them in the context of our discussion. The final conclusion (7) is that the variations either originate in the earth very close to the surface or they reach the earth's surface from above and thus are not subject to the same attenuation as that which takes place in the earth.

$$\frac{\sigma}{\omega\epsilon} = \frac{0.01}{1.\ 8.8\times10^{-12}} \qquad \epsilon_0 = 8.8\times10^{-12}\ \text{F/m}$$

Joseph

Quals Questions
J.W. Goodman

Orthogonal functions play an extremely important role in signal analysis.

1. Under what conditions are the following two functions orthogonal?

$$g_1(t) = \mathrm{sinc}(t)$$
$$g_2(t) = \mathrm{sinc}(t - \tau)$$

Here, τ is a delay and $\mathrm{sinc}(t) = \dfrac{\sin \pi t}{\pi t}$.

Answer:

We assume that the region over which orthogonality is defined is the entire real line. In that case the condition for orthogonality is

$$\int_{-\infty}^{\infty} \mathrm{sinc}(t)\mathrm{sinc}(t - \tau)\, dt = 0$$

Attempts to evaluate this integral in the time domain yield little of use. However, the integral can be evaluated with the help of Parseval's theorem.

Let $\mathrm{rect}(f) = \begin{cases} 1 & |f| < 1/2 \\ 1/2 & |f| = 1/2. \\ 0 & |f| > 1/2 \end{cases}$

Then Parseval's theorem implies that

$$\int_{-\infty}^{\infty} \mathrm{sinc}(t)\mathrm{sinc}(t - \tau)\, dt = \int_{-\infty}^{\infty} \mathrm{rect}(f)\mathrm{rect}(f) e^{j2\pi f\tau} df = \int_{-\infty}^{\infty} \mathrm{rect}(f)\, e^{j2\pi f\tau} df = \mathrm{sinc}(\tau).$$

We see that orthogonality is obtained when the sinc function is zero, or whenever the delay is a non-zero integer.

2. We wish to expand a function $g(t)$ in a series of orthogonal functions of the type discussed above:

$$g(t) = \sum_{n=-\infty}^{\infty} a_n\, \mathrm{sinc}(t - n).$$

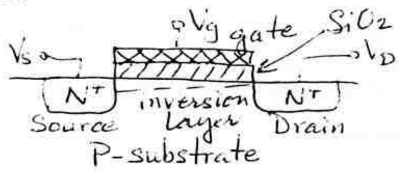
What restrictions must I place on $g(t)$ for such an expansion to be valid? I am not looking for mathematical subtleties, but rather fairly obvious and gross restrictions.

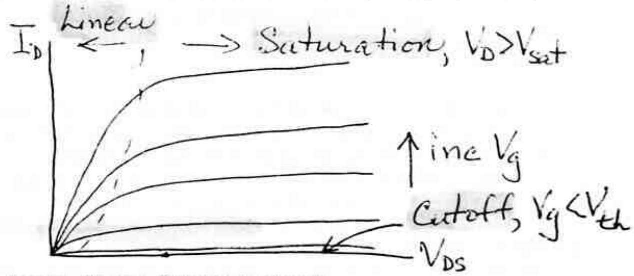
Devices

1997 PhD Quals Question, James Harris

1. Can you sketch the cross-section of a of a MOSFET, draw the I-V characteristic and briefly describe the 2 or 3 most important regions of the I-V characteristic and how these are related to the physics within the device?

A.



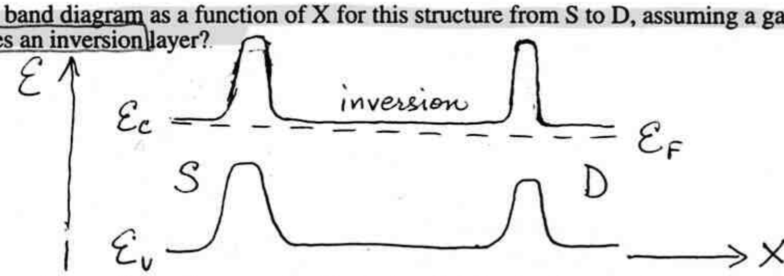


2. What happens if the gate does not overlap the Source &Drain regions?
A. Neglecting fringing fields, the transistor does not function.

3. Can an inversion region still form under the gate with applied V_{gs}?
A. This is a MOS capacitor and upon application of $V_g > V_{th}$, the device first into deep depletion and eventually forms an inversion layer under the gate from thermal generation of carriers.

4. Can you draw band diagram as a function of X for this structure from S to D, assuming a gate bias that creates an inversion layer?

A.

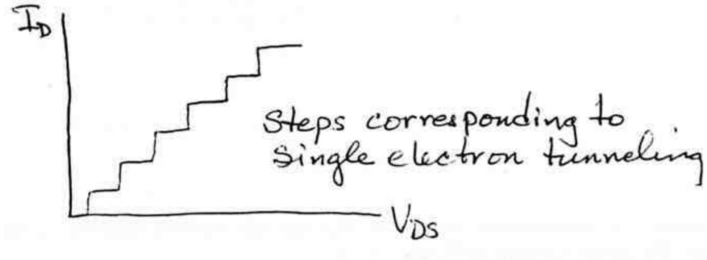


5. If I make the p-regions sufficiently narrow in this device, I claim that I can make a new type of transistor. What might be the important physical mechanisms operative in such a device?
A. Tunneling. If I apply a positive bias to the D, an electron could tunnel from the S into the inversion region which then causes an electron to tunnel into the D, thus allowing a current to flow from S to D thru the barriers shown in the band diagram above.

6. I would like to make a device that could operate at room temperature. If I could make the gate (and hence the inversion region) so small that the tunneling of a single electron into this small region increases the potential by 100meV, how small would this have to be (for instance compared to the gate capacitace of a MOSFET)?
$C = \frac{dq}{dV}$
A. C=DQ/DV=1.6E-19/0.1=1.6E-18 F or 1.6 aF, which is very small-a typical MOSFET gate C_g= 1E-15 F/μm^2 or 1 nF/μm^2, thus the "island" would have to be ~10E-3 μm or 30 nm ? diameter.

7. What might you guess the I-V characterisitc to look like if I had such a small structure and current transport limited by tunneling of single electrons?

A.



leave

Mime-Version: 1.0
Date: Wed, 12 Feb 1997 06:46:18 -0800
To: Diane Shankle <shankle@ee.Stanford.EDU>
From: "C. R. Helms" <helms@ee.Stanford.EDU>
Subject: Re: Quals Questions 1997
Cc: Helms, Fox

1. Consider two capacitors in series with a voltage applied.
   Determine the voltage across each capacitor for the dc case.
2. Discuss how a real world capacitor would differ.
   Recalculate 1 with a model for leakage through the caps.
   Why is this "real world" solution incorrect?
3. How is the capacitance of an individual capacitor determined by the physical properties of the capacitor.
4. Describe where two series capacitances might be found in and MOS device.
5. How would you design the gate insulator for scaled MOS device?

BOB HELMS, PROFESSOR
DIRECTOR - SOLID STATE INDUSTRIAL AFFILIATES
ASSOC. DIRECTOR - NSF/SRC ENGINEERING RESEARCH CENTER
ENVIRONMENTALLY BENIGN SEMIC. MFG.

From: Lambertus Hesselink <bert@optitek.com>
To: "'shankle@ee.stanford.edu'" <shankle@ee.stanford.edu>
Cc: "'bert@kaos.stanford.edu'" <bert@kaos.stanford.edu>
Subject: RE: Quals Questions 1997
Date: Tue, 4 Feb 1997 20:37:08 -0800
MIME-Version: 1.0

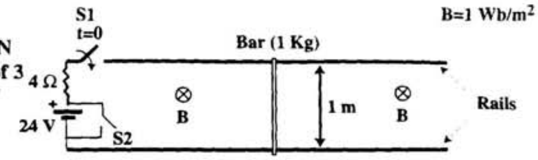
Electro magnetics

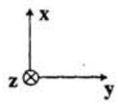
Diane:

Here is the question:
1. Please determine the focallengths of a thick camera lens using a laser beam?
2. How many focal lengths are there?
3. What are the principal planes?
4. Please locate them for the given camera lens using the laser beam
5. By using the laserbeam and the lens I show the diffraction pattern of a glass object placed on top of the lens. By looking at the diffraction pattern, what can you tell me a bout the properties of the glass element?
6. Please compute the spatial frequency of the grating

Electromagnetics

INAN
Q1 of 3
W97





Q: S1 closed at t=0. What does the bar do? Assume no friction.

Biot - Savart Law

A: Initial acceleration (F=ILxB) is 6 m/s^2; bar eventually attains a steady velocity v_o=24 m/s, at which V_{ind} due to v_oxB=24 V

B · V$_o$ L = 24 V

but still form close loop

Q: What would happen now if we take away the battery (S2)

A: Bar slows down and stops.

Q: What if we turn off the magnetic field instead?

A: Bar continues to move at 24 m/s.

no force any more
what is about the magnetic field generated by current in conduction bar itself?

INAN
Q3 of 3
W97

Q: This is a cross section of a waveguide. What is the plotted quantity (E or B)?

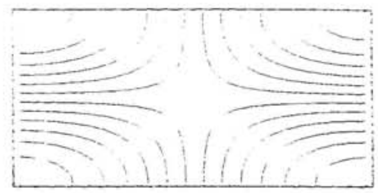
A: E-lines

Q: What mode is it, TE or TM?

A: TE_{11}

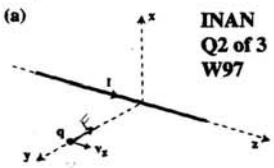
Q: Can you draw the H-lines?

A: See any E&M text.

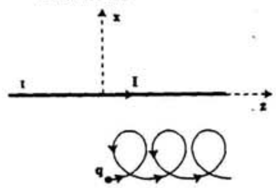


INAN
Q2 of 3
W97

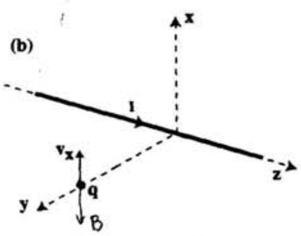
(a)



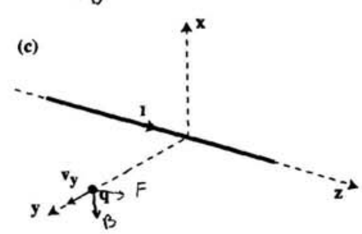
ANSWER:



(b)



(c)



Q: How does charge q move in cases (a), (b) and (c) ?

A: Basic motion is gyration; however, curvature is tighter nearer the wire, so particle drifts to the right. Case (a) and (c) are very similar.

Tom Kailath

Supplementary remarks (3)

(not asked in the exam.)

To understand complex-valued quantities (r.v.s) we should see what the implications of our results are for the real & imaginary parts. So writing

$$X(f) = X_r(f) + j\,X_i(f)$$

and pursuing the implications of the previously derived formulas

$$E[X(f_1)X^*(f_2)] = S_x(f_1)\,\delta(f_1 - f_2)$$

$$E[X(f_1)X(f_2)] = S_x(f_1)\,\delta(f_1 + f_2)$$

gives, after some calculation:

1) $X_r(f_1)$ and $X_i(f_2)$ are uncorrelated for all f_1 and f_2, including $f_1 = f_2$.

2) $X_r(f_1)$ and $X_r(f_2)$ are uncorrelated for $|f_1| \neq |f_2|$

3) $X_i(f_1)$ & $X_i(f_2)$ " " ".

4) $E\,X_r(f_1)X_r(f_2) = \frac{1}{2} S_x(f_1)\,\delta(f_1 - f_2)$
$= E\,X_i(f_1)X_i(f_2)$.

1997 Ph.D Qualifying Exam Question

R.A. Kiehl

The student was shown a schematic of the cross section of a field-effect transistor which identified a two layer (doped on undoped) physical structure and was told some basic things about how the device was fabricated (e.g., S & D contacts made from alloyed metals, gate formed by depositing a metal on directly on a clean semiconductor surface). The schematic also showed depletion regions under the gate for three different bias conditions. Several nonideal I-V characteristics were also shown.

As a warm up, the student was asked to generally explain how the device worked and to explain the general shape of the three depletion regions and identify their corresponding points in the Id-Vds characteristic. Most students correctly described the basic MESFET or JFET operation, although some tried to force the structure to be a MOSFET by assuming the presence oxides, substrate doping, and inversion layers that were not indicated. Others had some difficulty in making a link between the depletion region shapes and the operating points in the I-V plane.

The main part of the exam involved the student speculating on various physical factors that could cause nonideal behavior and suggesting possible solutions to these problems. First they were asked to consider the possibility of whether a source-drain leakage path could occur due to electron injection from the doped layer into the undoped layer and to explain why this could or could not happen. If they were unclear or simply mentioned the conductivity differences between the layers they were asked to sketch the relevant energy band diagrams and consider the problem in that way. They were asked to speculate on possible causes of a pronounced downward bowing of the depletion region near the drain, nonsaturation of the Id-Vds characteristic (high output conductance), lack of drain current pinch off, runaway (sharp increasing) drain current at high Vds, and hysterisis in the Id vs Vgs characteristic.

I was more interested in the persons ability to reason through things and make creative suggestions than in "textbook" answers, which by themselves did not yield full credit. So the students were often asked to more fully explain their answers and to suggest other possibilities. People gained points by supporting their answers with good arguments or by making clever suggestions, even if these were unlikely or not very practical.

Date: Tue, 4 Feb 1997 13:17:45 -0800 (PST)
From: Greg Kovacs <kovacs@cis.Stanford.EDU>
To: shankle@ee.stanford.edu
Subject: Re: Quals Questions 1997

circuits (?)

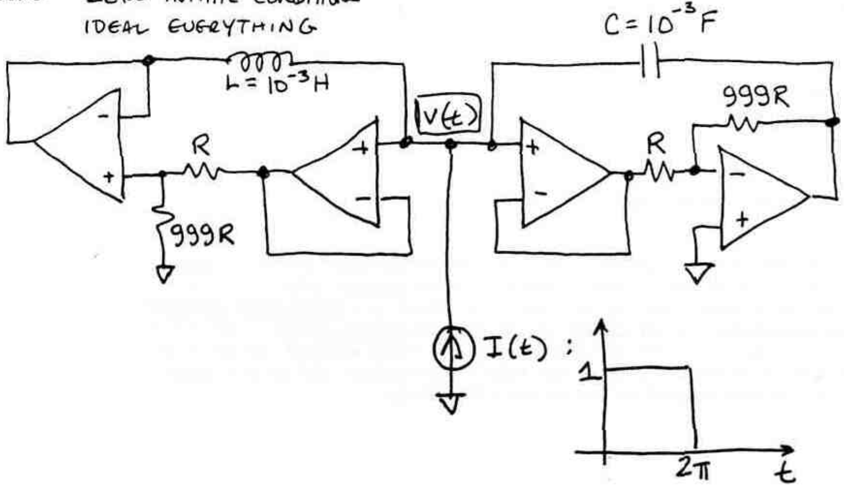
Diane,
Here is my quals question.
Thanks,
Greg

G. Kovacs Quals Question 1997

The student was asked to put on a pair of glasses and look at a multi-colored image. The question was to try to explain the mechanisms underlying the perceived three-dimensional effect, which was based on a wavelength-dependent light bending effect of the optics in the glasses (like a prism). The point of the question was to observe the way the student looked at/used the available information, proposed and carried out simple experiments, and maintained an organized and logical path to his or her conclusions.

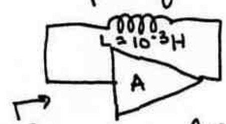
circuits

Assume: ZERO INITIAL CONDITIONS
IDEAL EVERYTHING



Q: Given the current pulse as an input, what is V(t) for $t > 0$?

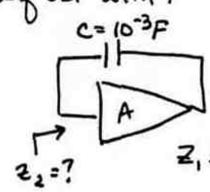
Outline of answer: Replace left-half ckt with this abstraction:



A = 0.999

It is easy to show that $Z_1 = sL \cdot \frac{1}{1-A} = s(10^3)(L) = s \cdot 1\,\text{henry}$

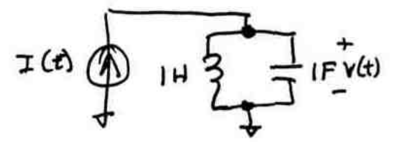
Then, replace right half ckt with:

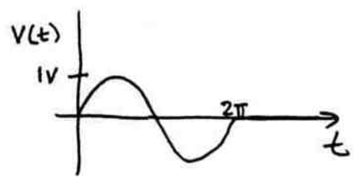


A = −999

$Z_1 = \frac{1}{sC(1-A)} = \frac{1}{s(10^{-3})(10^3)} = \frac{1}{s \cdot 1F}$

So, equivalent ckt is:





circuit oscillates for exactly one cycle, then stays at zero.

David Miller

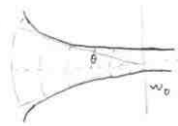
Devices

Part 1

A laser beam has the classic Gaussian form. Sketch the diameter of the beam and the phase fronts of the beam, starting from the beam waist (i.e., starting with the beam collimated) and proceeding to the far field.

Answer

If the candidate happens to be very familiar with Gaussian beams, he or she may know the actual formulae, and be able to describe the transition from plane parallel phase fronts at the waist, through phase fronts with their minimum radius of curvature at the Rayleigh length, through to spherical phase fronts in the far field with radius of curvature given by the distance from the waist and a beam diameter expanding linearly with distance in the far field.



If the candidate does not know the explicit formulae for Gaussian beams, that is fine as long as they can describe that the beam does have plane phase fronts at the waist and spherical wavefronts in the far field (with approximately the correct radius of curvature) with a diffraction angle that is approximately the wavelength divided by the initial beam waist diameter, and ideally be able to give some description of approximately what the diffraction length is (the distance when the diffraction spreading starts to become significant), which is approximately the initial beam area divided by the wavelength. Some basic understanding of diffraction or even basic wave phenomena, together with some common sense dimensional analysis, should enable them to figure most of this out, possibly with a little help from the examiner. If the candidate figured most of this out (or knew it in advance), they would get a high mark on the examination.

$$\theta \approx \frac{2}{k w_0} = \frac{2\lambda}{2\pi w_0} = \frac{\lambda}{\pi w_0}$$

$$\ell = \frac{\pi w_0^2}{\lambda}$$

Part 2

As a bonus to test the candidate who might make their way all through the first part, and to get the highest possible marking on the question, with some candidates I got to the second part.

I put a positive lens precisely at the waist of a laser beam (i.e., where the laser beam is collimated). I want to use the lens to form another beam waist (i.e., a focus) as far away as possible, and I may choose the focal length of the lens to be of any value. Approximately what is the maximum distance at which I can form another beam waist, i.e., how far away can I bring the beam to another focus? (An intelligent guess is a sufficient answer - no need for a proof.)

Answer

This is not a question to which I expect the student to know the answer in advance. Obviously, if I use a short focal length lens, I will form the focus near to the lens. As I increase the focal length, the focus will move farther away, as I want. But if I try to move the focus too far away, the beam will start diffracting out before I get to the desired focus. There is a precise answer to this question for a Gaussian beam, but what I was looking for was looking for was the intelligent guess that the maximum distance is approximately the diffraction length (Rayleigh length). That there is a minimum radius of curvature to the phase fronts at a Rayleigh length in a Gaussian beam is also a hint that the answer is of the order of the Rayleigh length. There is however no other obvious length to use in the answer other than the Rayleigh length, and for that reason alone it is also a good guess.

Throughout the whole examination, I was not interested in the candidate getting factors of two or the like correct. I was interested in at least some intuitive understanding and ability to reason to approximately the correct answer, possibly with some hints from the examiner.

```
From: Ed McCluskey <ejm@shasta.Stanford.EDU>
Subject: please ack receipt of questions from quals
To: shankle@ee.Stanford.EDU (Diane Shankle)
Date: Thu, 6 Feb 97 8:55:52 PST
Cc: ejm@shasta.Stanford.EDU (Ed McCluskey),
        munda@shasta.Stanford.EDU (Siegrid Munda)

Center for Reliable Computing
X-Mailer: ELM [version 2.3 PL11]
```

architecture

```
>
>
> Good Afternoon!
>
```

```
Q1 Draw a multiplexer circuit -- gate level, transistor lever

Q2 Draw a D latch circuit

Q3 Design a field reprogrammable universal combinational logic element
        with 5 inputs
Q4 describe how to test the answer to Q3
>
>
>
EJMc
```

Mime-Version: 1.0
Date: Fri, 7 Feb 1997 11:07:52 -0800
To: Diane Shankle <shankle@ee.Stanford.EDU>
From: len@nova.stanford.edu (Len Tyler)
Subject: Quals question
Cc: len@nova.stanford.edu

Electro magnetics

Quals 97

"The current '911' phone system automatically provides the dispatcher (who answers the phone) with the location of the caller---for calls made over the regular wired system. But this feature does not exist in the cellular phone system.

"Given an assignment to develop a proposal for a caller 'localization' system for cellular phones, how would you proceed?

1. In this application, what do you suggest as a target requirement or goal for localization accuracy?

2. What engineering solutions occur to you?

3. On what physical principles are your solutions based?

4. What, roughly, is the expected accuracy of your various suggestions?

5. What factors control the accuracy?"

Len Tyler
Jan. 1997

G. Leonard Tyler
Professor
Department of Electrical Engineering
Stanford University, California
94305-9515

(415) 723 3535
FAX (415) 723 9251
tyler@ee.stanford.edu

EE Qualifying Exam Questions
(Jan., 97)

physics

Shan Wang
Solid State Lab

1. Consider a metal/insulator/metal junction. If the insulator is very thin, electrons can tunnel through the insulator barrier from metal 1 to metal 2, or vice versa. Now assume that:

(1) The electrons will tunnel through (tunnel probability = 1) if the electrons in both metals are all spin up, or all spin down.
(2) The electrons will not tunnel through (tunnel probability = 0) if the electrons in the metal 1 have opposite spins from the electrons in metal 2.

What would be the tunneling current (or probability) of the electrons when the spin polarization of metal 1 and metal 2 are p_1 and p_2, respectively. The spin polarization is defined as

$$p \equiv a\uparrow - a\downarrow$$

where $a\uparrow$ and $a\downarrow$ are the fraction of spin up electrons and spin down electrons, respectively.

Answer:

For tunneling to occur, the spin directions have to be preserved. Therefore, the probability of electrons tunneling through is

$$P = a\uparrow_1 * a\uparrow_2 + a\downarrow_1 * a\downarrow_2$$

Now that

$$p \equiv a\uparrow - a\downarrow, \quad a\uparrow + a\downarrow = 1$$

so

$$a\uparrow = (1+p)/2, \quad a\downarrow = (1-p)/2$$

Therefore, the tunneling current is

$$I \propto P = (1+p_1)(1+p_2)/4 + (1-p_1)(1-p_2)/4$$

$$= (1+p_1 * p_2)/2.$$

You can check that this formula is valid for the P =1 or P =0 cases.

To: Diane Shankle <shankle@ee.Stanford.EDU>
Subject: Re: Quals Questions 1997
Date: Fri, 31 Jan 1997 15:58:23 -0800
From: Jennifer Widom <widom@DB.Stanford.EDU>

CS

Question

Consider a relational schema with three tables:

Took (studID, courseNum, year, grade)
Student (ID, name, address, birthdate)
Course (num, dept)

Assume that Took is very large; Student and Course are quite a bit smaller. (E.g., Took is 50 times the size of Student, 1000 times the size of Course.)

(a) Write a query in SQL that finds the names of all students who took a class in the EE department and got an A.

(b) Draw a simple relational algebra tree showing how this query might be evaluated. The tree can closely reflect the SQL construction -- don't worry about efficiency.

(c) Suggest several modifications to the tree that would enable more efficient execution.

(d) [If time premits] Suppose we add a condition to the query that we're only interested in students where some "magic" predicate applied to their address and birthdate returns true. Add to the query:

    AND Magic(address,birthdate) = TRUE

Where should this selection condition best be placed in the tree for the most efficient evaluation?

Answer

(a)
```
SELECT name
FROM   Took, Student, Course
WHERE  studID = ID
  AND  courseNum = num
  AND  grade = 'A'
  AND  dept = 'EE'
```

Okay to write using EXISTS subquery, but then rewrite using join.

(b)
```
                PROJECT[name]
                     |
        SELECT[grade=A AND dept=EE]
                     |
          JOIN[courseNum=num]
              /          \
    JOIN[studID=ID]     Course
       /      \
    Took    Student
```

Okay to use CROSS-PRODUCT instead of JOIN and put the two join conditions in the SELECT, but then rewrite using JOIN.

(c) Push selections down to reduce number of intermediate tuples:
    Add SELECT[grade=A] above Took
    Add SELECT[dept=EE] above Course
    Remove top SELECT node

  Change join order to reduce number of intermediate tuples:
    Join (cross-product) Student and Course first

  Add projections to reduce width of intermediate tuples:
    Add PROJECT[studID,courseNum] to above SELECT above Took
    Add PROJECT[ID,name] to above Student
    Add PROJECT[num] to above SELECT above Course

To: Diane Shankle <shankle@ee.Stanford.EDU>
Subject: Re: Quals Questions 1997
Date: Fri, 31 Jan 1997 15:58:28 -0800
From: Jennifer Widom <widom@DB.Stanford.EDU>

CS

Question

Consider a relational schema with three tables:

Took (studID, courseNum, year, grade)
Student (ID, name, address, birthdate)
Course (num, dept)

Assume that Took is very large; Student and Course are quite a bit smaller. (E.g., Took is 50 times the size of Student, 1000 times the size of Course.)

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(b) Draw a simple relational algebra tree showing how this query might be evaluated. The tree can closely reflect the SQL construction -- don't worry about efficiency.

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AND Magic(address,birthdate) = TRUE

Where should this selection condition best be placed in the tree for the most efficient evaluation?

Answer

(a)

```
SELECT name
FROM   Took, Student, Course
WHERE  studID = ID
  AND  courseNum = num
  AND  grade = 'A'
  AND  dept = 'EE'
```

Okay to write using EXISTS subquery, but then rewrite using join.

(b)

```
                PROJECT[name]
                      |
        SELECT[grade=A AND dept=EE]
                      |
          JOIN[courseNum=num]
             /              \
     JOIN[studID=ID]       Course
        /       \
     Took     Student
```

Okay to use CROSS-PRODUCT instead of JOIN and put the two join conditions in the SELECT, but then rewrite using JOIN.

(c) Push selections down to reduce number of intermediate tuples:
    Add SELECT[grade=A] above Took
    Add SELECT[dept=EE] above Course
    Remove top SELECT node

Change join order to reduce number of intermediate tuples:
    Join (cross-product) Student and Course first

Add projections to reduce width of intermediate tuples:
    Add PROJECT[studID,courseNum] to above SELECT above Took
    Add PROJECT[ID,name] to above Student
    Add PROJECT[num] to above SELECT above Course

If anything is unclear, please ask System ?

Idea: for a linear system, a larger input yields a larger output

consider a linear time-invariant system

input u yields output y; input v yields output z

(1.) is it always true that

$$\int_{-\infty}^{\infty} u(t)^2\,dt \ge \int_{-\infty}^{\infty} v(t)^2\,dt$$
$$\Longrightarrow \quad \int_{-\infty}^{\infty} y(t)^2\,dt \ge \int_{-\infty}^{\infty} z(t)^2\,dt \ \ ?$$

discuss

(2.) is it always true that

$$\max_t |u(t)| \ge \max_t |v(t)| \quad \Longrightarrow \max_t |y(t)| \ge \max_t |z(t)| \ \ ?$$

discuss

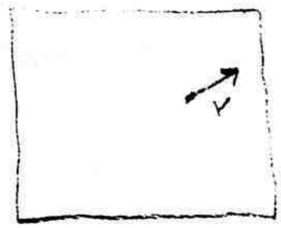
S. Boyd 1997

Qual Questions
Tom Cover, Feb 1997

signals

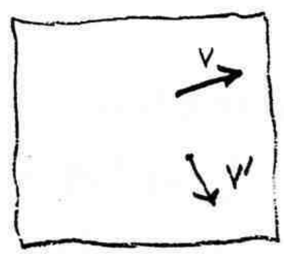
1. An atom with position **x** and velocity **v** is confined to the unit square. What is the time average distribution of the position and velocity of the atom?

a)



Now put two atoms with velocities **v** and **v'** in the box. Same question.

b)



Answer: 1a) $\underline{X} \sim$ uniform
$\underline{V}$ takes on 4 values (flip x component and flip y component of velocity)

b) $\underline{X}_1, \underline{X}_2$ indep. unif.
$(\underline{V}, \underline{V}')$ is roughly uniform on 4-dim [illegible] $\|V\|^2 + \|V'\|$[illegible]

Tom Cover

2. Two Envelopes

You are offered two envelopes. One contains a certain amount of money and the other contains twice that amount.
One of these envelopes is given to you at random. You inspect the contents.

You may keep the contents of this envelope or exchange envelopes.

Question: Should you keep the envelope or should you switch?

Comment: This starts the discussion

Tom Cover

Analysis

Consider the following reasoning.
Suppose the amount observed in the first envelope is y.
Half the time this will be the lesser of the two amounts.
Thus half the time the other envelope will contain $2y$ and half the time the other envelope will contain $y/2$.
Thus the other envelope will contain $5y/4$ on the average..

Thus the factor by which you improve is $5/4$ on the average.

This suggests you should always switch.

Comment: This is the origin of the paradox. The conclusion is fishy.

Tom Cover

Further Analysis

Assume one envelope contains x dollars and the other contains 2x. The amount x is fixed but unknown to you.
One of these envelopes is given to you at random. Call this the first envelope.

Question: What is the expected amount in the first envelope?
What is the expected amount in the second?

Is there any advantage in switching?

<u>Answer</u> The expected amount in each envelope is $3x/2$.

Random switching

As before, assume one envelope contains x and the other contains 2x. The amount x is fixed but unknown.

When you receive one of these envelopes at random, inspect the contents y and switch envelopes with probability 1/y.

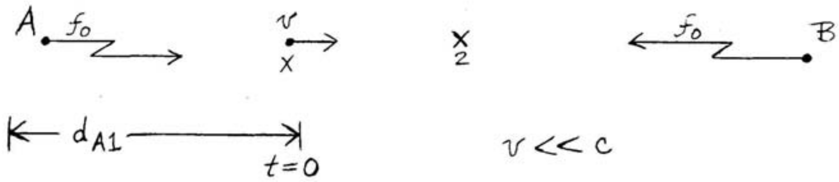
Calculate the expected payoff

Is this switching rule better than not switching?

Answer: Amazingly, random switching helps. The expected amount becomes $\frac{3x}{2} + \frac{1}{4}$ for all $x \geq 1$.

January 1997 Oral Ph.D. Qualification Exam Question
Donald C. Cox Signal

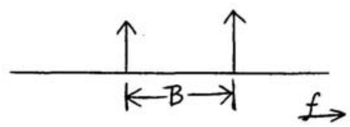
Free Space Large Distances



Two transmitters located in free space at points A and B are radiating continuous wave radio signals towards each other at a single frequency f_0. A spaceship is traveling in a straight line from A to B at a constant velocity v. The velocity v is much less than the velocity of light, i.e. $v << c$. At some reference time $t = 0$, the spaceship is at point 1 a distance d_{A1} from A.

I. First look only at the signal received on the spaceship from point A as the spaceship travels from point 1 towards point 2. Express the signal from point A received on the spaceship in terms of f_0, v, d_{A1} and other constants as appropriate.

II.



$B = 200kHz$
$f_o = 1GHz$

As the spaceship passes point 1 you observe the signals received on the spaceship from points A and B on a spectrum analyzer. You see the 2 delta functions above on the spectrum analyzer display. Can you explain the picture on the display? If $f_0 = 1GHz$ and $B = 200kHz$, what is v?

III. As the spaceship travels between point 1 and point 2, you count the cycles of the Doppler shift frequency (note: few students proceeded this far and by now those had all noted Doppler shift.) accumulated between points 1 and 2. If 100,000 cycles of the Doppler shift frequency were counted between points 1 and 2, what is the distance d_{12} between the two points?

Qualifying Examination 1997

Giovanni De Micheli

January 1997

Architecture

1) What characterizes a Boolean Algebra?

A Boolean algebra is characterized by a set of values (which includes two distinctive values 0 and 1), and by two operations (say AND and OR or conjunction and disjunction) on this set having the commutative and distributive properties. The 1 value is the identity for the AND operation (i.e., x AND $1 = x$) and the 0 value is the identity for the OR operation (i.e., x OR $0 = x$). Moreover, any Boolean value x has a complement x' such that: x AND $x' = 0$ and x OR $x' = 1$.

2) Consider the algebra defined by:

- Carrier: $\{0, 1, 2, 3, 4\}$.
- $+ \rightarrow MAX$
- $* \rightarrow MIN$

Is it a Boolean algebra? Why?

The MAX and MIN operators are commutative and distributive, and their results are always elements of the carrier. Value 0 is the identity element for the MAX operator while value 4 is the identity element for the MIN operator. Unfortunately, it is not possible to define a rule yielding, for each element x in the carrier, a complement element x' satisfying the rule: x MIN $x' = 0$ and x MAX $x' = 4$. Thus this algebra is not Boolean.

3) Consider the algebra defined by:

- Carrier: $\{1, 2, 3, 5, 6, 10, 15, 30\}$.
- $+ \rightarrow lcm$
- $* \rightarrow gcd$

Giovanni De Micheli

Is it a Boolean algebra? Why?

The LCM and GCD operators are commutative and distributive, and their results are always elements of the carrier. Value 1 is the identity element for the LCM operator, while value 30 is the identity element for the GCD operator. Define $x' = 30/x$ to be the complement of x. Then x LCM $x' = 1$ and x GCD $x' = 30$. Thus this algebra is Boolean.

4) Why would you use multiple-valued logic circuits in microelectronic design? Can you list some advantages and disadvantages?

Multiple-valued logic gates use different electric voltages to express the values. The gates are more complex than those implementing binary-valued functions, but circuits that use multiple-valued logic require **less wiring**. *A major disadvantage is coping with* **noise** *that can make it difficult to discriminate logic values.*

Date: Thu, 6 Feb 1997 12:58:50 -0800
From: Michael Flynn <flynn@hobbes.Stanford.EDU>
To: shankle@ee.Stanford.EDU
Subject: Re: Quals Meeting

Qual question (M J Flynn)....two questions...
1) A sequence of addresses reference a cache. If the cache has a 90% hit rate what can you tell me about the information content of the address sequence....(the question can be elaborated to include line size, etc.)
2) What must be done to a multiplier that was designed to take unsigned 8 bit operands and produce a 16 bit product, that is now used to operate on two's complement operands....how can the multiplier be modified to operate correctly?

J. W. Goodman

Answer:

Since the right-hand side of this equation has a spectrum that is bandlimited to (-1/2,1/2), the left-hand side must also be bandlimited to that interval.

3. Find the expansion coefficients in the above expansion.

Answer:

We use the orthogonality of the sinc functions to find coefficients:

$$\int_{-\infty}^{\infty} \operatorname{sinc}(t-m)\, g(t)\, dt = \int_{-\infty}^{\infty} \operatorname{sinc}(t-m) \sum_{n=-\infty}^{\infty} a_n \operatorname{sinc}(t-n)\, dt$$

$$= \sum_{n=-\infty}^{\infty} a_n \int_{-\infty}^{\infty} \operatorname{sinc}(t-m)\operatorname{sinc}(t-n)\, dt = a_m \int_{-\infty}^{\infty} \operatorname{sinc}^2(t-m)\, dt = a_m$$

where we have used the fact that the area under the square of the sinc function is unity. In the case of a bandlimited function, the evaluation can be carried further. Using Parseval's theorem again, and noting that $\mathrm{rect(f)G(f) = G(f)}$,

$$a_m = \int_{-\infty}^{\infty} \operatorname{sinc}(t-m)\, g(t)\, dt = \int_{-\infty}^{\infty} \mathrm{rect(f)G(f)}\, e^{-j2\pi fm} \mathrm{df} = \int_{-\infty}^{\infty} \mathrm{G(f)}\mathrm{e}^{\mathrm{j}2\pi \mathrm{fm}}\, \mathrm{df} = g(-m)$$

Thus the expansion becomes (with $k = -m$)

$$\mathrm{g(t)} = \sum_{\mathrm{k}=-\infty}^{\infty} \mathrm{g(k)\,sinc(t+k)}.$$

You have proved the Nyquist sampling theorem!!

Systems

R.M. Gray's 1997 EE Quals Question

$\{X_n;\ n = 0, 1, \ldots, N-1\}$ is a collection of independent random numbers with probability mass functions

$$p_{X_n}(k) = \Pr(X_n = k) = \frac{1}{2};\ k = -1, +1;\ n = 0, 1, \ldots, N-1$$

Define the random variables Y_k by

$$Y_k = \sum_{n=0}^{N-1} X_n e^{\frac{-j2\pi kn}{N}};\ k = 0, 1, \ldots, N-1$$

---

1. Evaluate $E(Y_k)$, the expected value of Y_k.

   Solution and Comment: Starter question to make sure student familiar with basic probability. Subsequent questions recast in deterministic framework if not. By linearity of expectation,

   $$E(Y_k) = E[\sum_{n=0}^{N-1} X_n e^{\frac{-j2\pi kn}{N}}] = \sum_{n=0}^{N-1} E[X_n] e^{\frac{-j2\pi kn}{N}} = 0$$

   since $E[X_n] = 0$. Many students did this by inspection without writing anything down.

2. Evaluate $E(\frac{1}{N}\sum_{k=0}^{N-1}|Y_k|^2)$

   Solution and Comment: Looked for solution was to think before plugging in and doing detailed evaluation of individual $E[|Y_n|^2]$. This is the energy in the Fourier coefficients and hence by Parsevals theorem should be the same as the sum of the X_n^2, which is just N. Probability and expectation is not needed here.

   Some essentially derived Parseval's theorem, which can be done as

   $$\frac{1}{N}\sum_{k=0}^{N-1} Y_k Y_k^* = \frac{1}{N}\sum_{k=0}^{N-1} Y_k \sum_{n=0}^{N-1} X_n e^{\frac{j2\pi kn}{N}} =$$

   $$\sum_{n=0}^{N-1} X_n \frac{1}{N}\sum_{k=0}^{N-1} Y_k e^{\frac{j2\pi kn}{N}} = \sum_{n=0}^{N-1} |X_n|^2$$

   using the Fourier inversion formula (which can also be proved if needed). Since $|X_n| = 1$ here,

   $$\frac{1}{N}\sum_{k=0}^{N-1} |Y_k|^2 = N.$$

Robert Gray

The probability approach was longer here, but still worked. Plugging in gives

$$E[\frac{1}{N}\sum_{k=0}^{N-1} Y_kY_k^*] = \frac{1}{N}\sum_{k=0}^{N-1} E[Y_kY_k^*]$$

and each of the summand terms is

$$E[Y_kY_k^*] = \left(\sum_{n=0}^{N-1} X_n e^{\frac{-j2\pi kn}{N}}\right)\left(\sum_{n=0}^{N-1} X_n e^{\frac{j2\pi kn}{N}}\right)^* = \sum_{n=0}^{N-1}\sum_{l=0}^{N-1} E[X_nX_m]e^{\frac{-j2\pi kn}{N}}e^{\frac{j2\pi kl}{N}}$$

$$= \sum_{n=0}^{N-1} |X_n|^2$$

since the expections are 0 if $n = m$ and 1 otherwise. Thus $E[Y_kY_k^*] = N$ and the final answer is again N.

3. Evaluate or approximate the following sums assuming that N is very large:

$$\frac{1}{N}\sum_{n=0}^{N-1} X_n\ , \quad \frac{1}{N}\sum_{n=0}^{N-1} Y_n\ , \quad \frac{1}{\sqrt{N}}\sum_{n=0}^{N-1} X_n\ , \quad \frac{1}{\sqrt{N}}\sum_{n=0}^{N-1} Y_n$$

Solution and Comment: Since the X_n are independent and identically distributed, the law of large numbers says that the sample mean is approximately the expectation, so the leftmost sum is close to 0 with high probability.

From Fourier inversion, the second sum is just X_0. If this is not recognized, it can be derived by plugging in:

$$\frac{1}{N}\sum_{n=0}^{N-1} Y_n = \frac{1}{N}\sum_{n=0}^{N-1}\sum_{l=0}^{N-1} X_l e^{\frac{-j2\pi ln}{N}}$$

$$= \sum_{l=0}^{N-1} X_l \frac{1}{N}\sum_{n=0}^{N-1} e^{\frac{-j2\pi ln}{N}} = X_0.$$

From the central limit theorem, the third sum will be approximately a Gaussian random variable with mean 0 and variance 1.

The last sum is just $\sqrt{N}$ times the second sum, hence $\sqrt{N}X_0$

4. Define the matrix A by

$$A = \begin{bmatrix} X_0 & X_1 & X_2 & \dots & X_{N-2} & X_{N-1} \\ X_1 & X_2 & X_3 & \dots & X_{N-1} & X_0 \\ X_2 & X_3 & X_4 & \dots & X_0 & X_1 \\ \vdots & \vdots & \vdots & \dots & \vdots & \vdots \\ X_{N-1} & X_0 & X_1 & \dots & X_{N-3} & X_{N-2} \end{bmatrix}$$

Find an eigenvalue and eigenvector of A.

Solution and Comment: You need to find a vector u and a scalar λ such that $Au = \lambda u$
The matrix equation can be written out as

$$\sum_{n=0}^{N-1} X_{n-k} u_n = \lambda u_k; \; k = 0, 1, \ldots, N-1, \qquad ($$

where the subscript $n-k$ is taken modulo N and where each k corresponds to multiplyin
the kth row of A (the kth cyclic shift of the original vector) by the eigenvector
$u = (u_0, \ldots, u_{N-1})^t$.

The looked for solution was to realize that these sums look like the definition of Y_k. Thus, for example, if the vector u has all ones, multiplication by any row will give $\sum_n X_n = Y_0$. Thus Y_0 is an eigenvalue and the vector of all ones an eigenvector.

A few people realized that the eigenvalues are simply the Y_k, the DFT coefficients, and th eigenvectors look like powers of complex exponentials. (Follows from the shift theorem of Fourier analysis.) But this was more than what was needed in the problem.

Tom Kailath

Other reasons: The power spectrum could ② be bandlimited, so the properties of $X(f)$ could vary drastically as f changes.

We could first try to answer b) & come back to a).

b). $E\,X(f) = 0$. The covariance function of the generally complex $X(f)$ is $E\,X(f_1)X^*(f_2)$. Why the conjugate? One answer: when $f_1 = f_2$ the variance $= E\,|X(f)|^2 > 0$, whereas $E\,X^2(f)$ may be complex (or zero, even though $X(f) \neq 0$).

Now

$$E\,X(f_1)X^*(f_2) = \iint E\,x(t)\,x^*(s)\, e^{-j2\pi f_1 t}\, e^{+j2\pi f_2 s}\,dt\,ds$$

$$= \iint R_x(t-s)\, e^{-j2\pi f_1 (t-s)}\,dt\; e^{+j2\pi (f_2 - f_1)s}\,ds$$

$$= S_x(f_1)\int e^{j2\pi(f_2 - f_1)s}\,ds = S_x(f_1)\,\delta(f_1 - f_2)$$

This is not a function only of $|f_2 - f_1|$, again showing the nonstationarity of $X(f)$.

Comments There is also relevant information about $X(f)$ in $E\,X(f_1)\,X(f_2)$, which some people started with. The answer is

$$E\left[X(f_1)\,X(f_2)\right] = S_x(f_1)\,\delta(f_1 + f_2)$$

What happens when $f_1 = f_2$?

1997 Quals Question
— T. Kailath

1.

$$u(t) \longrightarrow \boxed{\begin{array}{c}\text{LTI}\\ \text{determin.}\end{array}} \xrightarrow{\;y(t)\;}$$

1. $u(t)$ is random, $Eu(t) = 0.3$
2. The step response of the LTI system ~~is so~~ $g(t)$, is such that $g(t) \to 0$ as $t \to \infty$.

FIND $Ey(t)$.

Solution: Let $h(t)$ be the impulse response.

Then $y(t) = \int_{-\infty}^{\infty} h(\tau)\, u(t-\tau)\, d\tau$ & $Ey(t) = 0.3 \int_{-\infty}^{\infty} h(\tau)\, d\tau$.

Now $g(t) = \int_{-\infty}^{t} h(\tau)\, d\tau$. $\therefore \int_{-\infty}^{\infty} h(\tau)\, d\tau = g(\infty) = 0$

& so $\underline{E\, y(t) = 0}$.

2. $x(t)$ is a zero-mean realvalued (wide-sense) stationary random process. Define $X(f) = \int_{-\infty}^{\infty} x(t) e^{-j2\pi f t}\, dt$.

a. Is $X(f)$ a stationary process (as a function of f)?

b. How would you characterize/~~st~~ specify it?

Soln: a. The Fourier transform operation is linear but not time-invariant: changing t to $t+\tau$ changes $X(f)$ to $X(f)e^{-j2\pi f \tau}$

~~other~~ $\therefore$ $X(f)$ is not (wide sense) stationary.

Nicholas McKeown ?

Question 1.

A densely populated country decides to enforce a strict policy on the number of children per family. The problem is that there is a preference for sons, and every family would like to have at least one boy. The policy is as follows:

> *Each family may have as many daughters as they want, but they must stop having children after their first son is born.*

Assume the following:

1. Every family continues to have children until their first son.
2. All children survive.
3. Boys and girls are equally likely to be born.

Part a: Will the number of girls in the country exceed the number of boys?

Part b: Keeping assumptions (2) and (3), we try and devise a new policy such that the expected number of boys exceeds the expected number of girls in the population as a whole. Does such a policy exist?

Nicholas McKeown

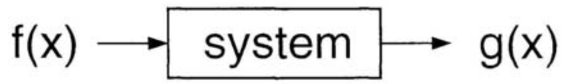
Question 2.

Imagine a population in which the distribution of children per woman is constant over time, and is $a_i = 1/3$, where $i = 1,2,3$. We ask a number of women the following two questions:

1. How many children did you have? (Assume their childbearing days are over).
2. How many children did your mother have?

Part a: What is the average response to (1)?
Part b: What is the average response to (2)?
Part c: Will the population remain constant, increase or decrease?

Dwight Nishimura
system



Let $\mathcal{FT}\{f(x)\} = F(s)$ and $\mathcal{FT}\{g(x)\} = G(s)$,

Are the following systems linear? time-invariant?
What is the transfer function?

1.

$$G(s) = |F(s)|^2$$

2.

$$G(s) = F(-s)$$

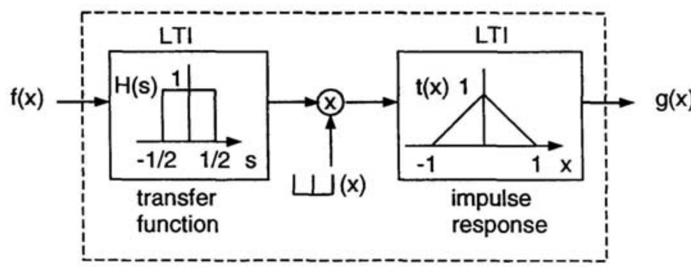
3.

$$g(x) = \frac{df(x-1)}{dx}$$

4.

$$g(x) = \sum_{n=-\infty}^{\infty} f(x-n)$$

5.



6.

$$g(x) = \int_{-\infty}^{x} f(u)\, du$$

Nishimura

Answers

1. g(x) = autocorrelation of f(x). Nonlinear (consider 2 f(x)), Time-Variant (lose phase information).

2. g(x) = f(-x). Linear, Time-Variant (shift input one way, output shifts the other way).

3. Linear, Time-Invariant. H(s) = $j2\pi s \, e^{-i2\pi s}$

4. Linear, Time-Invariant. g(x) = f(x) * comb(x). H(s) = comb(s).

5. Linear (all elements linear), Time-Variant (If input shifted, linear phase in s-domain that should result does not because of replication islands that are not filtered out. Also, interpolation filter does linear interpolation. So consider a slightly shifted version of the input. The piecewise linear output cannot be a shifted version of the original piecewise linear output. Note that if the interpolation filter is a sinc interpolator of appropriate bandwidth, the system is LTI).

6. Linear, Time-Invariant. g(x) = f(x) * step(x). H(s) = 1/2 $\delta(s)$ - i/2πs.

Kunle Olukotun

Computer Architecture

Which application class will benefit most from each enhancement? Explain your reasoning.

	Floating point Scientific	Large DB applicaiton
Large Icache		
Static branch prediciton		
Lots of registers in the ISA		
Victim cache		
Software Prefetching		
4-way S.A. Icache		
Fetch both branch target and fall-through		
Wide Instruction issue with dynamic scheduling		
More branch displacement bits		
Nonblocking cache with many outstanding misses		
Large Dcache block size		
Deep pipelining		
Hardware Prefetching		
High Main Memory bandwidth		

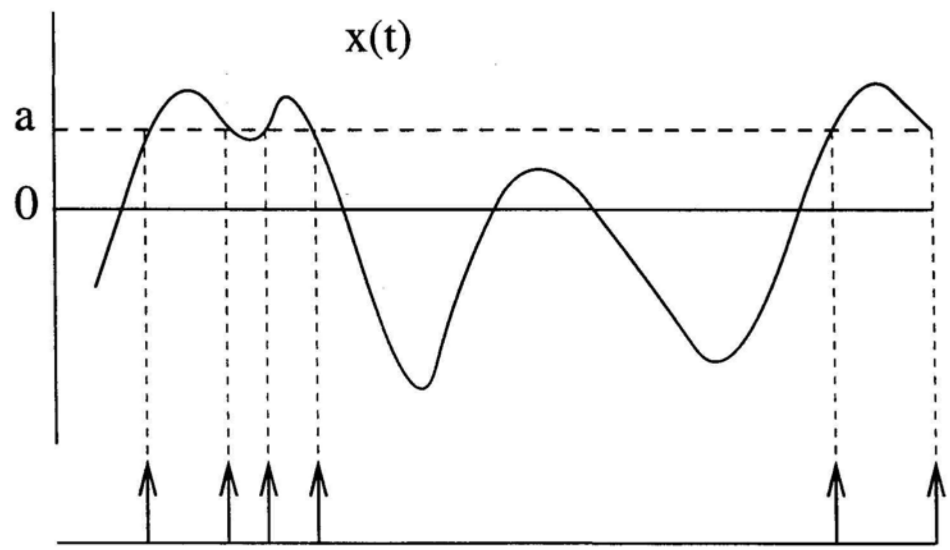
? Paulraj: Quals Questions 1997

Let $x(t)$ be a zero mean, stationary random process.

Let λ_a be the rate parameter for level crossing at level a. i.e.,

$$\lambda_a \;=\; \mathrm{Lim}_{T\to 0}\frac{E[n_a(T)]}{T}$$

Where $n_a(T)$ is number of level crossings at level a in time interval T.



Questions:

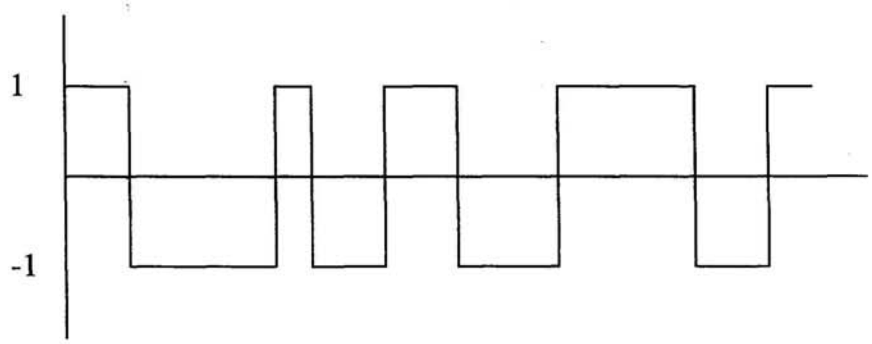
1. On factors does λ_a depend on:

Paulraj

- On a?
- On $p_x(\cdot)$, the probability density of $x(t)$?
- On the bandwidth of $x(t)$?
- On $E[|\frac{dx(t)}{dt}|_{x=a}]$

2. What is the formula for λ_a in terms of the above factors.

3. If $x(t)$ is random telegraph wave taking values ± 1 and with a rate parameter λ



- What is λ_a, $\quad a = 0$?
- What is λ_a, $0 < a < 1$
- What is λ_a, $a = 2$

raulray

4. If $x(t)$ is a zero mean gaussian random process with power spectrum $S_{xx}(\omega) = 1, \quad -\Omega < \omega < \Omega$ or autocorrelation $R_{xx}(\tau)$

- What is relationship between λ_0 and Ω
- What is relationship between λ_0 and $R_{xx}(\tau)$

97 Qualifying Exam
S. Simon Wong

Name : ____________________

Undergrad : ____________________

Area of Interest : ____________________

The students are asked to compare the transit time of carriers inside the base of a BJT and that along the channel of a MOSFET.

The grading is not based on how well one remembers the equations, but the physical arguments that lead to the answers.