High Probability Analysis of the Condition Number of Sparse Polynomial Systems

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Abstract

Let $f := (f^1, \ldots, f^n)$ be a polynomial system with an arbitrary fixed n-tuple of supports and random coefficients. Our main result is an upper bound on the probability that the condition number of f in a region U is larger than $1/\varepsilon$. The bound depends on an integral of a differential form on a toric manifold and admits a simple explicit upper bound when the Newton polytopes (and underlying variances) are all identical.

We also consider polynomials with real coefficients and give bounds for the expected number of real roots and (restricted) condition number. Using a Kähler geometric framework throughout, we also express the expected number of roots of f inside a region U as the integral over U of a certain **mixed volume** form, thus recovering the classical mixed volume when $U = (\mathbb{C}^*)^n$.

Key words: mixed volume, condition number, polynomial systems, sparse, random.

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1 Introduction

From the point of view of numerical analysis, it is not only the number of complex solutions of a polynomial system which make it hard to solve numerically but the sensitivity of its roots to small perturbations in the coefficients. This is formalized in the condition number, $\mu(f,\zeta)$ (cf. Definition 3 and Section 1.1 below), which dates back to work of Alan Turing (Tur36). In essence, $\mu(f,\zeta)$ measures the sensitivity of a solution ζ to perturbations in a problem f, and a large condition number is meant to imply that f is intrinsically hard to solve numerically. Such analysis of numerical conditioning, while having been applied for decades in numerical linear algebra (see, e.g., (Dem97)), has only been applied to computational algebraic geometry toward the end of the twentieth century (see, e.g., (SS93b)).

Here we use Kähler geometry to analyze the numerical conditioning of sparse polynomial systems, thus setting the stage for more realistic complexity bounds for the numerical solution of polynomial systems. Our bounds generalize some earlier results of Kostlan (Kos93) and Shub and Smale (SS96) on the more restrictive dense case, and also yield new formulae for the expected number of roots (real and complex) in a region. The appellations "sparse" and "dense" respectively refer to either (a) taking into account the underlying monomial term structure or (b) ignoring this finer structure and simply working with degrees of polynomials.

Since many polynomial systems occurring in practice have rather restricted monomial term structure, sparsity is an important consideration and we therefore strive to state our complexity bounds in terms of this refined information. In particular, it is now understood that algorithmic algebraic geometry can be sped up tremendously by taking advantage of sparsity, e.g., (Roj00; Som02; MPR03; Lec03). Here, we demonstrate analogous improvements in a context closer to numerical analysis.

To give the flavor of our results, let us first make some necessary definitions. We must first formalize the spaces of polynomial systems we work with and how we measure perturbations in the spaces of problems and solutions.

Definition 1 *Given any finite subset $A \subset \mathbb{Z}^n$, let $\mathcal{F}_{\mathbb{C}}(A)$ (resp. $\mathcal{F}_{\mathbb{R}}(A)$) denote the vector space of all polynomials in $\mathbb{C}[x_1,\ldots,x_n]$ (resp. $\mathbb{R}[x_1,\ldots,x_n]$) of the form $\sum_{a\in A} c_a x^a$ where the notation $x^a := x_1^{a_1}\cdots x_n^{a_n}$ is understood. For any finite subsets $A_1,\ldots,A_n \subset \mathbb{Z}^n$ we then let $\mathcal{A} := (A_1,\ldots,A_n)$ and $\mathcal{F}_{\mathbb{C}}(\mathcal{A}) := \mathcal{F}_{\mathbb{C}}(A_1)\times\cdots\times\mathcal{F}_{\mathbb{C}}(A_n)$ (resp. $\mathcal{F}_{\mathbb{R}}(\mathcal{A}) := \mathcal{F}_{\mathbb{R}}(A_1)\times\cdots\times\mathcal{F}_{\mathbb{R}}(A_n)$).* $\diamond$

The n-tuple $\mathcal{A}$ will thus govern our notion of sparsity as well as the perturbations allowed in the coefficients of our polynomial systems. It is then easy to speak of random polynomial systems and the distance to the nearest degenerate system. Recall that a **degenerate** root of f is simply a root of f having Jacobian of rank $<n$.

Definition 2 *By a* **complex (resp. real) random sparse polynomial system** *we will mean a choice of* $\mathcal{A} := (A_1, \ldots, A_n)$ *and an assignment of a probability measure to each* $\mathcal{F}_{\mathbb{C}}(A_i)$ *(resp.* $\mathcal{F}_{\mathbb{R}}(A_i)$*) as follows: endow* $\mathcal{F}_{\mathbb{C}}(A_i)$ *(resp.* $\mathcal{F}_{\mathbb{R}}(A_i)$*) with an independent complex (resp. real) Gaussian distribution having mean* $\mathbf{O}$ *and a (positive definite and diagonal) variance matrix* C_i*. Finally, let the* **discriminant variety***,* $\Sigma(\mathcal{A})$*, denote the closure of the set of all* $f \in \mathcal{F}_{\mathbb{C}}(\mathcal{A})$ *(resp.* $f \in \mathcal{F}_{\mathbb{R}}(\mathcal{A})$*) with a degenerate root and define* $\mathcal{F}_{\zeta}(\mathcal{A}) := \{f \in \mathcal{F}_{\mathbb{C}}(\mathcal{A}) \mid f(\zeta) = \mathbf{O}\}$ *(resp.* $\mathcal{F}_{\zeta}(\mathcal{A}) := \{f \in \mathcal{F}_{\mathbb{R}}(\mathcal{A}) \mid f(\zeta) = \mathbf{O}\}$*) and* $\Sigma_{\zeta}(\mathcal{A}) := \mathcal{F}_{\zeta}(\mathcal{A}) \cap \Sigma(\mathcal{A})$*.* $\diamond$

A classic result from numerical linear algebra (the Eckart-Young Theorem (GvL96)) relates the condition number of a linear system $Ax=b$ to the distance of A to the set of singular matrices. Our first main result below generalizes this to sparse polynomial systems.

Definition 3 *If* $A = \{a_1, \ldots, a_m\}$ *has cardinality* m*, and* $C = [c_{ij}]$ *is any positive definite diagonal* $m \times m$ *matrix, then we define a norm on* $\mathcal{F}_{\mathbb{C}}(A)$ *by* $\|g\|_C^2 := \sum_{i=1}^m c_{ii}^{-1} |g_i|^2$ *where* $g(x) = \sum_{i=1}^m g_i x^{a_i}$*. Also, if* m_i *is the cardinality of* A_i *and* C_i *is an* $m_i \times m_i$ *positive definite diagonal matrix for all* i*, then we define a norm on* $\mathcal{F}_{\mathbb{C}}(\mathcal{A})$ *by* $\|f\|^2 := \sum\limits_{i=1}^n \|f^i\|_{C_i}^2$*, and a metric* $d_{\mathbb{P}}$ *on the product of projective spaces* $\mathbb{P}(\mathcal{F}_{\mathbb{C}}(\mathcal{A})) := \mathbb{P}(\mathcal{F}_{\mathbb{C}}(A_1)) \times \cdots \times \mathbb{P}(\mathcal{F}_{\mathbb{C}}(A_n))$ *by* $d_{\mathbb{P}}(f, g) := \sum\limits_{i=1}^n \min\limits_{\lambda \in \mathbb{C}^*} \frac{\|f^i - \lambda g^i\|_{C_i}}{\|f^i\|_{C_i}}$*.* [3] *Finally, if* $C_1 = \cdots = C_n = C$ *and* $A_1 = \cdots = A_n = A$*, then we define the* **condition number of** f **at a root** ζ*,* $\mu(f, \zeta)$*, as the operator norm* $\left\| DG|_f \right\| := \max\limits_{\|\dot{f}\|=1} \left\| DG|_f \, \dot{f} \right\|_C$*, where* G *is the unique branch of the implicit function satisfying* $G(f) = \zeta$ *and* $g(G(g)) = \mathbf{O}$ *for all* g *sufficiently near* f*, and* DG *is the derivative of* G*.* [4] $\diamond$

Theorem 1 *If* $\mathcal{A} := (\underbrace{A, \ldots, A}_{n})$ *then* $\mu(f, \zeta) = \frac{1}{d(f, \Sigma_{\zeta}(\mathcal{A}))}$*. Furthermore,*

$$\text{Prob}\left[\mu(f,\zeta) \geq \tfrac{1}{\varepsilon} \text{ for some root } \zeta \in (\mathbb{C}^*)^n \text{ of } f\right] \leq n^3(n^2+n)(n^2+n-1)(n+1)\text{Vol}(A)\varepsilon^4,$$

[3] Each of the terms in the sum defining $d_{\mathbb{P}}$ corresponds to the square of the sine of the Fubini (or angular) distance between f^i and g^i. Therefore, $d_{\mathbb{P}}$ is never larger than the Hermitian distance between points in $\mathcal{F}_{\mathbb{C}}(\mathcal{A})$, but is a correct first-order approximation of the distance when $g \to f$ in $\mathbb{P}(\mathcal{F}_{\mathbb{C}}(\mathcal{A}))$ (compare with (BCSS98, Ch. 12)). Note also that we implicitly used the natural embedding of $\mathbb{P}(\mathcal{F}_{\mathbb{C}}(A_i))$ into the unit hemisphere of $\mathcal{F}_{\mathbb{C}}(A_i)$.

[4] We set the condition number $\mu(f, \zeta) := +\infty$ in the event that Df does not have full rank and G thus fails to be uniquely defined.

where f is a complex random sparse polynomial system, m denotes the number of points in A, and $\mathrm{Vol}(A)$ *denotes the volume of the convex hull of A (normalized so that* $\mathrm{Vol}(\mathbf{O}, e_1, \ldots, e_n) = 1$*).*

The above theorem is in fact a simple corollary of two much more general theorems (Theorems 4 and 5) which also include as a special case an analogous result of Shub and Smale in the dense case (BCSS98, Thm. 1, Pg. 237). We also note that theorems such as the one above are natural precursors to explicit bounds on the number of steps required for a homotopy algorithm (SS93b) to solve f. We will pursue the latter topic in a future paper. Indeed, one of our long term goals is to provide a rigorous and explicit complexity analysis of the numerical homotopy algorithms for sparse polynomial systems developed by Verschelde et. al. (VVC94), Huber and Sturmfels (HS95), and Li and Li (LL01).

The framework underlying our first main theorem involves Kähler geometry, which is the intersection of Riemannian metrics and symplectic and complex structures on manifolds. For technical reasons, we will mainly work with **logarithmic coordinates**. That is, we will let $\mathcal{T}^n$ be the n-fold product of cylinders $(\mathbb{R} \times (\mathbb{R} \bmod 2\pi))^n \subset \mathbb{C}^n$, and use coordinates $p + iq := (p_1 + iq_1, \ldots, p_n + iq_n) \in \mathcal{T}^n$ to stand for a root $\zeta := \exp(p + iq) := (e^{p_1+iq_1}, \ldots, e^{p_n+iq_n})$ of f. Roots with zero coordinates can be handled by then working in a suitable **toric compactification** and this is made precise in Section 2.[5] On a more concrete level, we can give new formulae for the expected number of roots of f in a region U.

Theorem 2 *Let $A_1, \ldots, A_n$ be finite subsets of $\mathbb{Z}^n$ and $U \subseteq \mathcal{T}^n$ be a measurable region. Pick positive definite diagonal variance matrices $C_1, \ldots, C_n$ and consider a complex random polynomial system as in Definition 2, for some $(A_1, C_1, \ldots, A_n, C_n)$. Then there are natural real 2-forms $\omega_{A_1}, \ldots, \omega_{A_n}$ on $\mathcal{T}^n$ such that the expected number of roots of f in $\exp U \subseteq (\mathbb{C}^*)^n$ is exactly*

$$\frac{(-1)^{n(n-1)/2}}{\pi^n} \int_U \omega_{A_1} \wedge \cdots \wedge \omega_{A_n} \quad .$$

In particular, when $U = (\mathbb{C}^*)^n$, the above expression is exactly the mixed volume of the convex hulls of $A_1, \ldots, A_n$ (normalized so that the mixed volume of n standard n-simplices is 1).

See (BZ88; SY93) for the classical definition of mixed volume and its main properties. The result above generalizes the famous connection between root counting and mixed volumes discovered by David N. Bernshtein (Ber76).

[5] The idea of working with roots of polynomial systems in logarithmic coordinates seems to be extremely classical, yet it gives rise to interesting and surprising connections (see the discussions in (MZ01a; MZ01b; Vir01)).

The special case of unmixed systems with identical coefficient distributions ($A_1 = \cdots = A_n$, $C_1 = \cdots = C_n$) recovers a particular case of Theorem 8.1 in (EK95). However, comparing Theorem 2 and (EK95, Theorem 8.1), this is the only overlap since neither theorem generalizes the other. The very last assertion of Theorem 2 (for uniform variance $C_i = I$ for all i) was certainly known to Gromov (Gro90), and a version of Theorem 2 was known to Kazarnovskii (Kaz81, p. 351) and Khovanskii (Kho91, Prop. 1, Sec. 1.13). In (Kaz81), the supports A_i are even allowed to have complex exponents. However, uniform variance is again assumed. His method may imply this special case of Theorem 2, but the indications given in (Kaz81) were insufficient for us to reconstruct a proof. Also, there is some intersection with a result by Passare and Rullgård (Theorem 5 in (PR00) and Theorem 20 in (Mikh01)). However, the latter result is about a more restrictive choice of the domain U and a more general class of functions (holomorphic instead of polynomial) under a different averaging process.

As a consequence of our last result, we can also give a coarse estimate on the expected number of real roots in a region.

Theorem 3 *Let U be a measurable subset of $\mathbb{R}^n$ with Lebesgue volume $\lambda(U)$. Then, following the notation above, suppose instead that f is a real random polynomial system. Then the average number of real roots of f in $\exp U \subset \mathbb{R}^n_+$ is bounded above by*

$$(4\pi^2)^{-n/2}\sqrt{\lambda(U)}\sqrt{\int_{(p,q)\in U\times[0,2\pi)^n} (-1)^{n(n-1)/2}\omega_{A_1}\wedge\cdots\wedge\omega_{A_n}}.$$

This bound is of interest when n and U are fixed, in which case the expected number of positive real roots grows as the square root of the mixed volume.

1.1 Stronger Results Via Mixed Metrics

Note that while $\mathcal{F}_{\mathbb{C}}(\mathcal{A})$ admits a natural Hermitian structure, the solution-space $\mathcal{T}^n$ admits n different natural Hermitian structures (one from each support A_i, as we shall see in the next section). This complicates defining $\mu(f, p+iq)$ for general $\mathcal{A}$. Nevertheless, we can extend our earlier provisional definition to many additional cases, and give useful bounds on the resulting generalized condition number.

Theorem 4 (Condition Number Theorem) *If $(p,q) \in \mathcal{T}^n$ is a non-degenerate root of f then, following the preceding notation,*

$$\max_{\|\dot{f}\|\leq 1}\min_i \|DG_f\dot{f}\|_{C_i} \leq \frac{1}{d_{\mathbb{P}}(f,\Sigma_{(p,q)})} \leq \max_{\|\dot{f}\|\leq 1}\max_i \|DG_f\dot{f}\|_{C_i} \quad .$$

In particular, if $A_1 = \cdots = A_n$ and $C_1 = \cdots = C_n$, then

$$\max_{\|\dot{f}\|\leq 1} \min_i \|DG_f \dot{f}\|_{C_i} = \max_i \max_{\|\dot{f}\|\leq 1} \|DG_f \dot{f}\|_{C_i} = \frac{1}{d_{\mathbb{P}}(f, \Sigma_{(p,q)})}$$

and we can define $\boldsymbol{\mu}(f;(p,q))$ to be any of the three last quantities.

This generalizes (BCSS98, Thm. 3, Pg. 234) which is essentially equivalent to the last assertion above, in the special case where A_i is an n-column matrix whose rows $\{A_i^\alpha\}_\alpha$ consist of all partitions of d_i into n non-negative integers and $C_i = \mathrm{Diag}_\alpha\left(\frac{(d_i-1)!}{(A_i)_1^\alpha!(A_i)_2^\alpha!\cdots(A_i)_n^\alpha!(d_i-\sum_{j=1}^n (A_i)_j^\alpha)!}\right)$ — in short, the case where one considers complex random polynomial systems with f^i a degree d_i polynomial and the underlying probability measure is invariant under a natural action of the unitary group $U(n+1)$ on the space of roots. The last assertion of Theorem 4 also bears some similarity to Theorem D of (Ded96) where the notion of metric is considerably loosened to give a statement which applies to an even more general class of equations. However, our philosophy is radically different: we consider the inner product in $\mathcal{F}_{\mathbb{C}}(\mathcal{A})$ as the starting point of our investigation and we do **not** change the metric in the fiber $\mathcal{F}_{(p,q)}$. Theorem 4 thus gives us some insight about reasonable intrinsic metric structures on $\mathcal{T}^n$.

Recall that T_pM denotes the tangent space at p of a manifold M. In view of the preceding theorem, we can define a restricted condition number with respect to any measurable sub-region $U \subset \mathcal{T}^n$ as follows:

Definition 4 *We let $\boldsymbol{\mu}(f;U) := \frac{1}{\min_{(p,q)\in U} d_{\mathbb{P}}(f,\Sigma_{(p,q)})}$. Also, via the natural $GL(n)$-action on $T_{(p,q)}\mathcal{T}^n$ defined by $(\dot{p},\dot{q}) \mapsto (L\dot{p}, L\dot{q})$ for any $L \in GL(n)$, we define the* **mixed dilation** *of the tuple $(\omega_{A_1}, \cdots, \omega_{A_n})$ as:*

$$\kappa(\omega_{A_1}, \cdots, \omega_{A_n}; (p,q)) := \min_{L\in GL(n)} \max_i \frac{\max_{\|u\|=1} (\omega_{A_i})_{(p,q)}(Lu, JLu)}{\min_{\|u\|=1} (\omega_{A_i})_{(p,q)}(Lu, JLu)} \quad ,$$

where $J : T\mathcal{T}^n \longrightarrow T\mathcal{T}^n$ is the canonical complex structure (MS98; CCL99) of $\mathcal{T}^n$. Finally, we define $\kappa_U := \sup_{(p,q)\in U} \kappa(\omega_{A_1}, \cdots, \omega_{A_n}; (p,q))$, provided the supremum exists, and $\kappa_U := +\infty$ otherwise. $\diamond$

We can then bound the expected number of roots with condition number $\boldsymbol{\mu} > \varepsilon^{-1}$ on U in terms of the mixed volume form, the mixed dilation κ_U and the expected number of ill-conditioned roots in the **linear case**. The linear case corresponds to the point sets and variance matrices below:

$$A_i^{\mathrm{Lin}} = \begin{bmatrix} 0 & \cdots & 0 \\ 1 & & \\ & \ddots & \\ & & 1 \end{bmatrix} \qquad\qquad C_i^{\mathrm{Lin}} = \begin{bmatrix} 1 & & & \\ & 1 & & \\ & & \ddots & \\ & & & 1 \end{bmatrix}$$

where both matrices have n columns.

Theorem 5 (High Probability Estimate of the Condition Number) *Let $\nu^{\mathrm{Lin}}(n, \varepsilon)$ be the probability that a complex random system of n polynomials in n variables has condition number larger than ε^{-1}. Let $\nu^{\mathcal{A}}(U, \varepsilon)$ be the probability that $\boldsymbol{\mu}(f, U) > \varepsilon^{-1}$ for a complex random polynomial system f with supports $A_1, \cdots, A_n$ and variance matrices $C_1, \cdots, C_n$. Then*

$$\nu^{\mathcal{A}}(U, \varepsilon) \leq \frac{\int_U \bigwedge \omega_{A_i}}{\int_U \bigwedge \omega_{A_i^{\mathrm{Lin}}}} \ \nu^{\mathrm{Lin}}(n, \sqrt{\kappa_U}\varepsilon).$$

Our final main result concerns the distribution of the real roots of a real random polynomial system. Let $\nu_{\mathbb{R}}(n, \varepsilon)$ be the probability that a real random linear system of n polynomials in n variables has condition number larger than ε^{-1}.

Theorem 6 *Let $A = A_1 = \cdots = A_n$ and $C = C_1 = \cdots = C_n$ and let $U \subseteq \mathbb{R}^n$ be measurable. Let f be a real random polynomial system. Then,*

$$\mathrm{Prob}\left[\boldsymbol{\mu}(f, U) > \varepsilon^{-1}\right] \leq E(U) \ \nu_{\mathbb{R}}(n, \varepsilon)$$

where $E(U)$ is the expected number of real roots on U.

Note that $E(U)$ depends on C, so even if we make $U = \mathbb{R}^n$ we may still obtain a bound depending on C. Shub and Smale showed in (SS93a) that the expected number of real roots in the dense case (with a particular choice of probability measure) is exactly the square root of the expected number of complex roots. The sparse analogue of this result seems hard to prove even in the general unmixed case: Explicit formulæ for the unmixed case are known only in certain special cases, e.g., certain systems of bounded multi-degree (Roj96; McL02). Hence our last theorem can be interpreted as another step toward a fuller generalization.

2 Symplectic Geometry and Polynomial Systems

2.1 Some Basic Definitions and Examples

For the standard definitions and properties of symplectic structures, complex structures, Riemannian manifolds, and Kähler manifolds, we refer the reader to (MS98; CCL99). A treatment focusing on toric manifolds can be found in (Gui94, Appendix A). We briefly review a few of the basics before moving on to the proofs of our theorems.

Definition 5 (Kähler manifolds) *Let M be a complex manifold, with complex structure J and a strictly positive symplectic $(1,1)$-form ω on M (considered as a real manifold). We then call the triple (M, ω, J) a* **Kähler manifold**. $\diamond$

Example 1 (Affine Space) *We identify $\mathbb{C}^m$ with $\mathbb{R}^{2m}$ and use coordinates $Z^i = X^i + \sqrt{-1}Y^i$. The* **canonical** *2-form $\omega_Z = \sum_{i=1}^m dX_i \wedge dY_i$ makes $\mathbb{C}^m$ into a symplectic manifold.*

The natural complex structure J is just the multiplication by $\sqrt{-1}$. The triple $(\mathbb{C}^m, \omega_Z, J)$ is a Kähler manifold. $\diamond$

Example 2 (Projective Space) *Projective space $\mathbb{P}^{m-1}$ admits a* **canonical** *2-form defined as follows. Let $Z = (Z^1, \cdots, Z^m) \in (\mathbb{C}^m)^*$, and let $[Z] = (Z^1 : \cdots : Z^m) \in \mathbb{P}^{m-1}$ be the corresponding point in $\mathbb{P}^{m-1}$. The tangent space $T_{[Z]}\mathbb{P}^{m-1}$ may be modeled by $Z^\perp \subset T_Z\mathbb{C}^m$. Then we can define a two-form on $\mathbb{P}^{m-1}$ by setting:*

$$\omega_{[Z]}(u, v) = \|Z\|^{-2} \omega_Z(u, v) \quad ,$$

where it is assumed that u and v are orthogonal to Z. The latter assumption tends to be quite inconvenient, and most people prefer to pull $\omega_{[Z]}$ back to $\mathbb{C}^m$ by the canonical projection $\pi : Z \mapsto [Z]$. It is standard to write the pull-back $\tau = \pi^*\omega_{[Z]}$ as:

$$\tau_Z = -\frac{1}{2} dJ^* d \ \frac{1}{2} \log \|Z\|^2 \quad ,$$

using the notation $d\eta = \sum_i \frac{\partial \eta}{p_i} \wedge dp_i + \frac{\partial \eta}{q_i} \wedge dq_i$, and where J^* denotes the pull-back by J.

Projective space also inherits the complex structure from $\mathbb{C}^m$. Then $\omega_{[Z]}$ is a strictly positive $(1,1)$-form. The corresponding metric is called **Fubini-Study** *metric in $\mathbb{C}^m$ or $\mathbb{C}^{m-1}$.* $\diamond$

Remark 1 *Some authors prefer to write $\sqrt{-1}\partial\bar{\partial}$ instead of $-\frac{1}{2}dJ^*d$. They then assume the notation $\partial\eta = \sum_i \frac{\partial\eta}{Z_i} \wedge dZ_i$ and $\bar{\partial}\eta = \sum_i \frac{\partial\eta}{\bar{Z}_i} \wedge d\bar{Z}_i$, and then* *write τ_Z as*

$$\frac{\sqrt{-1}}{2}\left(\frac{\sum_i dZ_i \wedge d\bar{Z}_i}{\|Z\|^2} - \frac{\sum_i Z_i d\bar{Z}_i \wedge \sum_j \bar{Z}_j dZ_j}{\|Z\|^4}\right) \quad . \diamond$$

Example 3 (Toric Kähler Manifolds from Point Sets) *Let A be any $m \times n$ matrix with integer entries whose row vectors have n-dimensional convex hull and let C be any positive definite diagonal $m \times m$ matrix. Define the map $\hat{V}_A$ from $\mathbb{C}^n$ into $\mathbb{C}^m$ by*

$$\hat{V}_A : z \mapsto C^{1/2} \begin{bmatrix} z^{A^1} \\ \vdots \\ z^{A^m} \end{bmatrix} .$$

We can also compose with the projection into projective space to obtain a slightly different map $V_A = \pi \circ \hat{V}_A : \mathbb{C}^n \to \mathbb{P}^{m-1}$ defined by $V_A : z \mapsto [\hat{V}_A(z)]$. When C is the identity, the Zariski closure of the image of V_A is called the **Veronese variety** *and the map V_A is called the* **Veronese embedding**. *Note that V_A is not defined for certain values of z, e.g., $z = 0$. Those values comprise the* **exceptional set** *which is a subset of the coordinate hyper-planes.*

There is then a natural symplectic structure on the closure of the image of V_A, given by the restriction of the Fubini-Study 2-form τ: We will see below (Lemma 1) that by our assumption on the convex hull of the rows of A, we have that DV_A is of rank n for $z \in (\mathbb{C}^*)^n$. Thus, we can pull-back this structure to $(\mathbb{C}^*)^n$ by $\Omega_A = V_A^* \tau$. Also, we can pull back the complex structure of $\mathbb{P}^{m-1}$, so that Ω_A becomes a strictly positive $(1,1)$-form. Therefore, the matrix A defines a Kähler manifold $((\mathbb{C}^*)^n, \Omega_A, J)$. $\diamond$

The reason we introduced C in the definition of $\hat{V}_A$ is as follows: if f denotes also the row-vector of the **scaled** coefficients of f, then $f(z) = \sum_a f_a (C_a)^{1/2} z^a = f\hat{V}_A(z)$ (the sum being over the row vectors of A). This way, the 2-norm of the row vector f is also the norm of the polynomial f in $(F_A, \| \cdot \|_{C^{-1}})$. A random normal polynomial with variance matrix C corresponds to a random normal row vector f with unit variance.

Example 4 (Toric Manifolds in Logarithmic Coordinates) *For any matrix A as in the previous example, we can pull-back the Kähler structure of $((\mathbb{C}^*)^n, \Omega_A, J)$ to obtain another Kähler manifold $(\mathcal{T}^n, \omega_A, J)$. (Actually, it is the same object in logarithmic coordinates, minus points at "infinity".) An equivalent definition is to pull back the Kähler structure of the Veronese variety by $\hat{v}_A \stackrel{\text{def}}{=} \hat{V}_A \circ \exp$. $\diamond$*

Remark 2 *The Fubini-Study metric on $\mathbb{C}^m$ was constructed by applying the operator $-\frac{1}{2} dJ^* d$ to a certain convex function (in our case, $\frac{1}{2} \log \|Z\|^2$). This is a general standard way to construct Kähler structures. In (Gro90), it is* *explained how to associate a (non-unique) convex function to any convex body, thus producing an associated Kähler metric.* $\diamond$

For the record, we state explicit formulæ for several of the invariants associated to the Kähler manifold $(\mathcal{T}^n, \omega_A, J)$. First of all, the function $g_A = g \circ \hat{v}_A$ is precisely:

Formula 2.1.1: The canonical Integral g_A (or **Kähler potential**) of the convex set associated to A

$$g_A(p) := \frac{1}{2} \log \Big((\exp(A \cdot p))^T \, C \, (\exp(A \cdot p)) \Big)$$

The terminology **integral** is borrowed from mechanics, and it refers to the invariance of g_A under a $[0, 2\pi)^n$-action. Also, the gradient of g_A is called the **momentum map**. Recall that the Veronese embedding takes values in projective space. We will use the following notation: $v_A(p) = \hat{v}_A(p)/\|\hat{v}_A(p)\|$. This is independent of the representative of equivalence class $v_A(p)$. Now, let $v_A(p)^2$ mean coordinate-wise squaring and $v_A(p)^{2T}$ be the transpose of $v_A(p)^2$. The gradient of g_A is then:

Formula 2.1.2: The Momentum Map associated to A

$$\nabla g_A = v_A(p)^{2T} A$$

Formula 2.1.3: Second derivative of g_A

$$D^2 g_A = 2 D v_A(p)^T D v_A(p)$$

We also have the following formulae:

Formula 2.1.4: The symplectic 2-form associated to A:

$$(\omega_A)_{(p,q)} = \frac{1}{2} \sum_{ij} (D^2 g_A)_{ij} dp_i \wedge dq_j$$

Formula 2.1.5: Hermitian structure of $\mathcal{T}^n$ associated to A:

$$(\langle u, w \rangle_A)_{(p,q)} = u^H (\frac{1}{2} D^2 g_A)_p w$$

In general, the function v_A goes from $\mathcal{T}^n$ into projective space. Therefore, its derivative is a mapping

$$(Dv_A)_{(p,q)} : T_{(p,q)}\mathcal{T}^n \to T_{v_A(p+q\sqrt{-1})}\mathbb{P}^{m-1} \simeq \hat{v}_A(p+q\sqrt{-1})^\perp \subset \mathbb{C}^m \quad .$$

For convenience, we will write this derivative as a mapping into $\mathbb{C}^m$, with range $\hat{v}_A(p+q\sqrt{-1})^\perp$. Let P_v be the projection operator

$$P_v = I - \frac{1}{\|v\|^2} v v^H \quad .$$

We then have the following formula.

Formula 2.1.6: Derivative of v_A

$$(Dv_A)_{(p,q)} = P_{\hat{v}_A(p+q\sqrt{-1})} \mathrm{Diag}\left(\frac{\hat{v}_A(p+q\sqrt{-1})}{\|\hat{v}_A(p+q\sqrt{-1}\|} \right) A$$

Lemma 1 *Let A be a matrix with non-negative integer entries, such that* $\mathrm{Conv}(A)$ *has dimension n. Then $(Dv_A)_p$ (resp. $(Dv_A)_{p+iq}$) is injective, for all $p \in \mathbb{R}^n$ (resp. for all $p+iq \in \mathbb{C}^n$).*

Proof: We prove only the real case (the complex case is analogous). The conclusion of this Lemma can fail only if there are $p \in \mathbb{R}^n$ and $u \neq 0$ with $(Dv_A)_p u = 0$. This means that

$$P_{v_A(p)} \mathrm{diag}(v_A)_p A u = 0 \quad .$$

This can only happen if $\mathrm{diag}(v_A)_p Au$ is in the space spanned by $(v_A)_p$, or, equivalently, Au is in the space spanned by $(1, 1, \cdots, 1)^T$. This means that all the rows a of A satisfy $au = \lambda$ for some λ. Interpreting a row of A as a vertex of $\mathrm{Conv}(A)$, this means that $\mathrm{Conv}(A)$ is contained in the affine plane $\{a : au = \lambda\}$ ■

An immediate consequence of Formula 2.1.6 is:

Lemma 2 *Let $f \in \mathcal{F}_A$ and $(p,q) \in \mathcal{T}^n$ be such that $f \cdot \hat{v}_A(p + q\sqrt{-1}) = 0$. Then,*

$$f \cdot (Dv_A)_{(p,q)} = \frac{1}{\|\hat{v}_A(p,q)\|} f \cdot (D\hat{v}_A)_{(p,q)}.$$

In other words, when $(f \circ \exp)(p+q\sqrt{-1})$ vanishes, Dv_A and $D\hat{v}_A$ are the same up to scaling. Noting that the Hermitian metric can be written $(\langle u, w\rangle_A)_{(p,q)} = u^h Dv_A(p,q)^H Dv_A(p,q) w$, we also obtain the following formula.

Formula 2.1.7: Volume element of $(\mathcal{T}^n, \omega_A, J)$

$$d\mathcal{T}^n_A = \det\left(\frac{1}{2}\ D^2 g_A(p)\right)\ dp_1 \wedge \cdots \wedge dp_n \wedge dq_1 \wedge \cdots \wedge dq_n$$

2.2 Toric Actions and the Momentum Map

The **momentum map**, also called **moment map**, was introduced in its modern formulation by Smale (Sma70) and Souriau (Sou70). The reader may consult one of the many textbooks in the subject (such as Abraham and Marsden (AM78) or McDuff and Salamon (MS98)) for a general exposition (See also the discussion at the end of (MR02)).

In this section we instead follow the point of view of Gromov (Gro90). The main results in this section are the two propositions below.

Proposition 1 *The momentum map ∇g_A maps $\mathcal{T}^n$ onto the interior of* $\mathrm{Conv}(A)$. *When ∇g_A is restricted to the real n-plane $[q = 0] \subset \mathcal{T}^n$, this mapping is a bijection.*

This would appear to be a particular case of the Atiyah-Guillemin-Sternberg theorem (Ati82; GS82). However, technical difficulties prevent us from directly applying this result here.[6]

Proposition 2 *The momentum map ∇g_A is a volume-preserving map from the manifold $(\mathcal{T}^n, \omega_A, J)$ into* $\mathrm{Conv}(A)$, *up to a constant, in the following sense: if U is a measurable region of* $\mathrm{Conv}(A)$, *then*

$$\mathrm{Vol}\left((\nabla g_A)^{-1}(U)\right) = (2\pi)^n\ \mathrm{Vol}(U) \quad .$$

[6] The Atiyah-Guillemin-Sternberg Theorem applies to compact symplectic manifolds and the implied compactification of $\mathcal{T}^n$ may have singularities.

Proof of Proposition 2: Consider the mapping

$$\begin{aligned} M: \quad \mathcal{T}^n &\to \mathrm{Conv}(A) \times \mathbb{T}^n \\ (p, q) &\mapsto (\nabla g_A(p), q) \end{aligned} .$$

Since we assume $\dim \mathrm{Conv}(A) = n$, we can apply Proposition 1 and conclude that M is a diffeomorphism.

The pull-back of the canonical symplectic structure in $\mathbb{R}^{2n}$ by M is precisely ω_A, because of Formulæ 2.1.3 and 2.1.4. Diffeomorphisms with that property are called **symplectomorphisms**. Since the volume form of a symplectic manifold depends only of the canonical 2-form, symplectomorphisms preserve volume. So we are done. ■

Before proving Proposition 1, we will need the following result about convexity which has been attributed to Legendre. (See also (Gro90, Convexity Theorem 1.2) and a generalization in (Avr76, Th. 5.1).)

Legendre's Theorem *If f is convex and of class $\mathcal{C}^2$ on $\mathbb{R}^n$, then the closure of the image $\{\nabla f_r : r \in \mathbb{R}^n\}$ in $\mathbb{R}^n$ is convex.*

By replacing f by g_A, we conclude that the image of the momentum map ∇g_A is convex.

Proof of Proposition 1: The momentum map ∇g_A maps $\mathcal{T}^n$ onto the interior of $\mathrm{Conv}(A)$. Indeed, let $a = A^\alpha$ be a row of A, associated to a vertex of $\mathrm{Conv}(A)$. Then there is a direction $v \in \mathbb{R}^n$ such that

$$a \cdot v = \max_{x \in \mathrm{Conv}(A)} x \cdot v$$

for some unique a.

We claim that $a \in \overline{\nabla g_A(\mathbb{R}^n)}$. Indeed, let $x(t) = v_A(tv)$, t a real parameter. If b is another row of A,

$$e^{a \cdot tv} = e^{ta \cdot v} \gg e^{tb \cdot v} = e^{b \cdot tv}$$

as $t \to \infty$. We can then write $\hat{v}_A(tv)^{2T}$ as:

$$\hat{v}_A(tv)^{2T} = \begin{bmatrix} \vdots \\ e^{ta \cdot v} \\ \vdots \end{bmatrix}^T C \mathrm{Diag} \begin{bmatrix} \vdots \\ e^{ta \cdot v} \\ \vdots \end{bmatrix} .$$

Since C is positive definite, $C_{\alpha\alpha} > 0$ and

$$\lim_{t\to\infty} v_A(tv)^{2T} = \lim_{t\to\infty} \frac{\hat{v}_A(tv)^{2T}}{\|\hat{v}_A(tv)\|^2} = \mathrm{e}_a^T \frac{C_{\alpha\alpha}}{C_{\alpha\alpha}} = \mathrm{e}_a^T \quad ,$$

where e_a is the unit vector in $\mathbb{R}^m$ corresponding to the row a. It follows that $\lim_{t\to\infty} \nabla g_A(tv) = a$

When we set $q = 0$, we have $\det D^2 g_A \neq 0$ on $\mathbb{R}^n$, so we have a local diffeomorphism at each point $p \in \mathbb{R}^n$. Assume that $(\nabla g_A)_p = (\nabla g_A)_{p'}$ for $p \neq p'$. Then, let $\gamma(t) = (1-t)p + tp'$. The function $t \mapsto (\nabla g_A)_{\gamma(t)}\gamma'(t)$ has the same value at 0 and at 1, hence by Rolle's Theorem its derivative must vanish at some $t^* \in (0,1)$.

In that case,

$$(D^2 g_A)_{\gamma(t^*)}(\gamma'(t^*), \gamma'(t^*)) = 0$$

and since $\gamma'(t^*) = p' - p \neq 0$, $\det D^2 g_A$ must vanish in some $p \in \mathbb{R}^n$. This contradicts Lemma 1. ∎

2.3 The Condition Matrix

Recall that the **evaluation map**, $ev_{\mathcal{A}}$, is defined as follows:

$$\begin{aligned} ev_{\mathcal{A}} : \quad & \mathcal{F} \times \mathcal{T}^n \to \mathbb{C}^n \\ & ((f^1, \dots, f^n), p + iq) \mapsto (f^1(\exp(p+iq)), \dots, f^n(\exp(p+iq))). \end{aligned}$$

Following (BCSS98), we look at the linearization of the implicit function $p + q\sqrt{-1} = G(f)$ for the equation $ev_{\mathcal{A}}(f, p + q\sqrt{-1}) = 0$.

Definition 6 *The* **condition matrix** *of ev at $(f, p + q\sqrt{-1})$ is*

$$DG = D_{\mathcal{T}^n}(ev)^{-1} D_{\mathcal{F}}(ev) \quad ,$$

where $\mathcal{F} = \mathcal{F}_{A_1} \times \cdots \times \mathcal{F}_{A_n}$.

Above, $D_{\mathcal{T}^n}(ev)$ is a linear operator from an n-dimensional complex space into $\mathbb{C}^n$, while $D_{\mathcal{F}}(ev)$ goes from an $(m_1 + \cdots + m_n)$-dimensional complex space into $\mathbb{C}^n$.

Lemma 3 *If $p + iq \in \mathcal{T}^n$ and $f(\exp(p+iq)) = \mathbf{O}$ then*

$$\det\left(DG\,DG^H\right)^{-1} dp_1 \wedge dq_1 \wedge \cdots \wedge dp_n \wedge dq_n = (-1)^{n(n-1)/2} \bigwedge_{i=1}^{n} \sqrt{-1} f^i \cdot (Dv_{A_i})_{(p,q)} dp \wedge \bar{f}^i \cdot (Dv_{A_i})_{(p,-q)} dq.$$

Note that although $f^i \cdot (Dv_{A_i})_{(p,q)} dp$ is a complex-valued form, each wedge $f^i \cdot (Dv_{A_i})_{(p,q)} dp \wedge \bar{f}^i \cdot (Dv_{A_i})_{(p,-q)} dq$ is a real-valued 2-form.

Proof of Lemma 3: We compute:

$$D_{\mathcal{F}}(ev)|_{(p,q)} = \begin{bmatrix} \sum_{\alpha=1}^{m_1} \hat{v}_{A_1}^{\alpha}(p+q\sqrt{-1})df_{\alpha}^{1} \\ \vdots \\ \sum_{\alpha=1}^{m_n} \hat{v}_{A_n}^{\alpha}(p+q\sqrt{-1})df_{\alpha}^{n} \end{bmatrix} \quad ,$$

and hence

$$D_{\mathcal{F}}(ev)D_{\mathcal{F}}(ev)^H = \text{diag } \|\hat{v}_{A_i}\|^2.$$

Also,

$$D_{\mathcal{T}^n}(ev) = \begin{bmatrix} f^1 \cdot D\hat{v}_{A_1} \\ \vdots \\ f^n \cdot D\hat{v}_{A_n} \end{bmatrix}.$$

Therefore,

$$\det\left(DG_{(p,q)}DG^H_{(p,q)}\right)^{-1} = \left|\det \begin{bmatrix} f^1 \cdot \frac{1}{\|\hat{v}_{A_1}\|} D\hat{v}_{A_1} \\ \vdots \\ f^n \cdot \frac{1}{\|\hat{v}_{A_n}\|} D\hat{v}_{A_n} \end{bmatrix}\right|^2.$$

We can now use Lemma 2 to conclude the following:

Formula 2.3.1: Determinant of the Condition Matrix

$$\det\left(DG_{(p,q)}DG^H_{(p,q)}\right)^{-1} = \left|\det \begin{bmatrix} f^1 \cdot Dv_{A_1} \\ \vdots \\ f^n \cdot Dv_{A_n} \end{bmatrix}\right|^2$$

We can now write the same formula as a determinant of a block matrix:

$$\det\left(DG_{(p,q)}DG^H_{(p,q)}\right)^{-1} = \det\begin{bmatrix} f^1\cdot Dv_{A_1} & \\ \vdots & \\ f^n\cdot Dv_{A_n} & \\ & \bar{f}^1\cdot D\bar{v}_{A_1} \\ & \vdots \\ & \bar{f}^n\cdot D\bar{v}_{A_n} \end{bmatrix}$$

and replace the determinant by a wedge. The factor $(-1)^{n(n-1)/2}$ comes from replacing $dp_1\wedge\cdots\wedge dp_n\wedge dq_1\wedge\cdots\wedge dq_n$ by $dp_1\wedge dq_1\wedge\cdots\wedge dp_n\wedge dq_n$. ■

We are now ready to prove our main theorems.

3 The Proofs of Theorems 1–6

We first prove that Theorem 1 follows from Theorem 4. Then we will prove our remaining main theorems in the following order: 2, 4, 5, 3, 6.

3.1 The Proof of Theorem 1

The first assertion, modulo an exponential change of coordinates and using the multi-projective metric $d_{\mathbb{P}}(\cdot,\cdot)$, follows immediately from Theorem 4.

As for the rest of Theorem 1, Theorem 5 implies that our probability in question is bounded above by $\frac{\int_U \bigwedge \omega_{A_i}}{\int_U \bigwedge \omega_{A_i^{\text{Lin}}}}\,\nu^{\text{Lin}}(n,\sqrt{\kappa_U}\varepsilon)$. In particular, Theorem 2 (and the fact that mixed volume reduces to ordinary volume for unmixed n-tuples) immediately implies that this bound reduces to $\text{Vol}(A)\nu^{\text{Lin}}(n,\varepsilon)$. So, by another application of Theorem 4, it suffices to prove the inequality

$$\text{Prob}[d_{\mathbb{P}}(f,\Sigma_{(p,q)})<\varepsilon]\leq n^3(n+1)\text{Vol}(A)(n^2+n)(n^2+n-1)\varepsilon^4,$$

for the **linear** case.

To prove the latter inequality, recall that by the definition of the multi-projective distance $d_{\mathbb{P}}(\cdot,\cdot)$, we have the following equality:

$$d_{\mathbb{P}}(f,\Sigma_{(p,q)})^2 = \min_{\substack{g\in\Sigma_{(p,q)}\\ \lambda\in(\mathbb{C}^*)^n}} \sum_{i=1}^n \frac{\|f^i-\lambda_i g^i\|^2}{\|f^i\|^2}.$$

So let g be so that the above minimum is attained. Without loss of generality, we may scale the g^i so that $\lambda_1 = \cdots = \lambda_n = 1$. In that case,

$$d_{\mathbb{P}}(f, \Sigma_{(p,q)})^2 = \sum_{i=1}^{n} \frac{\|f^i - g^i\|^2}{\|f^i\|^2} \geq \frac{\sum_{i=1}^n \|f^i - g^i\|^2}{\sum_{j=1}^n \|f^j\|^2}.$$

We are then in the setting of (BCSS98, pp. 248–250) where we identify our linear f with a normally distributed $(n+1)\times n$ complex matrix. By (BCSS98, Rem. 2, Pg. 250), the right-hand side in the above inequality is then precisely the Frobenius distance[7] $d_F(f, \Sigma_{(p,q)})$. So it follows that

$$\operatorname{Prob}\left[d_{\mathbb{P}}(f, \Sigma_{(p,q)}) < \varepsilon\right] \leq \operatorname{Prob}\left[d_{\mathrm{F}}(f, \Sigma_x) < \varepsilon\right]$$

and the last probability is bounded above by $n^3(n+1)\frac{\Gamma(n^2+n)}{\Gamma(n^2+n-2)}\varepsilon^4$ via (BCSS98, Thm. 6, Pg. 254), where Γ denotes Euler's well-known generalization of the factorial function. So Theorem 1 follows. ■

3.2 The Proof of Theorem 2

Using (BCSS98, Theorem 5, Pg. 243) (or Proposition 5, Pg. 31 below), we deduce that the average number of complex roots is:

$$\text{Avg} = \int_{(p,q)\in U} \int_{f\in\mathcal{F}_{(p,q)}} \left(\prod \frac{e^{-\|f^i\|^2/2}}{(2\pi)^{m_i}}\right) \det\left(DG_{(p,q)} DG^H_{(p,q)}\right)^{-1}.$$

By Lemma 3, we can replace the inner integral by a $2n$-form valued integral:

$$\begin{aligned}\text{Avg} = (-1)^{n(n-1)/2} \int_{(p,q)\in U} \int_{f\in\mathcal{F}_{(p,q)}} \bigwedge_i \frac{e^{-\|f^i\|^2/2}}{(2\pi)^{m_i}} f^i \cdot (Dv_{A_i})_{(p,q)} dp \wedge \\ \wedge \bar{f}^i \cdot (Dv_{A_i})_{(p,-q)} dq \quad .\end{aligned}$$

Since the image of Dv_{A_i} is precisely $\mathcal{F}_{A_i,(p,q)} \subset \mathcal{F}_{A_i}$, one can add n extra variables corresponding to the directions $v_{A_i}(p+q\sqrt{-1})$ without changing the integral: we write $\mathcal{F}_{A_i} = \mathcal{F}_{A_i,(p,q)} \times \mathbb{C} v_{A_i}(p+q\sqrt{-1})$. Since $\left(f^i + t v_{A_i}(p+q\sqrt{-1})\right) Dv_{A_i}$

[7] Recall that the Frobenius distance $d_F(A,B)$ between any two $M \times N$ matrices $A := [a_{ij}]$ and $B := [b_{ij}]$ is just $\sqrt{\sum_{ij}(a_{ij} - b_{ij})^2}$. It then makes sense to speak of the Frobenius distance of a matrix to any compact subset of matrix space.

is equal to $f^i Dv_{A_i}$, the average number of roots is indeed:

$$\mathrm{Avg} = (-1)^{n(n-1)/2} \int_{(p,q)\in U} \int_{f\in\mathcal{F}} \bigwedge_i \frac{e^{-\|f^i\|^2/2}}{(2\pi)^{m_i+1}} f^i \cdot (Dv_{A_i})_{(p,q)} dp \wedge \\ \wedge \bar{f}^i \cdot (Dv_{A_i})_{(p,-q)} dq \quad .$$

In the integral above, all the terms that are multiple of $f^i_\alpha \bar{f}^i_\beta$ for some $\alpha \neq \beta$ will cancel out. Therefore,

$$\mathrm{Avg} = (-1)^{n(n-1)/2} \int_{(p,q)\in U} \int_{f\in\mathcal{F}} \bigwedge_i \frac{e^{-\|f^i\|^2/2}}{(2\pi)^{m_i+1}} \sum_\alpha |f^i_\alpha|^2 (Dv_{A_i})^\alpha_{(p,q)} dp \wedge \\ \wedge (Dv_{A_i})^\alpha_{(p,-q)} dq \quad .$$

Now, we apply the integral formula:

$$\int_{x\in\mathbb{C}^m} |x_1|^2 \frac{e^{-\|x\|^2/2}}{(2\pi)^m} = \int_{x_1\in\mathbb{C}} |x_1|^2 \frac{e^{-|x_1|^2/2}}{2\pi} = 2$$

to obtain:

$$\mathrm{Avg} = \frac{(-1)^{n(n-1)/2}}{\pi^n} \int_{(p,q)\in U} \bigwedge \sum_\alpha (Dv_{A_i})^\alpha_{(p,q)} dp \wedge (Dv_{A_i})^\alpha_{(p,-q)} dq \quad .$$

According to formulæ 2.1.3 and 2.1.4, the integrand is just $2^{-n} \bigwedge \omega_{A_i}$, and thus

$$\mathrm{Avg} = \frac{(-1)^{n(n-1)/2}}{\pi^n} \int_U \bigwedge_i \omega_{A_i} = \frac{n!}{\pi^n} \int_U d\mathcal{T}^n \blacksquare$$

3.3 The Proof of Theorem 4

Let $(p,q) \in \mathcal{T}^n$ and let $f \in \mathcal{F}_{(p,q)}$. Without loss of generality, we can assume that f is scaled so that for all i, $\|f^i\| = 1$.

Let $\delta f \in \mathcal{F}_{(p,q)}$ be such that $f + \delta f$ is singular at (p,q), and assume that $\sum \|\delta f^i\|^2$ is minimal. Then, due to the scaling we choose,

$$d_{\mathbb{P}}(f, \Sigma_{(p,q)}) = \sqrt{\sum \|\delta f^i\|^2} \quad .$$

Since $f + \delta f$ is singular, there is a vector $u \neq 0$ such that

$$\begin{bmatrix} (f^1 + \delta f^1) \cdot (D\hat{v}_{A_1})_{(p,q)} \\ \vdots \\ (f^n + \delta f^n) \cdot (D\hat{v}_{A_n})_{(p,q)} \end{bmatrix} u = 0$$

and hence

$$\begin{bmatrix} (f^1 + \delta f^1) \cdot (Dv_{A_1})_{(p,q)} \\ \vdots \\ (f^n + \delta f^n) \cdot (Dv_{A_n})_{(p,q)} \end{bmatrix} u = 0 \quad .$$

This means that

$$\begin{cases} f^1 \cdot Dv_{A_1} u \; = -\delta f^1 \cdot Dv_{A_1} u \\ \qquad \vdots \\ f^n \cdot Dv_{A_n} u = -\delta f^n \cdot Dv_{A_n} u \end{cases} .$$

Let $D(f)$ denote the matrix

$$D(f) \stackrel{\text{def}}{=} \begin{bmatrix} f^1 \cdot (Dv_{A_1})_{(p,q)} \\ \vdots \\ f^n \cdot (Dv_{A_n})_{(p,q)} \end{bmatrix} \quad .$$

Given $v = D(f)\ u$, we obtain:

$$\begin{cases} v_1 \; = -\delta f^1 \cdot Dv_{A_1} D(f)^{-1} v \\ \quad \vdots \\ v_n = -\delta f^n \cdot Dv_{A_n} D(f)^{-1} v \end{cases} \tag{3.3.1}$$

We can then scale u and v, such that $\|v\| = 1$.

Claim 1 *Under the assumptions above, δf^i is colinear to $(Dv_{A_i} D(f)^{-1} v)^H$.*

Proof: Assume that $\delta f^i = g+h$, with g colinear and h orthogonal to $(Dv_{A_i} D(f)^{-1} v)^H$. As the image of Dv_{A_i} is orthogonal to v_{A_i}, g is orthogonal to $v_{A_i}^H$, so $ev(g^i, (p,q)) = 0$ and hence $ev(h^i, (p,q)) = 0$. We can therefore replace δf^i by g without compromising equality (3.3.1). Since $\|\delta f\|$ was minimal, this implies $h = 0$. ■

We obtain now an explicit expression for δf^i in terms of v:

$$\delta f^i = -v_i \frac{\left(D v_{A_i} D(f)^{-1} v\right)^H}{\|D v_{A_i} D(f)^{-1} v\|^2} \quad .$$

Therefore,

$$\|\delta f^i\| = \frac{|v_i|}{\|D v_{A_i} D(f)^{-1} v\|} = \frac{|v_i|}{\|\left(D(f)^{-1} v\right)\|_{A_i}} \quad .$$

So we have proved the following result:

Lemma 4 *Fix v so that $\|v\| = 1$ and let $\delta f \in \mathcal{F}_{(p,q)}$ be such that equation (3.3.1) holds and $\|\delta f\|$ is minimal. Then,*

$$\|\delta f^i\| = \frac{|v_i|}{\|D(f)^{-1} v\|_{A_i}} \quad .$$

Lemma 4 provides an immediate lower bound for $\|\delta f\| = \sqrt{\sum \|\delta f^i\|^2}$: Since

$$\|\delta f^i\| \geq \frac{|v_i|}{\max_j \|D(f)^{-1} v\|_{A_j}} \quad ,$$

we can use $\|v\| = 1$ to deduce that

$$\sqrt{\sum_i \|\delta f^i\|^2} \geq \frac{1}{\max_j \|D(f)^{-1} v\|_{A_j}} \geq \frac{1}{\max_j \|D(f)^{-1}\|_{A_j}} \quad .$$

Also, for any v with $\|v\| = 1$, we can choose δf minimal so that equation (3.3.1) applies. Using Lemma 4, we obtain:

$$\|\delta f^i\| \leq \frac{|v_i|}{\min_j \|D(f)^{-1} v\|_{A_j}} \quad .$$

Hence

$$\sqrt{\sum_i \|\delta f^i\|^2} \leq \frac{1}{\min_j \|D(f)^{-1} v\|_{A_j}} \quad .$$

Since this is true for any v, and $\|\delta f\|$ is minimal for all v, we have

$$\sqrt{\sum_i \|\delta f^i\|^2} \leq \frac{1}{\max_{\|v\|=1} \min_j \|D(f)^{-1}\|_{A_j}}$$

and this proves Theorem 4.

The proof of Theorem 5 is long. We first sketch the idea of the proof. Recall that $\mathcal{F}_{(p,q)}$ is the set of all $f \in \mathcal{F}$ such that $ev(f; p + q\sqrt{-1}) = 0$, and that $\Sigma_{(p,q)}$ is the restriction of the discriminant to the fiber $\mathcal{F}_{(p,q)}$:

$$\Sigma_{(p,q)} \stackrel{\text{def}}{=} \{f \in \mathcal{F}_{(p,q)} : D(f)_{(p,q)} \text{ does not have full rank}\} \quad .$$

The space $\mathcal{F}$ is endowed with a Gaussian probability measure, with volume element

$$\frac{e^{-\|f\|^2/2}}{(2\pi)^{\sum m_i}} d\mathcal{F} \quad ,$$

where $d\mathcal{F}$ is the usual volume form in $\mathcal{F} = (\mathcal{F}_{A_1}, \langle\cdot,\cdot\rangle_{A_1}) \times \cdots \times (\mathcal{F}_{A_n}, \langle\cdot,\cdot\rangle_{A_n})$ and $\|f\|^2 = \sum \|f^i\|^2_{A_i}$. For U a set in $\mathcal{T}^n$, we defined earlier (in the statement of Theorem 5) the quantity:

$$\nu^{\mathcal{A}}(U, \varepsilon) \stackrel{\text{def}}{=} \text{Prob}[\boldsymbol{\mu}(f, U) > \varepsilon^{-1}] = \text{Prob}[\exists (p,q) \in U : d_{\mathbb{P}}(f, \Sigma_{(p,q)}) < \varepsilon] \quad .$$

The naïve idea for bounding $\nu^{\mathcal{A}}(U, \varepsilon)$ is as follows: Let $V(\varepsilon) \stackrel{\text{def}}{=} \{(f, (p,q)) \in \mathcal{F} \times U : ev(f; (p,q)) = 0 \text{ and } d_{\mathbb{P}}(f, \Sigma_{(p,q)}) < \varepsilon\}$. We also define $\pi : V(\varepsilon) \to \mathcal{F}$ as the canonical projection mapping $\mathcal{F} \times U$ to $\mathcal{F}$, and set $\#_{V(\varepsilon)}(f) \stackrel{\text{def}}{=} \#\{(p,q) \in U : (f, (p,q)) \in V(\varepsilon)\}$. Then,

$$\begin{aligned} \nu^{\mathcal{A}}(U, \varepsilon) &= \int_{f \in \mathcal{F}} \chi_{\pi(V(\varepsilon))}(f) \; \frac{e^{-\|f\|^2/2}}{(2\pi)^{\sum m_i}} d\mathcal{F} \\ &\leq \int_{f \in \mathcal{F}} \#_{V(\varepsilon)} \; \frac{e^{-\|f\|^2/2}}{(2\pi)^{\sum m_i}} d\mathcal{F} \end{aligned}$$

with equality in the linear case and when $\epsilon > \sqrt{n}$.

Now we apply the coarea formula (BCSS98, Theorem 5, Pg. 243) to obtain:

$$\nu^{\mathcal{A}}(U, \varepsilon) \leq \int_{(p,q) \in U \subset \mathcal{T}^n} \int_{\substack{f \in \mathcal{F}_{(p,q)} \\ d_{\mathbb{P}}(f, \Sigma_{(p,q)}) < \varepsilon}} \frac{1}{NJ(f; (p,q))} \; \frac{e^{-\|f\|^2/2}}{(2\pi)^{\sum m_i}} d\mathcal{F} \; dV_{\mathcal{T}^n} \quad ,$$

where $dV_{\mathcal{T}^n}$ stands for Lebesgue measure in $\mathcal{T}^n$. Again, in the linear case, we have equality.

We already know from Lemma 3 that

$$1/NJ(f;(p,q)) = \bigwedge_{i=1}^{n} f^i \cdot (Dv_{A_i})_{(p,q)} dp \wedge \bar{f}^i \cdot (D\bar{v}_{A_i})_{(p,q)} dq \quad .$$

We should focus now on the inner integral. In each coordinate space $\mathcal{F}_{A_i}$, we can introduce a new orthonormal system of coordinates (depending on (p,q)) by decomposing:

$$f^i = f^i_{\mathrm{I}} + f^i_{\mathrm{II}} + f^i_{\mathrm{III}} \quad ,$$

where f^i_{I} is the component colinear to $v^H_{A_i}$, f^i_{II} is the projection of f^i to $(\text{range } Dv_{A_i})^H$, and f^i_{III} is orthogonal to f^i_{I} and f^i_{II}.

Of course, $f^i \in (\mathcal{F}_{A_i})_{(p,q)}$ if and only if $f^i_{\mathrm{I}} = 0$.

Also,

$$\bigwedge_{i=1}^{n} f^i \cdot (Dv_{A_i})_{(p,q)} dp \wedge \bar{f}^i \cdot (D\bar{v}_{A_i})_{(p,q)} dq =$$

$$= \bigwedge_{i=1}^{n} f^i_{\mathrm{II}} \cdot (Dv_{A_i})_{(p,q)} dp \wedge \bar{f}^i_{\mathrm{II}} \cdot (D\bar{v}_{A_i})_{(p,q)} dq \quad .$$

It is an elementary fact that

$$d_{\mathbb{P}}(f^i_{\mathrm{II}} + f^i_{\mathrm{III}}, \Sigma_{(p,q)}) \le d_{\mathbb{P}}(f^i_{\mathrm{II}}, \Sigma_{(p,q)}) \quad .$$

It follows that for $f \in \mathcal{F}_{(p,q)}$:

$$d_{\mathbb{P}}(f, \Sigma_{(p,q)}) \le d_{\mathbb{P}}(f_{\mathrm{II}}, \Sigma_{(p,q)}) \quad ,$$

with equality in the linear case. Hence, we obtain:

$$\nu^{\mathcal{A}}(U,\varepsilon) \le \int_{(p,q)\in U \subset \mathcal{T}^n} \int_{\substack{f \in \mathcal{F}_{(p,q)} \\ d_{\mathbb{P}}(f_{\mathrm{II}}, \Sigma_{(p,q)}) < \varepsilon}} \left(\bigwedge_{i=1}^{n} f^i_{\mathrm{II}} \cdot (Dv_{A_i})_{(p,q)} dp \wedge \bar{f}^i_{\mathrm{II}} \cdot (D\bar{v}_{A_i})_{(p,q)} dq \right) \cdot$$

$$\cdot \frac{e^{-\|f^i_{\mathrm{II}} + f^i_{\mathrm{III}}\|^2/2}}{(2\pi)^{\sum m_i}} d\mathcal{F} \; dV_{\mathcal{T}^n} \quad ,$$

with equality in the linear case. We can integrate the $\sum(m_i - n - 1)$ variables f_{III} to obtain:

Proposition 3

$$\nu^{\mathcal{A}}(U,\varepsilon) \leq \int_{(p,q)\in U\subset \mathcal{T}^n} \int_{\substack{f_{\mathrm{I\!I}}\in\mathbb{C}^{n^2} \\ d_{\mathbb{P}}(f_{\mathrm{I\!I}},\Sigma_{(p,q)})<\varepsilon}} \left(\bigwedge_{i=1}^{n} f_{\mathrm{I\!I}}^i \cdot (Dv_{A_i})_{(p,q)} dp \wedge \bar{f}_{\mathrm{I\!I}}^i \cdot (D\bar{v}_{A_i})_{(p,q)} dq \right) \cdot \\ \cdot \frac{e^{-\|f_{\mathrm{I\!I}}^i\|^2/2}}{(2\pi)^{n(n+1)}} \; dV_{\mathcal{T}^n} \quad .$$

with equality in the linear case. ■

3.5 The Proof of Theorem 5

The domain of integration in Proposition 3 makes integration extremely difficult. In order to estimate the inner integral, we will need to perform a change of coordinates.

Unfortunately, the Gaussian in Proposition 3 makes that change of coordinates extremely hard, and we will have to restate Proposition 3 in terms of integrals over a product of projective spaces.

The domain of integration will be $\mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}$. Translating an integral in terms of Gaussians to an integral in terms of projective spaces is not immediate, and we will use the following elementary fact about Gaussians:

Lemma 5 *Let $\varphi : \mathbb{C}^n \to \mathbb{R}$ be $\mathbb{C}^*$-invariant (in the sense of the usual scaling action). Then we can also interpret φ as a function from $\mathbb{P}^{n-1}$ into $\mathbb{R}$, and:*

$$\frac{1}{\mathrm{Vol}(\mathbb{P}^{n-1})} \int_{[x]\in\mathbb{P}^{n-1}} \varphi(x) d[x] = \int_{x\in\mathbb{C}^n} \varphi(x) \frac{e^{-\|x\|^2/2}}{(2\pi)^n} dx \quad ,$$

where, respectively, the natural volume forms on $\mathbb{P}^{n-1}$ and $\mathbb{C}^n$ are understood for each integral.

Now the integrand in Proposition 3 is not $\mathbb{C}^*$-invariant. This is why we will need the following formula:

Lemma 6 *Under the hypotheses of Lemma 5,*

$$\frac{1}{\mathrm{Vol}(\mathbb{P}^{n-1})} \int_{[x]\in\mathbb{P}^{n-1}} \varphi(x) d[x] = \frac{1}{2n} \int_{x\in\mathbb{C}^n} \|x\|^2 \varphi(x) \frac{e^{-\|x\|^2/2}}{(2\pi)^n} dx \quad .$$

where, respectively, the natural volume forms on $\mathbb{P}^{n-1}$ and $\mathbb{C}^n$ are understood for each integral.

Proof:

$$\int_{x\in\mathbb{C}^n} \|x\|^2 \varphi(x) \frac{e^{-\|x\|^2/2}}{(2\pi)^n} dx = \int_{\Theta\in S^{2n-1}} \int_{r=0}^{\infty} |r|^{2n+1} \varphi(\Theta) \frac{e^{-|r|^2/2}}{(2\pi)^n} dr d\Theta$$

$$= \int_{\Theta\in S^{2n-1}} \left(- \left[|r|^{2n} \frac{e^{-|r|^2/2}}{(2\pi)^n} \right]_0^{\infty} + 2n \int_{r=0}^{\infty} |r|^{2n-1} \frac{e^{-|r|^2/2}}{(2\pi)^n} dr \right) \varphi(\Theta) d\Theta$$

$$= 2n \int_{x\in\mathbb{C}^n} \varphi(x) \frac{e^{-\|x\|^2/2}}{(2\pi)^n} dx$$

■

We can now introduce the notation:

$$\mathrm{WEDGE}^A(f_{\mathrm{I\!I}}) \stackrel{\mathrm{def}}{=} \bigwedge_{i=1}^{n} \frac{1}{\|f_{\mathrm{I\!I}}^i\|^2} f_{\mathrm{I\!I}}^i \cdot (Dv_{A_i})_{(p,q)} dp \wedge \bar{f}_{\mathrm{I\!I}}^i \cdot (D\bar{v}_{A_i})_{(p,q)} dq \quad .$$

This function is invariant under the $(\mathbb{C}^*)^n$-action $\lambda \star f_{\mathrm{I\!I}} : f_{\mathrm{I\!I}} \mapsto (\lambda_1 f_{\mathrm{I\!I}}^1, \cdots, \lambda_n f_{\mathrm{I\!I}}^n)$.

We adopt the following conventions: $\mathcal{F}_{\mathrm{I\!I}} \subset \mathcal{F}$ is the space spanned by coordinates $f_{\mathrm{I\!I}}$ and $\mathbb{P}(\mathcal{F}_{\mathrm{I\!I}})$ is its quotient by $(\mathbb{C}^*)^n$.

We apply n times Lemma 6 and obtain:

Proposition 4 *Let* $\mathrm{VOL} \stackrel{\mathrm{def}}{=} \mathrm{Vol}(\mathbb{P}^{n-1})^n$. *Then,*

$$\nu^{\mathcal{A}}(U,\varepsilon) \le \frac{(2n)^n}{\mathrm{VOL}} \int_{(p,q)\in U\subset \mathcal{T}^n} \int_{\substack{f_{\mathrm{I\!I}}\in\mathbb{P}(\mathcal{F}_{\mathrm{I\!I}}) \\ d_{\mathbb{P}}(f_{\mathrm{I\!I}},\Sigma_{(p,q)})<\varepsilon}} \mathrm{WEDGE}^A(f_{\mathrm{I\!I}})\ d\mathbb{P}(\mathcal{F}_{\mathrm{I\!I}})\ dV_{\mathcal{T}^n}$$

with equality when $\epsilon > \sqrt{n}$. *In the linear case,*

$$\nu^{\mathrm{Lin}}(U,\varepsilon) = \frac{(2n)^n}{\mathrm{VOL}} \int_{(p,q)\in U\subset \mathcal{T}^n} \int_{\substack{g_{\mathrm{I\!I}}\in\mathbb{P}(\mathcal{F}_{\mathrm{I\!I}}^{\mathrm{Lin}}) \\ d_{\mathbb{P}}(g_{\mathrm{I\!I}},\Sigma_{(p,q)}^{\mathrm{Lin}})<\varepsilon}} \mathrm{WEDGE}^{\mathrm{Lin}}(g_{\mathrm{I\!I}})\ d(P\mathcal{F}_{\mathrm{I\!I}}^{\mathrm{Lin}}) dV_{\mathcal{T}^n} \ ■$$

Now we introduce the following change of coordinates. Let $L \in GL(n)$ be such that the minimum in Definition 4, Pg. 6 is attained:

$$\begin{aligned} \varphi : \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1} &\to \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1} \\ f_{\mathrm{I\!I}} &\mapsto g_{\mathrm{I\!I}} \stackrel{\mathrm{def}}{=} \varphi(f_{\mathrm{I\!I}}) \quad \text{, such} \\ &\qquad \text{that } g_{\mathrm{I\!I}}^i = f_{\mathrm{I\!I}}^i \cdot Dv_{A_i} L \quad . \end{aligned}$$

Without loss of generality, we scale L such that $\det L = 1$. The following property follows from the definition of WEDGE:

$$\mathrm{WEDGE}^{A}(f_{\rm I\!I}) = \mathrm{WEDGE}^{\mathrm{Lin}}(g_{\rm I\!I}) \prod_{i=1}^{n} \frac{\|g^i_{\rm I\!I}\|^2}{\|f^i_{\rm I\!I}\|^2} \tag{3.5.1}$$

Assume now that $d_{\mathbb{P}}(f_{\rm I\!I}, \Sigma_{(p,q)}) < \varepsilon$. Then there is $\delta f \in \mathcal{F}_{\rm I\!I}$, such that $f + \delta f \in \Sigma^{\mathrm{Lin}}_{(p,q)}$ and $\|\delta f\| \leq \varepsilon$ (assuming the scaling $\|f^i_{\rm I\!I}\| = 1$ for all i).

Setting $g_{\rm I\!I} = \varphi(f_{\rm I\!I})$ and $\delta g = \varphi(\delta g)$, we obtain that $g + \delta g \in \Sigma^{\mathrm{Lin}}_{(p,q)}$.

$$d_{\mathbb{P}}(g, \Sigma^{\mathrm{Lin}}_{(p,q)}) \leq \sqrt{\sum_{i=1}^{n} \frac{\|\delta g^i\|^2}{\|g^i_{\rm I\!I}\|^2}}$$

At each value of i,

$$\frac{\|\delta g^i\|}{\|g^i_{\rm I\!I}\|} \leq \frac{\|\delta f^i\|}{\|f^i_{\rm I\!I}\|} \kappa(D_{f^i_{\rm I\!I}} \varphi^i)$$

where κ denotes Wilkinson's condition number of the linear operator $D_{f^i_{\rm I\!I}} \varphi^i$. This is precisely $\kappa(Dv_{A_i} L)$. Thus,

$$d_{\mathbb{P}}(g, \Sigma^{\mathrm{Lin}}_{(p,q)}) \leq \varepsilon \max_i \kappa(Dv_{A_i} L) = \varepsilon \sqrt{\kappa_U}$$

Thus, an ε-neighborhood of $\Sigma^{A}_{(p,q)}$ is mapped into a $\sqrt{\kappa_U}\varepsilon$ neighborhood of $\Sigma^{\mathrm{Lin}}_{(p,q)}$.

We use this property and equation (3.5.1) to bound:

$$\nu^{\mathcal{A}}(U, \varepsilon) \leq \frac{(2n)^n}{\mathrm{VOL}} \int_{(p,q) \in U \subset \mathcal{T}^n} \int_{\substack{g_{\rm I\!I} \in \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1} \\ d_{\mathbb{P}}(g_{\rm I\!I}, \Sigma^{\mathrm{Lin}}_{(p,q)}) < \sqrt{\kappa_U}\varepsilon}} \mathrm{WEDGE}^{\mathrm{Lin}}(g_{\rm I\!I}) \cdot$$
$$\cdot \prod_{i=1}^{n} \frac{\|g^i_{\rm I\!I}\|^2}{\|f^i_{\rm I\!I}\|^2} |J_{g_{\rm I\!I}} \varphi^{-1}|^2 \, d(\mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}) \, dV_{\mathcal{T}^n} \tag{3.5.2}$$

where $J_{g_{\rm I\!I}} \varphi^{-1}$ is the Jacobian of φ^{-1} at $g_{\rm I\!I}$.

Remark 3 *Considering each Dv_{A_i} as a map from $\mathbb{C}^n$ into $\mathbb{C}^n$, the Jacobian is:*

$$J_{g_{\rm I\!I}} \varphi^{-1} = \prod_{i=1}^{n} \frac{\|\varphi^{-1}(g_{\rm I\!I})^i\|^n}{\|g^i_{\rm I\!I}\|^n} \left(\det Dv^H_{A_i} Dv_{A_i} \right)^{-1/2} .$$

We will not use this value in the sequel. $\diamond$

In order to simplify the expressions for the bound on $\nu^{\mathcal{A}}(U, \varepsilon)$, it is convenient to introduce the following notation:

$$
\begin{aligned}
dP &\stackrel{\text{def}}{=} \frac{(2n)^n}{\text{VOL}} \text{WEDGE}^{\text{Lin}}(g_{\mathbb{I}}) \frac{d(\mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1})}{n!\ (\omega_{\text{Lin}})^{\bigwedge n}} \\
H &\stackrel{\text{def}}{=} \prod_{i=1}^{n} \frac{\|g_{\mathbb{I}}^i\|^2}{\|f_{\mathbb{I}}^i\|^2} |J_g \varphi^{-1}|^2 \\
\chi_\delta &\stackrel{\text{def}}{=} \chi_{\{g \ | \ d_{\mathbb{P}}(g, \Sigma^{\text{Lin}}_{(p,q)}) < \delta\}}
\end{aligned}
$$

Now equation (3.5.2) becomes:

$$
\nu^{\mathcal{A}}(U, \varepsilon) \leq n! \int_{(p,q) \in U \subset \mathcal{T}^n} (\omega_{\text{Lin}})^{\bigwedge n} \int_{g_{\mathbb{I}} \in \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}} dP\ H(g_{\mathbb{I}})\ \chi_{\sqrt{\kappa_U} \varepsilon}(g_{\mathbb{I}}) \tag{3.5.3}
$$

Lemma 7 *Let (p, q) be fixed. Then $\mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}$ together with density function dP, is a probability space.*

Proof: The expected number of roots in U for a linear system is

$$
n! \int_{(p,q) \in U} \omega_{\text{Lin}}^{\bigwedge n} \int_{g_{\mathbb{I}} \in \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}} dP
$$

and is also equal to $n! \int_U \omega_{\text{Lin}}^{\bigwedge n}$. This holds for all U, hence the volume forms are the same and

$$
\int_{g_{\mathbb{I}} \in \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}} dP = 1. \ \blacksquare
$$

This allows us to interpret the inner integral of equation (3.5.3) as the expected value of a product. This is less than the product of the expected values, and:

$$
\begin{aligned}
\nu^{\mathcal{A}}(U, \varepsilon) \leq n! \int_{(p,q) \in U \subset \mathcal{T}^n} (\omega_{\text{Lin}})^{\bigwedge n} & \left(\int_{g_{\mathbb{I}} \in \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}} dP\ H(g_{\mathbb{I}}) \right) \cdot \\
& \cdot \left(\int_{g_{\mathbb{I}} \in \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}} dP\ \chi_{\sqrt{\kappa_U} \varepsilon}(g_{\mathbb{I}}) \right)
\end{aligned}
$$

Because generic (square) systems of linear equations have exactly one root, we can also consider U as a probability space, with probability measure $\frac{1}{\text{Vol}^{\text{Lin}}(U)} n! \omega_{\text{Lin}}^{\bigwedge n}$. Therefore, we can bound:

$$
\begin{aligned}
\nu^{\mathcal{A}}(U, \varepsilon) \leq \frac{1}{\text{Vol}^{\text{Lin}}(U)} & \left(\int_{(p,q) \in U} n! (\omega_{\text{Lin}})^{\bigwedge n} \int_{g_{\mathbb{I}} \in \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}} dP\ H(g_{\mathbb{I}}) \right) \cdot \\
& \cdot \left(\int_{(p,q) \in U} n! (\omega_{\text{Lin}})^{\bigwedge n} \int_{g_{\mathbb{I}} \in \mathbb{P}^{n-1} \times \cdots \times \mathbb{P}^{n-1}} dP\ \chi_{\sqrt{\kappa_U} \varepsilon}(g_{\mathbb{I}}) \right)
\end{aligned}
$$

The first parenthetical expression is $\mathrm{Vol}^A(U)$, the volume of U with respect to the toric volume form associated to $A = (A_1, \cdots, A_n)$. The second parenthetical expression is $\nu^{\mathrm{Lin}}(\sqrt{\kappa_U}\varepsilon, U)$. This concludes the proof of Theorem 5. ■

3.6 The Proof of Theorem 3

As in the complex case (Theorem 2), the expected number of roots can be computed by applying the coarea formula:

$$AVG = \int_{p\in U} \int_{f\in \mathcal{F}_p^{\mathbb{R}}} \prod_{i=1}^{n} \frac{e^{-\|f^i\|^2/2}}{\sqrt{2\pi}^{m_i}} \det(DG\ DG^H)^{-1/2} \quad .$$

Now there are three big differences. The set U is in $\mathbb{R}^n$ instead of $\mathcal{T}^n$, the space $\mathcal{F}_p^{\mathbb{R}}$ contains only real polynomials (and therefore has half the dimension), and we are integrating the square root of $1/\det(DG\ DG^H)$.

Since we do not know in general how to integrate such a square root, we bound the inner integral as follows. We consider the real Hilbert space of functions integrable in $\mathcal{F}_p^{\mathbb{R}}$ endowed with Gaussian probability measure. The inner product in this space is:

$$\langle \varphi, \psi \rangle \stackrel{\mathrm{def}}{=} \int_{\mathcal{F}_p^{\mathbb{R}}} \varphi(f)\psi(f) \prod_{i=1}^{n} \frac{e^{-\|f^i\|^2/2}}{\sqrt{2\pi}^{m_i-1}} dV \quad ,$$

where dV is Lebesgue volume. If $\mathbf{1}$ denotes the constant function equal to 1, we interpret

$$AVG = \int_{p\in U} (2\pi)^{-n/2} \left\langle \det(DG\ DG^H)^{-1/2}, \mathbf{1} \right\rangle \quad .$$

Hence Cauchy-Schwartz inequality implies:

$$AVG \le \int_{p\in U} (2\pi)^{-n/2} \| \det(DG\ DG^H)^{-1/2} \| \|\mathbf{1}\| \quad .$$

By construction, $\|\mathbf{1}\| = 1$, and we are left with:

$$AVG \le \int_{p\in U} (2\pi)^{-n/2} \sqrt{ \int_{\mathcal{F}_p^{\mathbb{R}}} \prod_{i=1}^{n} \frac{e^{-\|f^i\|^2/2}}{\sqrt{2\pi}^{m_i-1}} \det(DG\ DG^H)^{-1} } \quad .$$

As in the complex case, we add extra n variables:

$$AVG \leq (2\pi)^{-n/2} \int_{p \in U} \sqrt{\int_{\mathcal{F}^{\mathbb{R}}} \prod_{i=1}^{n} \frac{e^{-\|f^i\|^2/2}}{\sqrt{2\pi}^{m_i}} \det(DG\ DG^H)^{-1}} \quad ,$$

and we interpret $\det(DG\ DG^H)^{-1}$ in terms of a wedge. Since

$$\int_{x \in \mathbb{R}^m} |x_1|^2 \frac{e^{-\|x\|^2/2}}{\sqrt{2\pi}^m} = \int_{y \in \mathbb{R}} y^2 \frac{e^{-y^2/2}}{\sqrt{2\pi}} = \int_{y \in \mathbb{R}} \frac{e^{-y^2/2}}{\sqrt{2\pi}} = 1 \quad ,$$

we obtain:

$$AVG \leq (2\pi)^{-n/2} \int_{p \in U} \sqrt{n! d\mathcal{T}^n} = (2\pi)^{-n/2} \int_{p \in U} \sqrt{n! d\mathcal{T}^n} \quad .$$

Now we would like to use Cauchy-Schwartz again. This time, the inner product is defined as:

$$\langle \varphi, \psi \rangle \stackrel{\text{def}}{=} \int_{p \in U} \varphi(p)\psi(p) dV \quad .$$

Hence,

$$AVG \leq (2\pi)^{-n/2} \langle n! d\mathcal{T}^n, \mathbf{1} \rangle \leq (2\pi)^{-n/2} \|n! d\mathcal{T}^n\| \|\mathbf{1}\| \quad .$$

This time, $\|\mathbf{1}\|^2 = \lambda(U)$, so we bound:

$$\begin{aligned} AVG &\leq (2\pi)^{-n/2} \sqrt{\lambda(U)} \sqrt{\int_U n! d\mathcal{T}^n} \\ &\leq (4\pi^2)^{-n/2} \sqrt{\lambda(U)} \sqrt{\int_{(p,q) \in \mathcal{T}^n, p \in U} n! d\mathcal{T}^n} \blacksquare \end{aligned}$$

3.7 The Proof of Theorem 6

Let $\varepsilon > 0$. As in the mixed case, we define:

$$\begin{aligned} \nu_{\mathbb{R}}(U, \varepsilon) &\stackrel{\text{def}}{=} \text{Prob}_{f \in \mathcal{F}} \left[\boldsymbol{\mu}(f; U) > \varepsilon^{-1} \right] \\ &= \text{Prob}_{f \in \mathcal{F}} \left[\exists p \in U : ev(f; p) = 0 \text{ and } d_{\mathbb{P}}(f, \Sigma_p) < \varepsilon \right] \end{aligned}$$

where now $U \in \mathbb{R}^n$.

Let $V(\varepsilon) \stackrel{\text{def}}{=} \{(f, p) \in F_{\mathbb{R}} \times U : ev(f; p) = 0 \text{ and } d_{\mathbb{P}}(f, \Sigma_p) < \varepsilon\}$. We also define $\pi : V(\varepsilon) \to \mathbb{P}(\mathcal{F})$ to be the canonical projection mapping $F_{\mathbb{R}} \times U$ to $F_{\mathbb{R}}$ and set $\#_{V(\varepsilon)}(f) \stackrel{\text{def}}{=} \#\{p \in U : (f, p) \in V(\varepsilon)\}$. Then,

$$
\begin{aligned}
\nu_{\mathbb{R}}(U,\varepsilon) &= \int_{f\in\mathcal{F}^{\mathbb{R}}} \frac{e^{-\sum_i \|f^i\|^2/2}}{\sqrt{2\pi}^{\sum m_i}} \chi_{\pi(V(\varepsilon))}(f)\ d\mathcal{F}^{\mathbb{R}} \\
&\le \int_{f\in\mathcal{F}^{\mathbb{R}}} \frac{e^{-\sum_i \|f^i\|^2/2}}{\sqrt{2\pi}^{\sum m_i}} \#_{V(\varepsilon)} d\mathcal{F}^{\mathbb{R}} \\
&\le \int_{p\in U\subset\mathbb{R}^n} \int_{\substack{f\in\mathcal{F}_p^{\mathbb{R}} \\ d_{\mathbb{P}}(f,\Sigma_p)<\varepsilon}} \frac{e^{-\sum_i \|f^i\|^2/2}}{\sqrt{2\pi}^{\sum m_i}} \frac{1}{NJ(f;p)} d\mathcal{F}_p^{\mathbb{R}}\ dV_{\mathcal{T}^n}
\end{aligned}
$$

As before, we change coordinates in each fiber of $\mathcal{F}_A^{\mathbb{R}}$ by

$$
f = f_{\mathrm{I}} + f_{\mathrm{II}} + f_{\mathrm{III}}
$$

with f_{I}^i colinear to v_A^T, $(f_{\mathrm{II}}^i)^T$ in the range of Dv_A, and f_{III}^i orthogonal to f_{I}^i and f_{II}^i. This coordinate system is dependent on $p + q\sqrt{-1}$.

In the new coordinate system, formula 2.3.1 splits as follows:

$$
\begin{aligned}
&\det\left(DG_{(p)} DG_{(p)}^H\right)^{-1/2} dV_{\mathcal{T}^n} = \\
&\qquad = \left|\det \begin{bmatrix} (f_{\mathrm{II}}^1)_1 & \dots & (f_{\mathrm{II}}^1)_n \\ \vdots & & \vdots \\ (f_{\mathrm{II}}^n)_1 & \dots & (f_{\mathrm{II}}^n)_n \end{bmatrix}\right| \left|\det \begin{bmatrix} (Dv_A{}^{\mathrm{II}})_1^1 & \dots & (Dv_A{}^{\mathrm{II}})_n^1 \\ \vdots & & \vdots \\ (Dv_A{}^{\mathrm{II}})_1^n & \dots & (Dv_A{}^{\mathrm{II}})_n^n \end{bmatrix}\right| dV \\
&\qquad = \left|\det \begin{bmatrix} (f_{\mathrm{II}}^1)_1 & \dots & (f_{\mathrm{II}}^1)_n \\ \vdots & & \vdots \\ (f_{\mathrm{II}}^n)_1 & \dots & (f_{\mathrm{II}}^n)_n \end{bmatrix}\right| \sqrt{\det Dv_A^H Dv_A}
\end{aligned}
$$

The integral $E(U)$ of $\sqrt{\det Dv_A Dv_A^H}$ is the expected number of real roots on U, therefore

$$
\begin{aligned}
\nu_{\mathbb{R}}(U,\varepsilon) \le E(U) \int_{\substack{f_{\mathrm{II}}+f_{\mathrm{III}}\in\mathcal{F}_p^{\mathbb{R}} \\ d_{\mathbb{P}}(f_{\mathrm{II}}+f_{\mathrm{III}},\Sigma_p)<\varepsilon}} & \frac{e^{-\sum_i \|f_{\mathrm{II}}^i + f_{\mathrm{III}}^i\|^2/2}}{\sqrt{2\pi}^{\sum m_i}} \cdot \\
& \cdot \left|\det \begin{bmatrix} (f_{\mathrm{II}}^1)_1 & \dots & (f_{\mathrm{II}}^1)_n \\ \vdots & & \vdots \\ (f_{\mathrm{II}}^n)_1 & \dots & (f_{\mathrm{II}}^n)_n \end{bmatrix}\right| d\mathcal{F}_p^{\mathbb{R}} \quad .
\end{aligned}
$$

In the new system of coordinates, Σ_p is defined by the equation:

$$\det \begin{bmatrix} (f_{\mathrm{II}}^1)_1 & \dots & (f_{\mathrm{II}}^1)_n \\ \vdots & & \vdots \\ (f_{\mathrm{II}}^n)_1 & \dots & (f_{\mathrm{II}}^n)_n \end{bmatrix} = 0 \quad .$$

Since $\|f_{\mathrm{II}} + f_{\mathrm{III}}\| \geq \|f_{\mathrm{II}}\|$,

$$d_{\mathbb{P}}(f_{\mathrm{II}} + f_{\mathrm{III}}, \Sigma_p) < \varepsilon \Longrightarrow d_{\mathbb{P}}(f_{\mathrm{II}}, \Sigma_p) < \varepsilon \quad .$$

This implies:

$$\nu_{\mathbb{R}}(U, \varepsilon) \leq E(U) \int_{\substack{f_{\mathrm{II}} + f_{\mathrm{III}} \in F_p^{\mathbb{R}} \\ d_{\mathbb{P}}(f_{\mathrm{II}}, [\det = 0]) < \varepsilon}} \frac{e^{-\sum_i \|f_{\mathrm{II}}^i + f_{\mathrm{III}}^i\|^2/2}}{\sqrt{2\pi}^{\sum m_i}} \cdot \\ \cdot \left| \det \begin{bmatrix} (f_{\mathrm{II}}^1)_1 & \dots & (f_{\mathrm{II}}^1)_n \\ \vdots & & \vdots \\ (f_{\mathrm{II}}^n)_1 & \dots & (f_{\mathrm{II}}^n)_n \end{bmatrix} \right| \; d\mathcal{F}_p^{\mathbb{R}} \quad .$$

We can integrate the $(\sum m_i - n - 1)$ variables f_{III} to obtain:

$$\nu_{\mathbb{R}}(U, \varepsilon) = E(U) \int_{\substack{f_{\mathrm{II}} \in \mathbb{R}^{n^2} \\ d_{\mathbb{P}}(f_{\mathrm{II}}, [\det = 0]) < \varepsilon}} \frac{e^{-\sum_i \|f_{\mathrm{II}}^i\|^2/2}}{\sqrt{2\pi}^{n^2}} \left|\det f_{\mathrm{II}}\right|^2 \; d\mathbb{R}^{n^2} \quad .$$

This is $E(U)$ times the probability $\nu(n, \varepsilon)$ for the linear case. ■

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Appendix: The Coarea Formula

Here we give a short proof of the coarea formula, in a version suitable to the setting of this paper. This means we take all manifolds and functions smooth and avoid measure theory as much as possible.

Proposition 5

(1) Let X be a smooth Riemann manifold, of dimension m and volume form $|dX|$
(2) Let Y be a smooth Riemann manifold, of dimension n and volume form $|dY|$.
(3) Let U be an open set of X, and $F : U \rightarrow Y$ be a smooth map, such that DF_x is surjective for all x in U.
(4) Let $\varphi : X \rightarrow \mathbb{R}^+$ be a smooth function with compact support contained in U.

Then for almost all $z \in F(U)$, $V_z \stackrel{\text{def}}{=} F^{-1}(z)$ is a smooth Riemann manifold, and

$$\int_X \varphi(x) NJ(F;x)|dX| = \int_{z \in Y} \int_{x \in V_z} \varphi(x)|dV_z||dY|$$

where $|dV_z|$ is the volume element of V_z and $NJ(F,x) = \sqrt{\det DF_x^H DF_x}$ is the product of the singular values of DF_x.

By the implicit function theorem, whenever V_z is non-empty, it is a smooth $(m-n)$-dimensional Riemann submanifold of X. By the same reason, $V := \{(z,x) : x \in V_z\}$ is also a smooth manifold.

Let η be the following m-form restricted to V:

$$\eta = dY \wedge dV_z \quad .$$

This is **not** the volume form of V. The proof of Proposition 5 is divided into two steps:

Lemma 8
$$\int_V \varphi(x)|\eta| = \int_X \varphi(x)NJ(F;x)|dX|.$$

Lemma 9
$$\int_V \varphi(x)|\eta| = \int_{z\in Y}\int_{x\in V_z} \varphi(x)|dV_z||dY| \quad .$$

Proof of Lemma 8: We parametrize:

$$\begin{aligned} \psi : X &\to V \\ x &\mapsto (F(x), x) \end{aligned} \quad .$$

Then,
$$\int_V \varphi(x)|\eta| = \int_X (\varphi \circ \psi)(x)|\psi^*\eta| \quad .$$

We can choose an orthonormal basis $u_1, \cdots, u_m$ of T_xX such that $u_{n+1}, \cdots, u_m \in \ker DF_x$. Then,

$$D\psi(u_i) = \begin{cases} (DF_x u_i, u_i) & i = 1, \cdots, n \\ (0, u_i) & i = n+1, \cdots, m \end{cases} \quad .$$

Thus,

$$\begin{aligned} |\psi^*\eta(u_1, \cdots, u_m)| &= |\eta(D\psi u_1, \cdots, D\psi u_m)| \\ &= |dY(DF_x u_1, \cdots, DF_x u_n)|\ |dV_z(u_{n+1}, \cdots, u_m)| \\ &= |\det DF_x|_{\ker DF_x^\perp}| \\ &= NJ(F, x) \end{aligned}$$

and hence
$$\int_V \varphi(x)|\eta| = \int_X \varphi(x)NJ(F;x)|dX| \ \blacksquare$$

Proof of Lemma 9: We will prove this Lemma locally, and this implies the full Lemma through a standard argument (partitions of unity in a compact neighborhood of the support of φ).

Let x_0, z_0 be fixed. A small enough neighborhood of $(x_0, z_0) \subset V_{z_0}$ admits a fibration over V_{z_0} by planes orthogonal to $\ker DF_{x_0}$.

We parametrize:

$$\begin{aligned} \theta : Y \times V_{z_0} &\to V \\ (z, x) \quad &\mapsto (z, \rho(x, z)) \end{aligned} ,$$

where $\rho(x, z)$ is the solution of $F(\rho) = z$ in the fiber passing through (z_0, x). Remark that $\theta^* dY = dY$, and $\theta^* dV_z = \rho^* DV_z$. Therefore,

$$\theta^*(dY \wedge dV_z) = dY \wedge (\rho^* dV_z) \quad .$$

Also, if one fixes z, then ρ is a parametrization $V_{z_0} \to V_z$. We have:

$$\begin{aligned} \int_V \varphi(x)|\eta| &= \int_{Y \times V_{z_0}} \varphi(\rho(x,z))|\theta^*\eta| \\ &= \int_{z \in Y} \left(\int_{x \in V_{z_0}} \varphi(\rho(x,z)|\rho^* dV_z| \right) |dY| \\ &= \int_{z \in Y} \left(\int_{x \in V_z} \varphi(x)|dV_z| \right) |dY| \blacksquare \end{aligned}$$

The proposition below is essentially Theorem 3, Pg. 240 of (BCSS98). However, we do not require our manifolds to be compact. We assume all maps and manifolds are smooth, so that we can apply proposition 5.

Proposition 6

(1) Let X be a smooth m-dimensional manifold with volume element $|dX|$.
(2) Let Y be a smooth n-dimensional manifold with volume element $|dY|$.
(3) Let V be a smooth m-dimensional submanifold of $X \times Y$, and let $\pi_1 : V \to X$ and $\pi_2 : V \to Y$ be the canonical projections from $X \times Y$ to its factors.
(4) Let Σ' be the set of critical points of π_1, we assume that Σ' has measure zero and that Σ' is a manifold.
(5) We assume that π_2 is regular (all points in $\pi_2(V)$ are regular values).
(6) For any open set $U \subset V$, for any $x \in X$, we write: $\#_U(x) \stackrel{\text{def}}{=} \#\{\pi_1^{-1}(x) \cap U\}$. We assume that $\int_{x \in X} \#_V(x)|dX|$ is finite.

Then, for any open set $U \subset V$,

$$\int_{x \in \pi_1(U)} \#_U(x)|dX| = \int_{z \in Y} \int_{\substack{x \in V_z \\ (x,z) \in U}} \frac{1}{\sqrt{\det DG_x DG_x^H}} |dV_z||dY|$$

where G is the implicit function for $(\hat{x}, G(\hat{x})) \in V$ in a neighborhood of $(x, z) \in V \setminus \Sigma'$. ■

Proof: Every $(x, z) \in U \setminus \Sigma'$ admits an open neighborhood such that π_1 restricted to that neighborhood is a diffeomorphism. This defines an open covering of $U \setminus \Sigma'$. Since $U \setminus \Sigma'$ is locally compact, we can take a countable sub-covering and define a partition of unity $(\varphi_\lambda)_{\lambda \in \Lambda}$ subordinated to that sub-covering.

Also, if we fix a value of z, then $(\varphi_\lambda)_{\lambda \in \Lambda}$ becomes a partition of unity for $\pi_1(\pi_1^{-1}(V_z) \cap U)$. Therefore,

$$\begin{aligned}
\int_{x \in \pi_1(U)} \#_U(x) |dX| &= \sum_{\lambda \in \Lambda} \int_{x, z \in \mathrm{Supp}\varphi_\lambda} \varphi_\lambda(x, z) |dX| \\
&= \sum_{\lambda \in \Lambda} \int_{z \in Y} \int_{x, z \in \mathrm{Supp}\varphi_\lambda} \frac{\varphi_\lambda(x, z)}{NJ(G, x)} |dX| \\
&= \int_{z \in Y} \sum_{\lambda \in \Lambda} \int_{x, z \in \mathrm{Supp}\varphi_\lambda} \frac{\varphi_\lambda(x, z)}{NJ(G, x)} |dX| \\
&= \int_{z \in Y} \int_{x \in V_z} \frac{1}{NJ(G, x)} |dX|
\end{aligned}$$

where the second equality uses Proposition 5 with $\varphi = \varphi_\lambda / NJ$. Since $NJ = \sqrt{\det DG_x DG_x^H}$, we are done. ■

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