USING RANDOM MATRICES IN QUANTUM INFORMATION THEORY

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ABSTRACT. The goal of this series of lectures is to present some recent results in quantum information theory which make use of random matrices. After an introduction to random matrix theory, I will present the method of moments, one of the most successful methods used to study the spectra of large random matrices. This will be the occasion to discuss integration over Gaussian spaces. On the quantum information side, I will focus on two main topics, random quantum states and random quantum channels. I will then prove two recent results, one on the asymptotic eigenvalue distribution of the partial transposition of random quantum states, and another on the output set of random quantum channels. Both will require some terminology and results from free probability, which will also be discussed in detail.

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1. Lecture 1 — Generalities. Wishart matrices

1.1. Introduction. The birth of random matrix theory can be traced to statistics and physics. Wishart introduced the distribution that bears his name in the 1920’s [Wis28], in order to explain the discrepancy between the eigenvalues of a measured covariance matrix, and an expected covariance matrix. Later, Wigner was studying nuclear physics when he introduced [Wig55] the semi-circle distribution. Since then, random matrix theory has played a role in many fields of mathematics and science, including operator algebras [VDN92], combinatorics, complex analysis, theoretical physics and telecommunication theory, just to cite a few. Quantum information theory is definitely one of the most recent of fields of application; for more on this, we direct the interested reader to the recent review [CN16].

In quantum information theory, randomness is built in, by the axioms of quantum mechanics. Since quantum states are modeled by (unit trace, positive semidefinite) matrices, it is clear that the two fields intersect. However, we can see two more reasons for the use of random matrices in quantum information. First, we would like to understand the typical properties of quantum states and channels, relative to tasks ans paradigms in quantum information theory. Very early, properties such as the average entanglement of quantum states were studied [Pag93], and several probability distribution over the set of quantum states were introduced [ZS01]. Second, it turns out that some problems – in particular the minimum output entropy additivity problem, which we discuss at length here – did not have an obvious non-random answer, therefore it became not only natural, but also important, to consider random quantum objects.

One paper which popularized the use of random techniques in quantum information was [HLSW04]. This work pointed out that some well-established techniques in the mathematics of random matrices – measure concentration in this case – could be of use in quantum information.

Let us now gather here some basic definitions from quantum information theory and set up some notation.

A quantum state is a positive semidefinite matrix of unit trace. The set of all quantum states is a convex body denoted by

\[ \mathcal{M}_d^{1+}(\mathbb{C}) := \{ \rho \in \mathcal{M}_d(\mathbb{C}) : \rho \geq 0 \text{ and } \text{Tr} \rho = 1 \}. \]

The extremal points of \( \mathcal{M}_d^{1+}(\mathbb{C}) \) are the rank one projectors \( xx^* \) \((x \in \mathbb{C}^d, \|x\| = 1)\), and they are called pure states.

Of particular interest are states of multiple quantum systems, which are quantum states acting on the tensor power of the corresponding Hilbert spaces. Of particular importance are the separable states, which in the bipartite case can be described as

\[ \mathcal{SEP}_{d_1,d_2} := \text{conv}\{ \rho_1 \otimes \rho_2 \} \, | \rho_i \in \mathcal{M}_d^{1+}(\mathbb{C}) \}. \]

Non-separable states are called entangled, and among those, of particular importance is the maximally entangled state \( d^{-1} \Omega_d \in \mathcal{M}_d^{1+}(\mathbb{C}) \), where

\[ \Omega_d = \sum_{i=1}^{d} e_i \otimes e_i, \tag{1.1} \]

where \( \{e_i\} \) is an orthonormal basis of \( \mathbb{C}^d \).

1.2. Wishart matrices and their limit distribution. The probability density of the normal distribution is:

\[ f(x \mid \mu, \sigma) = \frac{1}{\sigma \sqrt{2\pi}} e^{-\frac{(x-\mu)^2}{2\sigma^2}} \]

Here, \( \mu \) is the mean. The parameter \( \sigma \) is its standard deviation with its variance then \( \sigma^2 \). A random variable with a Gaussian distribution is said to be normally distributed.
Suppose $X$ and $Y$ are random vectors in $\mathbb{R}^k$ such that $(X, Y)$ is a $2k$-dimensional normal random vector. Then we say that the complex random vector $Z = X + iY$ has the complex normal distribution. The normal distribution (resp. random vector) are also called Gaussian distribution (resp. random vectors).

Historically the first ensemble of random matrices having been studied is the Wishart ensemble [Wis28], see [BS10, Chapter 3] or [AGZ10, Section 2.1] for a modern presentation.

**Definition 1.1.** Let $G \in \mathcal{M}_{d \times s}(\mathbb{C})$ be a random matrix with complex, standard, i.i.d. Gaussian entries. The distribution of the positive-semidefinite matrix $W = GG^* \in \mathcal{M}_d(\mathbb{C})$ is called a Wishart distribution of parameters $(d, s)$ and is denoted by $\mathcal{W}_{d, s}$.

The study of the asymptotic behavior of Wishart random matrices is due to Marčenko and Pastur [MP67], while the stronger convergence results have been proved by analytic tools such as determinantal point processes; one can also recover the stronger forms of the theorem as direct consequences of the much more general results [Mal12]. Since we aim at giving complete proofs of our results, we state it here in a rather week form: the convergence in moments.

**Definition 1.2.** A sequence of random matrices $X_d$ is said to converge in moments to a probability distribution $\nu$ if for all positive integers $p$, we have

$$\lim_{d \to \infty} E \int t^p d\mu_{X_d} = E \frac{1}{d} \text{Tr}(X_d^p) = \int t^p d\nu,$$

where $\mu_{X_d}$ is the empirical eigenvalue distribution of $X_d$

$$\mu_{X_d} = \frac{1}{d} \sum_{i=1}^d \delta_{\lambda_i(X_d)}.$$

**Theorem 1.3.** Consider a sequence $s_d$ of positive integers which behaves as $s_d \sim cd$ as $d \to \infty$, for some constant $c \in (0, \infty)$. Let $W_d$ be a sequence of positive-semidefinite random matrices such that $W_d$ is distributed according to $\mathcal{W}_{d, s_d}$. Then, the sequence $W_d$ converges in moments to the Marčenko-Pastur distribution $\pi_c$ given by

$$\pi_c = \max(1 - c, 0) \delta_0 + \frac{\sqrt{(b - x)(x - a)}}{2\pi x} 1_{(a, b)}(x) \, dx,$$

(1.2)

where $a = (1 - \sqrt{c})^2$ and $b = (1 + \sqrt{c})^2$.

The Marčenko-Pastur distribution $\pi_c$ is sometimes called the free Poisson distribution, see [NS06, Proposition 12.11]. We plotted in Figure 1 its density in the cases $c = 1$ and $c = 4$.

![Figure 1](attachment:image.png)
Remark 1.4. The Dirac mass appearing in (1.2) is due to the fact that if \( c < 1 \), the matrix \( W_d \) is rank deficient. Since \( cd < d \), a fraction \( 1 - c \) of the eigenvalues of \( W_d \) are null, yielding the Dirac mass at zero.

We postpone the proof of Theorem 1.3 to Section 1.6.

We end this section by the statement of the so-called Carleman condition, which ensures that a sequence of moments defines a unique probability measure.

**Proposition 1.5.** Let \( \mu \) be a probability measure on \( \mathbb{R} \) having finite moments

\[
    m_n = \int t^n d\mu(t)
\]

which satisfy

\[
    \sum_{n=1}^{\infty} m_{2n}^{-1/(2n)} = +\infty.
\]

Then, \( \mu \) is the only measure on \( \mathbb{R} \) having the sequence \( (m_n) \) as moments.

1.3. **Graphical notation for tensors.** Most operations from linear and multilinear algebra (composition, tensor product, (partial) traces) can be efficiently represented graphically. The leading idea is that a string in a diagram means a tensor contraction. Many graphical theories for tensors and linear algebra computations have been developed in the literature [Pen05, Coe10]. Although they are all more or less equivalent, we will stick to the one introduced in [CN10b], as it allows to compute the expectation of random diagrams in a diagrammatic way subsequently. For more details on this method, we refer the reader to the paper [CN10b] and to other work which make use of this technique [CN11a, CN10a, CZ10, FŚ13, CGGPG13, Lan15].

In the graphical calculus, matrices (or, more generally, tensors) are represented by boxes. Each box has differently shaped symbols, where the number of different types of them equals that of different spaces (exceptions are mentioned below). Those symbols are empty (white) or filled (black), corresponding to primal or dual spaces. Wires connect these symbols, corresponding to tensor contractions. A diagram is a collection of such boxes and wires and corresponds to an element of an abstract tensor product space. Rather than going through the whole theory, we focus next on a few key examples.

Suppose that each diagram in Figure 2 comes equipped with two vector spaces \( V_1 \) and \( V_2 \) which we shall represent respectively by circle and square shaped symbols. In the first diagram, \( M \) is a tensor (or a matrix, depending on which point of view we adopt) \( M \in V_1^* \otimes V_1 \), and the wire applies the contraction \( V_1^* \otimes V_1 \to \mathbb{C} \) to \( M \). The result of the diagram \( D_a \) is thus \( T_{D_a} = \text{Tr}(M) \in \mathbb{C} \). In the second diagram, again there are no free decorations, hence the result is the complex number \( T_{D_b} = \langle y, Mx \rangle \). Finally, in the third example, \( N \) is a \((2, 2)\) tensor or a linear map \( N \in \text{End}(V_1 \otimes V_2, V_1 \otimes V_2) \). When one applies to the tensor \( N \) the contraction of the couple \((V_1, V_1^*)\), the result is the partial trace of \( N \) over the space \( V_1 \): \( T_{D_c} = \text{Tr}_{V_1}(N) \in \text{End}(V_2, V_2) \). We depict in Figure 3 the maximally entangled (un-normalized) state \( \Omega_d \) from (1.1), as well as its partial trace, \( \text{[id} \otimes \text{Tr]}(\Omega_d) = I_d \).
1.4. Wick formula, algebraic and diagrammatic formulations. The following theorem is the link between combinatorics and probability theory for Gaussian vectors: it allows to compute moments of any Gaussian vector thanks to its covariance matrix. A Gaussian space $V$ is a real vector space of random variables having moments of all orders, with the property that each of these random variables has centered Gaussian distributions. In order to specify the covariance information, such a Gaussian space comes with a positive symmetric bilinear form $(x, y) \to \mathbb{E}[xy]$. Gaussian spaces are in one-to-one correspondence with Euclidean spaces. In particular, the Euclidean norm of a random variable determines it fully (via its variance) and if two random variables are given, their joint distribution is determined by their angle. The following is usually called the Wick Lemma:

**Theorem 1.6.** Let $V$ be a Gaussian space and $x_1, \ldots, x_k$ be elements in $V$. If $k = 2l + 1$ then $\mathbb{E}[x_1 \cdots x_k] = 0$ and if $k = 2l$ then

$$
\mathbb{E}[x_1 \cdots x_k] = \sum_{p=\{(i_1,j_1),\ldots,(i_l,j_l)\}} \prod_{m=1}^{l} \mathbb{E}[x_{i_m}x_{j_m}]. \quad (1.3)
$$

In particular it follows that if $x_1, \ldots, x_p$ are independent standard Gaussian random variables, then

$$
\mathbb{E}[x_1^{2k_1} \cdots x_p^{2k_p}] = \prod_{i=1}^{p} (2k_i)!!. 
$$

The main difference between the real case discussed above and the complex case is that one has to pair Gaussian variables to their conjugates in the complex situation. This follows from the fact that if $Z$ is a standard complex Gaussian random variable,

$$
\mathbb{E}[Z^2] = \mathbb{E}[\overline{Z}^2] = 0, \quad \text{while} \quad \mathbb{E}[Z\overline{Z}] = 1.
$$

We shall now recast the Wick formula above in the graphical formalism described in the previous section. Consider a diagram which contains a new special box $G$ corresponding to a Gaussian random matrix. We shall compute the expected value of a random diagram with respect to the Gaussian probability measure; as we shall see, this operation will consist of expanding the diagram, by erasing the Gaussian boxes and replacing them with wires.

To start, consider $D$ a diagram which contains, amongst other constant tensors, boxes corresponding to independent Gaussian random matrices of covariance one (identity). One can deal with more general Gaussian matrices by multiplying the standard ones with constant matrices. Note that a box can appear several times, adjoints of boxes are allowed and the diagram may be disconnected. Also, Gaussian matrices need not be square.

The expectation value of such a random diagram $D$ can be computed by a removal procedure as in the unitary case. Without loss of generality, we assume that we do not have in our diagram adjoints of Gaussian matrices, but instead their complex conjugate box. This assumption allows for a more straightforward use of the Wick formula from Theorem 1.6. We can assume that $D$ contains only one type of random Gaussian box $G$; other independent random Gaussian matrices are assumed constant at this stage as they can be removed in the same manner afterwards.

A removal of the diagram $D$ is a pairing between Gaussian boxes $G$ and their conjugates $\overline{G}$. The set of removals is denoted by $\text{Rem}_G(D)$ and it may be empty: if the number of $G$ boxes is different
Figure 4. Pairing of boxes in the Gaussian case

Figure 5. Applying Theorem 1.7 to compute $E[GAG^*].$

from the number of $\bar{G}$ boxes, then $\text{Rem}_G(D) = \emptyset$ (since no pairing between matrices and their conjugates can exist). Otherwise, a removal $r$ can identified with a permutation $\alpha \in S_p,$ where $p$ is the number of $G$ and $\bar{G}$ boxes. In the Gaussian/Wick calculus, one pairs conjugate boxes: white and black decorations are paired in an identical manner, hence only one permutation is needed to encode the removal.

To each removal $r$ associated to a permutation $\alpha \in S_p$ corresponds a removed diagram $D_r$ constructed as follows. One starts by erasing the boxes $G$ and $\bar{G},$ but keeps the decorations attached to these boxes. Then, the decorations (white and black) of the $i$-th $G$ box are paired with the decorations of the $\alpha(i)$-th $\bar{G}$ box in a coherent manner, see Figure 4.

The graphical reformulation of the Wick formula from Theorem 1.6 becomes the following theorem, which we state without proof.

**Theorem 1.7.** The following holds true:

$$E_G[D] = \sum_{r \in \text{Rem}_G(D)} D_r.$$  

In Figure 5, we present an example of application of the theorem above. We consider, on the first row, the diagram corresponding to $E[GAG^*],\,$ where $G \in M_{n \times k}(\mathbb{C})$ is a $n \times k$ Gaussian matrix, and $A \in M_k(\mathbb{C})$ is a square, deterministic matrix. The first row contains the diagram $D$ associated to the algebraic expression. In the second row, we rewrite the same diagram, replacing $G^*$ by $\bar{G}^T,$ in order to be able to apply Theorem 1.7. The third row contains the result of the application: we erase the $G/\bar{G}$ boxed and we add the wires corresponding to the permutation $(1) \in S_1$ (in red). We recognize the diagrams for the identity matrix and for the trace of $A$: $E[GAG^*] = \text{Tr}(A)I_n.$

1.5. **Non-crossing partitions and permutations.** For a permutation $\sigma \in S_p,$ denote by $\# \sigma$ the number of its cycles, including the trivial ones (fixed points). Denote also by $|\sigma|$ its **length,** i.e. the minimum number of transposition which multiply to $\sigma.$ It is well known that for all permutations
The set of non-crossing partitions will play a crucial role in what follows. Recall that a partition \( \pi \) of \( [p] := \{1, 2, \ldots, p\} \) is called non-crossing if there are no quadruples \((a, b, c, d)\) such that \(a, b\) (resp. \(c, d\)) belong to the same block of \( \pi \), and \(a < c < b < d\); see Figure 6 for some examples. The are supremum and infimum operations on \( NC(p) \), which turn it into a lattice, see [NS06, Lecture 9]. The number of elements in the set \( NC(p) \) is the Catalan number

\[
\text{Cat}_p = \frac{1}{p+1} \binom{2p}{p}.
\]

These numbers satisfy the recurrence relation

\[
\text{Cat}_p = \sum_{i=1}^{p} \text{Cat}_{i-1}\text{Cat}_{p-i},
\]

and thus their generating series is given by

\[
M(z) = \sum_{p=0}^{\infty} \text{Cat}_p z^p = \frac{1 - \sqrt{1 - 4z}}{2z}.
\]

We collect now a some properties of the distance function over the symmetric group, which allow us to bijectively identify a subset of \( S_p \) with \( NC(p) \). This result can be traced back to [Bia97].

Lemma 1.8. The function \( d(\sigma, \tau) = |\sigma^{-1}\tau| \) is an integer valued distance on \( S_p \). Besides, it has the following properties:

- the diameter of \( S_p \) is \( p - 1 \);
- \( d(\cdot, \cdot) \) is left and right translation invariant;
- for three permutations \( \sigma_1, \sigma_2, \tau \in S_p \), the quantity \( d(\tau, \sigma_1) + d(\tau, \sigma_2) \) has the same parity as \( d(\sigma_1, \sigma_2) \);
- the set of geodesic points (elements which saturate the triangular inequality) between the identity permutation and some permutation \( \sigma \in S_p \) is in bijection with the set of non-crossing partitions smaller than \( \pi \), where the partition \( \pi \) encodes the cycle structure of \( \sigma \). Moreover, the preceding bijection preserves the lattice structure.

1.6. Proof of the Marcenko-Pastur theorem. Proof of the Marčenko-Pastur theorem We have now all the elements to present a short and elegant proof of Theorem 1.3.

Proof of Theorem 1.3. The proof will consist of three independent steps: computing the moments, at fixed \( d \), of the random matrix \( W_d \), letting \( d \to \infty \) and computing the limiting moments, and finally identifying the probability measure having precisely these moments.

**Step 1. Moment formula**

We are interested, for any fixed dimensions \( d, s \), in computing the \( p \)-th moment of the random matrix \( W_d = GG^* \), where \( G \) is a \( d \times s \) matrix with i.i.d. complex standard Gaussian random entries. To do this, we consider the diagram \( D \) corresponding to the random variable \( \text{Tr}(W_d^p) \). This diagram contains \( p \) pairs \((G, G')\) of Gaussian boxes, which are connected as in Figures 7 and 8. More precisely, the label corresponding to \( \mathbb{C}^d \) which is attached to the \( i \)-th \( G \)-box is connected to the
Figure 7. The first moment of a Wishart matrix using the graphical Wick calculus from Theorem 1.7. Round labels correspond to $\mathbb{C}^d$, while square labels correspond to $\mathbb{C}^s$.

Figure 8. The second moment of a Wishart matrix using the graphical Wick calculus. On the top row, the diagram for $E \text{Tr}(W_d^2)$. On the bottom row, the two diagrams corresponding to the permutations $\text{id} = (1)(2)$, on the left, and $(12)$, on the right. Their values are respectively $ds^2$ and $d^2s$.

corresponding label attached to the $(i-1)$-th $\tilde{G}$-box. On the other hand, the label corresponding to $\mathbb{C}^s$ which is attached to the $i$-th $G$-box is connected to the corresponding label attached to the $i$-th $\tilde{G}$-box. Using the graphical Wick formula from Theorem 1.7, we have

$$E \text{Tr}(W^p_d) = \mathbb{E}D = \sum_{\alpha \in \mathcal{S}_p} D_{\alpha},$$

where $D_{\alpha}$ is the removal diagram obtained by deleting the $G/\tilde{G}$ boxed and connecting the labels according to the permutation $\alpha$. It is clear that each diagram $D_{\alpha}$ consists only of loops of two types: ones coming from round labels corresponding to $\mathbb{C}^d$ spaces, and others coming from square labels corresponding to $\mathbb{C}^s$ spaces. The number of loops of each type is the number of cycles in the permutation $\beta^{-1} \alpha$, where $\beta$ encodes the initial wiring of the labels of each type; see Figures 7 and 8 for some examples. In conclusion, we have

$$E \text{Tr}(W^p_d) = \sum_{\alpha \in \mathcal{S}_p} d^{\#(\gamma^{-1} \alpha)} s^{\#\alpha}. \quad (1.4)$$

In the formula above, $\#(\cdot)$ is the number of cycles function, and $\gamma$ is the full cycle permutation

$$\gamma = (p(p-1) \cdots 3 2 1) \in \mathcal{S}_p.$$

**Step 2. Asymptotic moments**

Let us now consider the asymptotic regime we are interested in, $d \to \infty$ and $s \sim cd$, for some fixed parameter $c \in (0, \infty)$. Since the terms in (1.4) are all positive, we have

$$E \text{Tr}(W^p_d) \sim \sum_{\alpha \in \mathcal{S}_p} c^{\#\alpha} d^{\#(\gamma^{-1} \alpha) + \#\alpha}.$$

The dominating terms in the sum above are those maximizing the quantity $\#(\gamma^{-1} \alpha) + \#\alpha$ over the symmetric group. Using the properties of the distance function $|\cdot|$ on permutations from Lemma 1.8, we have

$$\#(\gamma^{-1} \alpha) + \#\alpha = 2p - (|\alpha| + |\gamma^{-1} \alpha|) \leq 2p - |\gamma| = p + 1,$$
where equality is attained iff \( \alpha \) is a geodesic permutation (it saturates the triangle inequality \(|\text{id}^{-1} \alpha| + |\alpha^{-1} \gamma| \geq |\text{id}^{-1} \gamma|\)). We conclude that
\[
E \text{Tr}(W_d^p) \sim d^{p+1} \sum_{\sigma \in \text{NC}(p)} c^\#\sigma.
\]
Notice that considering only the dominating terms from the sum (1.4), indexed over all permutations, selects the ones for which the permutations are non-crossing partitions.

**Step 3. The Marčenko-Pastur distribution**

We are going to treat here the case \( c = 1 \); the general case is similar. We can rewrite the asymptotic moment formula as
\[
\lim_{d \to \infty} \frac{1}{d} \text{Tr}[(d^{-1}W_d)^p] = \text{Cat}_p.
\]

We claim that the unique probability measure \( \mu \) having the Catalan numbers as moments is the one from (1.2):
\[
\pi_1 = \sqrt{x(4-x)} \frac{2\pi x}{(x,4)(x)} dx.
\]

To show this, recall that the generating function of the Catalan number must be the moment generating function of \( \mu \):
\[
M_\mu(z) = \sum_{p=0}^{\infty} z^p \int t^p d\mu = \frac{1 - \sqrt{1 - 4z}}{2z},
\]
where the relation above holds formally (as a power series in \( z \)), and analytically, in a small neighborhood of 0. The Cauchy transform of \( \mu \) reads now
\[
G_\mu(z) = \int \frac{1}{z - t} d\mu(t) = z^{-1}M_\mu(z^{-1}) = \frac{1 - \sqrt{1 - 4z^{-1}}}{2},
\]
which holds now on a neighborhood of the infinity in the complex plane. One recovers the density of \( \mu \) via the Stieltjes inversion formula, which says that if we denote by
\[
h_\varepsilon(t) := -\frac{1}{\pi} \Im G_\mu(t + i\varepsilon),
\]
then
\[
\frac{d\mu}{dt} = \lim_{\varepsilon \to 0} h_\varepsilon(t).
\]
In our case, we recover \( \mu = \pi_1 \).

The uniqueness claim comes from the fact that \( \pi_1 \) is compactly supported, hence it satisfies the Carleman condition from Proposition 1.5.

\[\square\]

2. **Lecture 2 — Partial transposition of random quantum states. Free probability**

2.1. **Some elements of free probability theory.** We have studied random matrices in the previous lecture by their moments: the only properties of the ambient probability space we have used were the fact that the random variables have an algebra structure, and the existence of the expectation functional. We abstract out these notions in the following definition [NS06, Lecture 1].

**Definition 2.1.** A non-commutative probability space is an algebra \( \mathcal{A} \) with unit endowed with a tracial state \( \varphi \). An element of \( \mathcal{A} \) is called a (non-commutative) random variable.
In classical probability theory, the notion of independence of random variables plays a very important role; in particular, it allows to compute the joint distribution of independent random variables in terms of the marginal distributions (i.e. the distributions of the individual random variables). The notion of freeness is a non-commutative alternative to classical independence.

**Definition 2.2.** Let $A_1, \ldots, A_k$ be subalgebras of $A$ having the same unit as $A$. They are said to be free if for all $a_i \in A_j$ ($i = 1, \ldots, k$) such that $\varphi(a_i) = 0$, one has

$$\varphi(a_1 \cdots a_k) = 0$$

as soon as $j_1 \neq j_2, j_2 \neq j_3, \ldots, j_{k-1} \neq j_k$. Collections $S_1, S_2, \ldots$ of random variables are said to be free if the unital subalgebras they generate are free.

Let $(a_1, \ldots, a_k)$ be a $k$-tuple of selfadjoint random variables and let $\mathbb{C}(X_1, \ldots, X_k)$ be the free $*$-algebra of non commutative polynomials on $\mathbb{C}$ generated by the $k$ indeterminates $X_1, \ldots, X_k$. The *joint distribution* of the family $\{a_i\}_{i=1}^k$ is the linear form

$$\mu(a_1, \ldots, a_k) : \mathbb{C}(X_1, \ldots, X_k) \to \mathbb{C}$$

$$P \mapsto \varphi(P(a_1, \ldots, a_k)).$$

In the case of a single, self-adjoint random variable $x$, if the moments of $x$ coincide with those of a compactly supported probability measure $\mu$, i.e.

$$\forall p \geq 1, \quad \varphi(x^p) = \int t^p d\mu(t),$$

we say that $x$ has distribution $\mu$. The most important distribution in free probability theory is the semicircular distribution

$$\mu_{SC(0,1)} = \frac{\sqrt{4 - x^2}}{2\pi} 1_{[-2,2]}(x)dx,$$

which is, for reasons we will not get into, the free world equivalent of the Gaussian distribution in classical probability (see [NS06, Lecture 8] for the details). A random variable $x$ having distribution $\mu_{SC(0,1)}$ has the Catalan number for moments:

$$\varphi(x^p) = \begin{cases} \text{Cat}_p := \frac{1}{p+1}(\frac{2p}{p})!, & \text{if } p \text{ is even} \\ 0, & \text{if } p \text{ is odd} \end{cases}$$

More generally, if $x$ has distribution $\mu_{SC(0,1)}$, we say that $y = \sigma x + m$ has distribution

$$\mu_{SC(m,\sigma^2)} = \frac{\sqrt{4\sigma^2 - (x-m)^2}}{2\pi \sigma^2} 1_{[m-2\sigma,m+2\sigma]}(x)dx. \quad (2.1)$$

![Figure 9](image-url) The density of the semicircular distributions $\mu_{SC(0,1)}$ (left) and $\mu_{SC(1,1/4)}$ (right).
The bounded operators on $\mathcal{F}(H)$, together with the vacuum state
\[ \tau(X) = \langle \Omega, X\Omega \rangle \]
form a non-commutative probability space. We also define, for a vector $f \in H$, the creation and annihilation operators $\ell(h)$ and $\ell(h)^*$, defined as follows:
\[ \ell(f)\Omega = f \]
\[ \ell(f)f_1 \otimes \cdots \otimes f_n = f \otimes f_1 \otimes \cdots \otimes f_n \]
and
\[ \ell(f)^*\Omega = 0 \]
\[ \ell(f)^*f_1 = \langle f, f_1 \rangle\Omega \]
\[ \ell(f)^*f_1 \otimes \cdots \otimes f_n = \langle f, f_1 \rangle f_2 \otimes \cdots \otimes f_n. \]

The following theorem is taken from [NS06, Section 7], where it is proven in a more general form.

**Theorem 2.5.** Let $f, g \in H$ be two orthogonal vectors. Then the non-commutative random variables $x = \ell(f) + \ell(f)^*$ and $y = \ell(g) + \ell(g)^*$ are semicircular and free.

**Proof.** Let us first show that both $x$ and $y$ have semicircular distributions; moreover, without loss of generality, let us assume that $\|f\| = 1$, and task to show that $x$ has $\mu_{SC(0,1)}$ distribution.

To do this, fix some moment order $p$, and consider $\tau(x^p)$:
\[ \tau(x^p) = \sum_{w: |p| \to \{1,*\}} \langle \Omega, \ell(f)^w(p)\ell(f)^{(p-1)} \cdots \ell(f)^{(2)}\ell(f)^{(1)}\Omega \rangle. \]

For each choice of the function $w$, the scalar product above is either 0 or 1; we have thus to count how many choices of $w$ give 1. It is clear that a function $w$ gives 1 if $p = 2q$ is even, and the lattice path induced by $w$ is a Dyck path. Recall that a Dyck path is a path in the lattice $\mathbb{Z}^2$, starting at $(0,0)$, ending at $(2q,0)$, having $(1, \pm 1)$ steps, and, importantly, staying above the $x$-axis at all times; see Figure 10 for an example. The number of such paths is given by the Catalan numbers.

**Remark 2.3.** If the non-commutative random variable $x$ has (standard) semicircular distribution, then $x^2$ has a free Poisson (or Marchenko-Pastur distribution) of parameter $c = 1$.

Given a $k$-tuple $(a_1, \ldots, a_k)$ of free random variables such that the distribution of $a_i$ is $\mu_{a_i}$, the joint distribution $\mu_{(a_1, \ldots, a_k)}$ is uniquely determined by the $\mu_{a_i}$’s. A family $(a_1^n, \ldots, a_k^n)_n$ of $k$-tuples of random variables is said to converge in distribution towards $(a_1, \ldots, a_k)$ iff for all $P \in \mathbb{C}(X_1, \ldots, X_k)$, $\mu_{(a_1^n, \ldots, a_k^n)}(P)$ converges towards $\mu_{(a_1, \ldots, a_k)}(P)$ as $n \to \infty$. Sequences of random variables $(a_1^n, \ldots, a_k^n)_n$ are called asymptotically free as $n \to \infty$ if the $k$-tuple $(a_1^n, \ldots, a_k^n)_n$ converges in distribution towards a family of free random variables.

Given two free random variables $a, b \in \mathcal{A}$, the distribution $\mu_{a+b}$ is uniquely determined by $\mu_a$ and $\mu_b$. The free additive convolution of $\mu_a$ and $\mu_b$ is defined by $\mu_a \boxplus \mu_b = \mu_{a+b}$. When $x = x^* \in \mathcal{A}$, we identify $\mu_x$ with the spectral measure of $x$ with respect to $\tau$. The operation $\boxplus$ induces a binary operation on the set of probability measures on $\mathbb{R}$. Similarly, we write $\mu_a \boxdot \mu_b = \mu_{a-b}$.

**2.2. The full Fock space, free semicircular random variables.** We discuss now a more abstract non-commutative probability space, in which freeness appears naturally.

**Definition 2.4.** Let $H$ be a complex Hilbert space. The full Fock space over $H$ is defined to be
\[ \mathcal{F}(H) = \bigoplus_{n=0}^{\infty} H^\otimes n = \mathbb{C}\Omega \oplus \bigoplus_{n=1}^{\infty} H^\otimes n. \]

The bounded operators on $\mathcal{F}(H)$, together with the vacuum state
\[ \tau(X) = \langle \Omega, X\Omega \rangle \]
form a non-commutative probability space. We also define, for a vector $f \in H$, the creation and annihilation operators $\ell(h)$ and $\ell(h)^*$, defined as follows:
\[ \ell(f)\Omega = f \]
\[ \ell(f)f_1 \otimes \cdots \otimes f_n = f \otimes f_1 \otimes \cdots \otimes f_n \]
and
\[ \ell(f)^*\Omega = 0 \]
\[ \ell(f)^*f_1 = \langle f, f_1 \rangle\Omega \]
\[ \ell(f)^*f_1 \otimes \cdots \otimes f_n = \langle f, f_1 \rangle f_2 \otimes \cdots \otimes f_n. \]

The following theorem is taken from [NS06, Section 7], where it is proven in a more general form.

**Theorem 2.5.** Let $f, g \in H$ be two orthogonal vectors. Then the non-commutative random variables $x = \ell(f) + \ell(f)^*$ and $y = \ell(g) + \ell(g)^*$ are semicircular and free.

**Proof.** Let us first show that both $x$ and $y$ have semicircular distributions; moreover, without loss of generality, let us assume that $\|f\| = 1$, and task to show that $x$ has $\mu_{SC(0,1)}$ distribution.

To do this, fix some moment order $p$, and consider $\tau(x^p)$:
\[ \tau(x^p) = \sum_{w: |p| \to \{1,*\}} \langle \Omega, \ell(f)^w(p)\ell(f)^{(p-1)} \cdots \ell(f)^{(2)}\ell(f)^{(1)}\Omega \rangle. \]

For each choice of the function $w$, the scalar product above is either 0 or 1; we have thus to count how many choices of $w$ give 1. It is clear that a function $w$ gives 1 if $p = 2q$ is even, and the lattice path induced by $w$ is a Dyck path. Recall that a Dyck path is a path in the lattice $\mathbb{Z}^2$, starting at $(0,0)$, ending at $(2q,0)$, having $(1, \pm 1)$ steps, and, importantly, staying above the $x$-axis at all times; see Figure 10 for an example. The number of such paths is given by the Catalan numbers, and the first part of the proof is complete.
Let us now show that $x$ and $y$ are free. Let us first identify which elements in the algebra generated by $\{1, \ell(f)\}$ are traceless. It is easy enough to see that, after some cancellations of the form $\ell(f)^*\ell(f) = \|f\|^2$, the only such elements are of the form
\[ \ell(f) \cdots \ell(f)\ell(f)^* \cdots \ell(f)^*, \]
where the product above is non empty. The conclusion follows by considering arbitrary alternating products of the above type for $f$ and $g$, and by noting that whenever $\ell(f)^*\ell(g)$ appears, the end result is zero; hence, the $*$-algebras generated by $\ell(f)$ and $\ell(g)$ are free. The conclusion follows.

2.3. The partial transpose of random quantum states. We study here the asymptotical eigenvalue distribution of the partial transposition of random quantum states. Here, by random, we mean the probability distributions on $M_{d,n}(\mathbb{C})$ known as the induced measures, which were introduced in [ZS01]. A random quantum state $\rho$ having the induced measure of parameters $(d,s)$ is simply a normalized Wishart matrix of the same parameters, see also [Nec07, ŽPNC11]

\[ \rho = \frac{W}{\text{Tr} W} = \frac{GG^*}{\text{Tr}(GG^*)}. \]

However, since we are just interested in the positivity of certain operators, it is enough to work with the cone of positive semidefinite matrices (and the Wishart matrices) instead of working with quantum states. Recall that the cone of separable matrices is defined as

\[ \mathcal{SEP}_{d,n} = \{ A \in M_{d,n} : A = \sum_i B_i \otimes C_i, \text{ where } B_i, C_i \geq 0 \} \subseteq \mathcal{PSD}_{d,n}. \]

The question whether a given mixed quantum state is separable or entangled has been proven to be an NP-hard one [Gur03]. To circumvent this worst-case intractability, entanglement criteria are used. These are efficiently computable conditions which are necessary for separability; in other words, an entanglement criterion is a (usually convex) super-set $\mathcal{X}_d$ of the set of separable states, for which the membership problem is efficiently solvable (see [AS15] for the number of such criteria needed to obtain a good approximation of the set of separable states). As in the previous section, from a probabilistic point of view, estimating the probability that a random quantum state (sampled from the induced ensemble) is an element of $\mathcal{X}_d$ is central.

In what follows we shall tackle this problem for one entanglement criterion in the framework of thresholds. Given a family $G_d \subseteq \mathcal{PSD}_d$ of convex cones, a pair of functions $(s'_d, s''_d)$ is called a threshold for the family $G_d$ if the following two properties are satisfied:

1. If $W_d$ is a sequence of Wishart random matrices of parameters $(d,s_d)$ with $s_d \geq s''_d$, then
\[ \lim_{d \to \infty} \mathbb{P}[W_d \in G_d] = 1. \]

2. If $W_d$ is a sequence of Wishart random matrices of parameters $(d,s_d)$ with $s_d \leq s'_d$, then
\[ \lim_{d \to \infty} \mathbb{P}[W_d \in G_d] = 0. \]
Let us start with the most used example, the positive partial transpose criterion (PPT). The PPT criterion has been introduced by Peres in [Per96]: if a positive semidefinite matrix \( A \in M_d \otimes M_n \) is separable, then
\[
A^\Gamma := [\text{id} \otimes \text{transp}](A) \geq 0.
\]
Note that the positivity of \( A^\Gamma \) is equivalent to the positivity of \( A^\Gamma = [\text{transp} \otimes \text{id}](A) \), so it does not matter on which tensor factor the transpose application acts. We denote by \( \mathcal{PPT}_{d,n} \) the PPT cone
\[
\mathcal{PPT}_{d,n} := \{ A \in M_{dn} : A^\Gamma \geq 0 \} \supseteq \mathcal{SEP}_{d,n}.
\]
This necessary condition for separability has been shown to be also sufficient for qubit-qubit and qubit-qutrit systems \((dn \leq 6)\) in [HHH96]; the result was a simple consequence of the fact that all the positive application from \( M_2 \) to \( M_{2,3} \) are decomposable. These non trivial facts are due to Woronowocz [Wor76]. The PPT criterion for random quantum states has first been studied in the balanced case and the relation to meanders. The analytic results in the following proposition are from [Aub12] (in the balanced case) and from [BN13] (in the unbalanced case); see also [FS13] for some improvements.

**Proposition 2.6.** Consider a sequence \( W_d \in M_{dn_d} \) of random Wishart matrices of parameters \((dn_d, cdn_d)\), where \( n_d \) is a function of \( d \) and \( c \) is a positive constant.

In the balanced regime \( n_d = d \), the (properly rescaled) empirical eigenvalue distribution of the matrices \( W_d^\Gamma \) converges to a semicircular measure \( \mu_{SC(1,1/c)} \) of mean 1 and variance \( 1/c \), see (2.1). In particular, the threshold for the sets \( \mathcal{PPT}_{d,d} \) (\( d \to \infty \)) is \( c_0 = 4 \).

In the unbalanced regime \( n_d = n \) fixed, the (properly rescaled) empirical eigenvalue distribution of the matrices \( d^{-1}W_d^\Gamma \) converges to a free difference of free Poisson distributions (see Section 2.1 for the definitions)
\[
\pi_{cn(n+1)/2} \square \pi_{cn(n-1)/2}.
\]
In particular, the threshold for the sets \( \mathcal{PPT}_{d,n} \) (\( n \) fixed, \( d \to \infty \)) is
\[
c_0 = 2 + 2\sqrt{1 - \frac{1}{n^2}}.
\]

**Proof.** We are going to sketch the proof of the convergence result in the unbalanced case; for the balanced case, see [Aub12] and for the threshold in the unbalanced case, see [BN13, Section 6].

Using again the graphical Wick formula, one can find the following expression for the (unnormalized) moments of \( W_d^\Gamma \):
\[
\mathbb{E} \text{Tr}[(W_d^\Gamma)^p] = \sum_{\sigma \in S_p} s^{\#\sigma} d^{\#(\gamma^{-1}\sigma)\sigma} n^{\#(\gamma\sigma)}.
\]
Using the fact that, for every noncrossing partition \( \sigma \in NC(p) \), denoting by \( e(\sigma) \) the number of blocks of even size of \( \sigma \), we have \( 1 + e(\sigma) = \#(\sigma\gamma) \), we arrive at the formula
\[
\mathbb{E}(dn)^{-1} \text{Tr}[(d^{-1}W_d^\Gamma)^p] \sim \sum_{\sigma \in NC(p)} n^{\#\sigma + e(\sigma)} c^{\#\sigma}
\sim \sum_{\sigma \in NC(p)} \prod_{\beta \in \sigma} c_{\sigma}^{1+1|\beta| \text{ is even}}
\sim \sum_{\sigma \in NC(p)} \prod_{\beta \in \sigma} \left( \frac{cn(n+1)}{2} + \frac{cn(n-1)}{2} (-1)^{|\beta|} \right).
\]
We can now identify the free difference of free Poisson operators using the free cumulant approach of [NS06]: the free cumulant of order \( p \) of the limiting measure is
\[
\frac{cn(n+1)}{2} + \frac{cn(n-1)}{2} (-1)^{|\beta|}.
\]
Remark 2.7. The computation of the limiting distribution of in the unbalanced case performed above was done using the method of moments. A more general approach, allowing to answer the same question for general maps and general matrix distributions, was provided in [ANV16] using operator valued free probability theory.

Remark 2.8. The value of the threshold in the theorem above has a practical significance: if one considers a random pure quantum state on \( H = \mathbb{C}^d \otimes \mathbb{C}^n \otimes \mathbb{C}^{cdn} \), takes the partial trace on the third subsystem, and the partial transposition on the second subsystem, then the resulting matrix is positive semidefinite if \( c > c_0 \), and has negative eigenvalues if \( c < c_0 \), with large probability as \( n \) is fixed and \( d \to \infty \).

3. Lecture 3 — Random quantum channels and their minimum output entropy

3.1. Quantum channels, minimum output entropies, additivity. In Quantum Information Theory, a quantum channel is the most general transformation of a quantum system. Quantum channels generalize the unitary evolution of isolated quantum systems to open quantum systems. Mathematically, we recall that a quantum channel is a linear completely positive trace preserving map \( \Phi \) from \( \mathcal{M}_n(\mathbb{C}) \) to itself. The trace preservation condition is necessary since quantum channels should map density matrices to density matrices. The complete positivity condition can be stated as

\[
\forall d \geq 1, \quad \Phi \otimes I_d : \mathcal{M}_{nd}(\mathbb{C}) \to \mathcal{M}_{nd}(\mathbb{C}) \text{ is a positive map.}
\]

The following three characterizations of quantum channels turn out to be very useful; they are due to Stinespring [Sti55] and Choi [Cho75].

**Proposition 3.1.** A linear map \( \Phi : \mathcal{M}_n(\mathbb{C}) \to \mathcal{M}_n(\mathbb{C}) \) is a quantum channel if and only if one of the following three equivalent conditions holds.

1. (Stinespring dilation) There exists a finite dimensional Hilbert space \( \mathcal{K} = \mathbb{C}^d \), a density matrix \( Y \in \mathcal{M}_d^{1,+}(\mathbb{C}) \) and an unitary operator \( U \in \mathcal{U}_{nd} \) such that

\[
\Phi(X) = \text{Tr}_{\mathcal{K}}[U(X \otimes Y)U^*], \quad \forall X \in \mathcal{M}_n(\mathbb{C}).
\]

2. (Kraus decomposition) There exists an integer \( k \) and matrices \( L_1, \ldots, L_k \in \mathcal{M}_n(\mathbb{C}) \) such that

\[
\Phi(X) = \sum_{i=1}^k L_iXL_i^*, \quad \forall X \in \mathcal{M}_n(\mathbb{C}).
\]

and

\[
\sum_{i=1}^k L_i^*L_i = I_n.
\]

3. (Choi matrix) The following matrix, called the Choi matrix of \( \Phi \)

\[
\mathcal{M}_{n^2}(\mathbb{C}) \ni C_\Phi = [\text{id} \otimes \Phi](\Omega_d) = \sum_{i,j=1}^n E_{ij} \otimes \Phi(E_{ij})
\]

is positive-semidefinite and satisfies \([\text{id} \otimes \text{Tr}](C_\Phi) = I_n\).

It can be shown that the dimension of the ancilla space \( \mathcal{K} \) in the Stinespring dilation theorem can be chosen \( d = \dim \mathcal{K} = n^2 \) and that the state \( Y \) can always be considered to be a rank one projector. A similar result holds for the number of Kraus operators: one can always find a decomposition with \( k = n^2 \) operators.

As in classical information theory [Sha48], entropic quantities play a very important role in quantum information theory. We define next the quantities of interest for the current work. Let
\( \Delta_k = \{ x \in \mathbb{R}^k_+ \mid \sum_{i=1}^{k} x_i = 1 \} \) be the \((k-1)\)-dimensional probability simplex. For a positive real number \( p \in (0, 1) \cup (1, \infty) \), define the Rényi entropy of order \( p \) of a probability vector \( x \in \Delta_k \) to be
\[
H_p(x) = \frac{1}{1-p} \log \sum_{i=1}^{k} x_i^p.
\]
Since \( \lim_{p \to 1} H_p(x) \) exists, we define the Shannon entropy of \( x \) to be this limit, namely:
\[
H(x) = H_1(x) = - \sum_{i=1}^{k} x_i \log x_i.
\]
We also define the values for the parameters \( p = 0, \infty \):
\[
H_0(x) = \log \# \{ i : x_i \neq 0 \}
\]
\[
H_\infty(x) = - \log \| x \|_\infty.
\]
We extend these definitions to density matrices by functional calculus: for \( \rho \in \mathcal{M}_n^{1+}(\mathbb{C}) \), we put
\[
H_0(\rho) = \log \text{rk}(\rho)
\]
\[
H_p(\rho) = \frac{1}{1-p} \log \text{Tr} \rho^p \quad p \in (0, 1) \cup (1, \infty)
\]
\[
H(\rho) = H_1(\rho) = - \text{Tr} \rho \log \rho
\]
\[
H_\infty(\rho) = - \log \| \rho \|_\infty.
\]
Of special interest for the computation of capacities of quantum channels to transmit classical information are the following quantities, called the minimum output entropies of the channel. As the Rényi entropies, they are indexed by some positive real parameter \( p \)
\[
H_p^{\min}(\Phi) = \min_{\rho \in \mathcal{M}_n^{1+}(\mathbb{C})} H_p(\Phi(\rho)). \tag{3.3}
\]
The following theorem summarizes some of the most important breakthroughs in quantum information theory in the last decade. It is based in particular on the papers [Has09, HW08], and concerns the minimum output entropies of quantum channels, defined in (3.3). The result came as a surprise to the community, since additivity (i.e. equality in (3.4)) was shown to hold for many examples of quantum channels.

**Theorem 3.2.** For every \( p \in [1, \infty] \), there exist quantum channels \( \Phi \) and \( \Psi \) such that
\[
H_p^{\min}(\Phi \otimes \Psi) < H_p^{\min}(\Phi) + H_p^{\min}(\Psi). \tag{3.4}
\]
Except for some particular cases \( (p > 4.79, [WH02] \text{ and } p > 2, [GHP10]) \), the proof of this theorem uses the random method, i.e. the channels \( \Phi, \Psi \) are random channels, and the above inequality occurs with non-zero probability. At this moment, we are not aware of any explicit, non-random choices for \( \Phi, \Psi \) in the case \( 1 \leq p \leq 2 \).

The additivity property for the minimum output entropy \( H^{\min}(\cdot) \) was related in [Sho04] to the additivity of another important entropic quantity, the Holevo quantity
\[
\chi(\Phi) = \max_{\{p_i, X_i\}} \left[ H \left( \sum_i p_i \Phi(X_i) \right) - \sum_i p_i H(\Phi(X_i)) \right].
\]
The regularized Holevo quantity provides [Hol98, SW97] the classical capacity of a quantum channel \( \Phi \), i.e. the maximum rate at which classical information can be reliably sent through the noisy channel. The importance of the additivity question stems mainly from the difficulty in computing the above regularized quantity. Indeed, if additivity holds, there would be no need for regularization, and the classical capacity of the channel \( \Phi \) would simply be equal to the Holevo (or one-shot) capacity \( \chi(\Phi) \).
3.2. Random quantum channels. As discussed before, all the known counter-examples from Theorem 3.2, at least in the case $p = 1$ or $p$ close to 1, come from random constructions. We will define now what we mean by a random quantum channel. Note that the model described below is just one of many possible. It has the merit of providing the largest violations for the inequality in Theorem 3.2, as well as the lowest output dimensions, see [BCN16].

We consider the probability distributions on quantum channels (i.e. trace preserving, completely positive maps)

$$\Phi : \mathcal{M}_d(\mathbb{C}) \to \mathcal{M}_k(\mathbb{C}), \quad \Phi(X) = [\text{id}_k \otimes \text{Tr}_n](V XV^*),$$

where $V : \mathbb{C}^d \to \mathbb{C}^k \otimes \mathbb{C}^n$ is a random Haar isometry; note that $\Phi$ is a quantum channel, by the Stinespring dilation result from Proposition 3.1. By a random Haar isometry we mean the unique invariant probability measure on the Stiefel manifold $\mathcal{V}_d(\mathbb{C}^{nk})$. A random isometry $V$ can also be seen as a truncation of a random Haar unitary $U \in \mathcal{U}_{nk}$.

It has been shown by many authors [FK10, ASW11, CN11b, BCN16] that sequences $(\Phi_n)$ of random quantum channels violate asymptotically, with probability one, the additivity relation from Theorem 3.2; most of the examples use the asymptotic regime $d_n \sim tkn$ for suitable fixed output dimension $k$ and input space ratio $t \in (0,1)$.

All the counter examples use the so-called Hayden-Winter trick, that is letting $\Psi = \Phi$ and lower bounding the left hand side of the additivity relation (3.4) by the output of the maximally entangled state $\Omega_2$. The details of this lower bound can be found in [CN10b], we shall not discuss them here. We study the behavior of the right hand side of (3.4) in the next subsection.

3.3. Strong convergence and the (t)-norm. Strong convergence and the (t)-norm

We have seen in Section 2.2 how freeness appears naturally in the full Fock space setting. Another very important situation where freeness manifests itself is the asymptotic theory of random matrices. The following result was one of Voiculescu’s breakthroughs [Voi98].

**Theorem 3.3.** Let $(A_n)$ and $(B_n)$ be sequences of $n \times n$ matrices such that $A_n$ and $B_n$ converge in distribution (with respect to $n^{-1}\text{Tr}$) for $n \to \infty$. Furthermore, let $(U_n)$ be a sequence of Haar unitary $n \times n$ random matrices. Then, $A_n$ and $U_nB_nU_n^*$ are asymptotically free for $n \to \infty$.

If $A_n, B_n$ are matrices of size $n$, whose spectra converge towards $\mu_a, \mu_b$, the spectrum of $A_n + U_nB_nU_n^*$ converges to $\mu_a \boxplus \mu_b$; for the definition of the free additive convolution $\boxplus$, see Section 2.1. Similarily, If $A_n, B_n$ are matrices of size $n$ such that $A_n \geq 0$, whose spectra converge towards $\mu_a, \mu_b$, the spectrum of $A_n^{1/2}U_nB_nU_n^*A_n^{1/2}$ converges to $\mu_a \boxtimes \mu_b$; the operation $\boxtimes$ is called the free multiplicative convolution.

Actually, if the matrices $A_n$ and $B_n$ have well-behaved eigenvalues, not only do the moments of $A_n + U_nB_nU_n^*$ converge to those of $a + b$ with $a, b$ free, but we also have a norm convergence, called strong convergence

$$\text{almost surely, } \lim_{n \to \infty} \|A_n + U_nB_nU_n^*\|_{\infty} = \|a + b\|.$$ 

Note that in the above setting, we need to consider $a, b \in (A, \tau)$ a $C^*$ non-commutative probability space. This result has been shown for GUE matrices in [HT05], and further extended in [Mal12, CM14].

Let us now consider an example, the truncation of random matrices. Let $P_n \in \mathcal{M}_n$ a projection of rank $n/2$; its eigenvalues are 0 and 1, with multiplicity $n/2$. Hence, the distribution of $P_n$ converges, when $n \to \infty$, to the Bernoulli probability measure $\frac{1}{2}\delta_0 + \frac{1}{2}\delta_1$. Let $C_n \in \mathcal{M}_{n/2}$ be the top $n/2 \times n/2$ corner of $U_nP_nU_n^*$, with $U_n$ a Haar random unitary matrix. Up to zero blocks, $C_n = Q_n(U_nP_nU_n^*)Q_n$, where $Q_n$ is the diagonal orthogonal projection on the first $n/2$ coordinates of $\mathbb{C}^n$. The distribution of $Q_n$ converges to $\frac{1}{2}\delta_0 + \frac{1}{2}\delta_1$. A direct computation in free probability...
theory tells us that the distribution of $C_n$ will converge to
$$\left[\frac{1}{2}\delta_0 + \frac{1}{2}\delta_1\right] \otimes \left[\frac{1}{2}\delta_0 + \frac{1}{2}\delta_1\right] = \frac{1}{\pi \sqrt{x(1-x)}} 1_{[0,1]}(x)dx,$$
which is the arcsine distribution.

\begin{figure}[h]
\centering
\includegraphics[width=0.5\textwidth]{histogram.png}
\caption{Histogram of eigenvalues of a truncated randomly rotated projector of relative rank 1/2 and size $n = 4000$; in red, the density of the arcsine distribution.}
\end{figure}

We introduce now a new norm, which will play a crucial role in computing MOE for random quantum channels.

**Definition 3.4.** For a positive integer $k$, embed $\mathbb{R}^k$ as a self-adjoint real subalgebra $\mathcal{R}$ of a $C^*$-ncps $(\mathcal{A}, \tau)$, so that $\tau(x) = (x_1 + \cdots + x_k)/k$. Let $p_t$ be a projection of rank $t \in (0, 1]$ in $\mathcal{A}$, free from $\mathcal{R}$. On the real vector space $\mathbb{R}^k$, we introduce the following norm, called the $(t)$-norm:
$$\|x\|_{(t)} := \|p_t x p_t\|_{\infty},$$
where the vector $x \in \mathbb{R}^k$ is identified with its image in $\mathcal{R}$.

One can show that $\|\cdot\|_{(t)}$ is indeed a norm, which is permutation invariant. When $t > 1 - 1/k$, $\|\cdot\|_{(t)} = \|\cdot\|_{\infty}$ on $\mathbb{R}^k$, and we can show that $\lim_{t \to 0^+} \|x\|_{(t)} = k^{-1} \sum_i x_i$.

For vectors $x$ with non-negative elements, $\|x\|_{(t)}$ is the right end of the support of the probability measure
$$\left[\sum_i \delta_{x_i}\right] \otimes \left[(1-t)\delta_0 + t\delta_1\right].$$

Let us now look at an example of a computation for the $t$-norm, in the case where $x$ has just two components taking two values.

**Theorem 3.5.** In $\mathbb{C}^n$, choose at random according to the Haar measure two independent subspaces $V_n$ and $V'_n$ of respective dimensions $q_n \sim sn$ and $q'_n \sim tn$ where $s, t \in (0, 1]$. Let $P_n$ (resp. $P'_n$) be the orthogonal projection onto $V_n$ (resp. $V'_n$). Then,
$$\lim_{n} \|P_n P'_n P_n\|_{\infty} = \varphi(s, t) = \sup \text{supp}((1-s)\delta_0 + s\delta_1) \otimes ((1-t)\delta_0 + t\delta_1),$$
with
$$\varphi(s, t) = \begin{cases} s + t - 2st + 2\sqrt{st(1-s)(1-t)} & \text{if } s + t < 1; \\
1 & \text{if } s + t \geq 1. \end{cases} \tag{3.6}$$

Hence, we can compute
$$\|1, \cdots, 1, 0, \cdots, 0\|_{(t)} = \varphi(j/k, t).$$
3.4. The MOE of a (large) random quantum channel. From now on, we shall abuse notation: recall that we are interested in random isometries $V : \mathbb{C}^d \to \mathbb{C}^k \otimes \mathbb{C}^n$. Since the quantities $H_p^{\min}$ only depend on the range of $V$, also write $V = \text{ran} V$. For a subspace $V \subset \mathbb{C}^k \otimes \mathbb{C}^n$, define
\[ H_p^{\min}(V) = \min_{y \in V, \|y\|=1} H_p(y), \]
the minimal $p$-entropy of vectors in $V$; for a channel as in (3.5), we have
\[ H_p^{\min}(V) = H_p^{\min}(\Phi). \]

For a subspace $V \subset \mathbb{C}^k \otimes \mathbb{C}^n$ of dimension $\dim V = d$, define the set eigen-/singular values or Schmidt coefficients
\[ K_V = \{ \lambda(x) : x \in V, \|x\| = 1 \}. \]
Our goal is to understand $K_V$, and thus, the particular statistic $H_p^{\min}(V)$. The set $K_V$ is a compact subset of the ordered probability simplex $\Delta_k^1$, having the following properties
- Local invariance: $K_{(U_1 \otimes U_2)V} = K_V$, for unitary matrices $U_1 \in \mathcal{U}(k)$ and $U_2 \in \mathcal{U}(n)$.
- Monotonicity: if $V_1 \subset V_2$, then $K_{V_1} \subset K_{V_2}$.
- Recovering minimum entropies:
\[ H_p^{\min}(\Phi) = H_p^{\min}(V) = \min_{\lambda \in K_V} H_p(\lambda). \]

Example 3.6. The anti-symmetric subspace provides the (explicit) counter-example for the additivity of the $p$-Rényi entropy [GHP10]. Let $k = n$ and put $V = \Lambda^2(\mathbb{C}^n)$. The subspace $V$ is almost half of the total space: $\dim V = n(n-1)/2$. Antisymmetric vectors in $V$ are typically
\[ V \ni x = \frac{1}{\sqrt{2}} (e \otimes f - f \otimes e). \]
Since singular values of vectors in $V$ come in pairs, the least entropy vector in $V$ is as above, with $e \perp f$ and $H(x) = \log 2$. Thus, $H^{\min}(V) = \log 2$ and one can show that
\[ K_V = \{ (\lambda_1, \lambda_1, \lambda_2, \lambda_2, \ldots) \in \Delta_n : \lambda_i \geq 0, \sum_i \lambda_i = 1/2 \}. \]

Problem 3.7. Find explicit (i.e. non-random) examples of subspaces $V \subset \mathbb{C}^k \otimes \mathbb{C}^n$ with
- large $\dim V$;
- large $H^{\min}(V)$.

We state now the main result of this section, a characterization of the set $K_V$, for a random isometry $V$ of large size [BCN12]. Recall that we are interested in random isometries/subspaces in the following asymptotic regime: $k$ fixed, $n \to \infty$, and $d \sim tkn$, for a fixed parameter $t \in (0, 1)$.

Theorem 3.8. For a sequence of uniformly distributed random subspaces $V_n$, the set $K_{V_n}$ of singular values of unit vectors from $V_n$ converges (almost surely, in the Hausdorff distance) to a deterministic, convex subset $K_{k,t}$ of the probability simplex $\Delta_k$
\[ K_{k,t} := \{ \lambda \in \Delta_k \mid \forall x \in \Delta_k, \langle \lambda, x \rangle \leq \|x\|_2 \}. \]
The theorem above allows for the computation of the asymptotic behavior of channel statistics which are related to the output set of random quantum channels. The problem is reduced to the analogous question on the deterministic set $K_{k,t}$. The case of the MOE was treated in [BCN16], where the following corollary was proved.

Corollary 3.9. For all $p \geq 1$,
\[ \lim_{n \to \infty} H_p^{\min}(\Phi) = \min_{\lambda \in K_{k,t}} H_p(\lambda) = H_p(a, b, b, \ldots, b). \]
where $a, b$ do not depend on $p$, $b = (1 - a)/(k - 1)$ and $a = \varphi(1/k, t)$, where the function $\varphi$ was defined in (3.6).

**Proof of Theorem 3.8.** The statement in the proof concerns the duals of the set $K_{k,t}$, so we are going to consider how far does the set $K_{V_n}$ extend into some given direction $a \in \Delta^n$.

Let us start by considering the direction $a = (1, 0, \ldots, 0)$. We would like to compute the largest maximal singular value $\max_{x \in V, \|x\|=1} \lambda_1(x)$ of vectors from the subspace $V$?

$$
\max_{x \in V, \|x\|=1} \lambda_1(x) = \max_{x \in V, \|x\|=1} \|[\text{id}_k \otimes \text{Tr}_n] P_x\|_{\infty} = \max_{x \in V, \|x\|=1} \max_{y \in \mathbb{C}^k, \|y\|=1} \text{Tr} \left( [(\text{id}_k \otimes \text{Tr}_n) P_x] \cdot P_y \right) = \max_{x \in V, \|x\|=1} \max_{y \in \mathbb{C}^k, \|y\|=1} \text{Tr} \left( P_x \cdot P_y \otimes I_n \right) = \max_{y \in \mathbb{C}^k, \|y\|=1} \|P_V \cdot P_y \otimes I_n \cdot P_V\|_{\infty}.
$$

For fixed $y$, $P_V$ and $P_y \otimes I_n$ are independent projectors of relative ranks $t$ and $1/k$ respectively. Thus, almost surely,

$$
\|P_V \cdot P_y \otimes I_n \cdot P_V\|_{\infty} \to \|(1 - t)\delta_0 + t\delta_1\| \otimes \|(1 - 1/k)\delta_0 + 1/k\delta_1\| = \varphi(t, 1/k) = \|(1, 0, \ldots, 0)\|_{(t)}.
$$

The computation above shows that the asymptotic behavior of $\|P_V \cdot P_y \otimes I_n \cdot P_V\|_{\infty}$ is independent of $y$. We can thus take the max over $y$ at no cost, by considering a finite net of $y$'s, since $k$ is fixed.

To get the full result $\limsup_{n \to \infty} K_{V_n} \subset K_{k,t}$, we have to consider $(\lambda, a)$ for all directions $a$; the computations are similar.

The inclusion $\liminf_{n \to \infty} K_{V_n} \supset K_{k,t}$ is much easier, and follows from the convergence in distribution/moments.

\[\square\]

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