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# Spectral monotonicity under Gaussian convolution ${ }^{(*)}$ 

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#### Abstract

We show that the Poincaré constant of a log-concave measure in Euclidean space is monotone increasing along the heat flow. In fact, the entire spectrum of the associated Laplace operator is monotone decreasing. Two proofs of these results are given. The first proof analyzes a curvature term of a certain time-dependent diffusion, and the second proof constructs a contracting transport map following the approach of Kim and Milman.

Résumé. - Nous montrons que la constante de Poincaré d'une mesure log-concave dans l'espace euclidien est croissante le long du flot de la chaleur. En fait, le spectre entier de l'opérateur de Laplace associé est décroissant. Deux preuves de ces résultats sont données. La première preuve analyse un terme de courbure d'une certaine diffusion dépendant du temps, et la seconde preuve construit une application de transport contractante en suivant l'approche de Kim et Milman.


## 1. Introduction

The Poincaré constant $C_{P}(\mu)$ of a Borel probability measure $\mu$ on $\mathbb{R}^{n}$ is the smallest constant $C \geqslant 0$ such that for any locally-Lipschitz function $f \in L^{2}(\mu)$,

$$
\operatorname{Var}_{\mu}(f) \leqslant C \cdot \int_{\mathbb{R}^{n}}|\nabla f|^{2} \mathrm{~d} \mu
$$

where $\operatorname{Var}_{\mu}(f)=\int_{\mathbb{R}^{n}} f^{2} \mathrm{~d} \mu-\left(\int_{\mathbb{R}^{n}} f \mathrm{~d} \mu\right)^{2}$ and $|\cdot|$ is the Euclidean norm. The Poincaré constant governs the rate of convergence to equilibrium of the Langevin dynamics in velocity space [33].

[^0]Suppose that $\mu$ admits a smooth, positive density $\rho$ in $\mathbb{R}^{n}$. The Laplace operator associated with $\mu$, defined a priori on smooth, compactly supported functions $u: \mathbb{R}^{n} \rightarrow \mathbb{R}$, is given by

$$
\begin{equation*}
L u=L_{\mu} u=\Delta u+\nabla(\log \rho) \cdot \nabla u \tag{1.1}
\end{equation*}
$$

It satisfies

$$
\int_{\mathbb{R}^{n}}(L u) v \mathrm{~d} \mu=-\int_{\mathbb{R}^{n}}\langle\nabla u, \nabla v\rangle \mathrm{d} \mu
$$

for any two smooth functions $u, v: \mathbb{R}^{n} \rightarrow \mathbb{R}$, one of which is compactly supported. The operator $L_{\mu}$ is essentially self-adjoint in $L^{2}(\mu)$, negative semi-definite, with a simple eigenvalue at 0 corresponding to the constant eigenfunction (see [2, Corollary 3.2.2]). The Poincaré constant is given by

$$
C_{P}(\mu)=1 / \lambda_{1}^{(\mu)}
$$

where $\lambda_{1}^{(\mu)}$ is the spectral gap of $L$, the infimum over all positive $\lambda>0$ that belong to the spectrum of $-L$. Under mild regularity assumptions the spectrum of $L$ is discrete (e.g., when $\rho$ is $C^{2}$ and $\Delta(\sqrt{\rho}) / \sqrt{\rho}$ tends to infinity at infinity [2, Corollary 4.10.9], or when $\rho$ is $\log$-concave and $|\log \rho(x)| /|x|$ tends to infinity at infinity, as shown in Appendix A below). In this case we write

$$
0=\lambda_{0}^{(\mu)}<\lambda_{1}^{(\mu)} \leqslant \lambda_{2}^{(\mu)} \leqslant \lambda_{3}^{(\mu)} \leqslant \ldots
$$

for the eigenvalues of $-L$, repeated according to their multiplicity.
A non-negative function $\rho$ on $\mathbb{R}^{n}$ is log-concave if $K=\left\{x \in \mathbb{R}^{n}\right.$; $\rho(x)>0\}$ is convex, and $\log \rho$ is concave in $K$. An absolutely continuous probability measure on $\mathbb{R}^{n}$ is called log-concave if it has a log-concave density. An arbitrary probability measure on $\mathbb{R}^{n}$ is called log-concave if it is the pushforward of some absolutely continuous log-concave probability measure on $\mathbb{R}^{k}$ under an injective affine map. An example of a log-concave probability measure is $\gamma_{s}$, the Gaussian probability measure on $\mathbb{R}^{n}$ of mean zero and covariance $s$. Id. In a minor abuse of notation, we use $\gamma_{s}$ to denote also its density function $\gamma_{s}(x)=(2 \pi s)^{-n / 2} \exp \left(-|x|^{2} /(2 s)\right)$. Another example of a log-concave probability measure is the uniform probability measure on any convex body in $\mathbb{R}^{n}$. The convolution of two log-concave probability measures is again log-concave, as follows from the Prékopa-Leindler inequality [8, Theorem 1.2.3] or from the earlier work by Davidovič, Korenbljum and Hacet [17].

The Poincaré constant is a particularly useful invariant in the class of logconcave probability measures. For example, when $\mu$ is absolutely-continuous and log-concave, its Poincaré constant is determined, up to a multiplicative
universal constant, by the isoperimetric constant

$$
h(\mu)=\inf _{A \subseteq \mathbb{R}^{n}} \frac{\int_{\partial A} \rho}{\min \{\mu(A), 1-\mu(A)\}}
$$

where the infimum runs over all open sets $A \subseteq \mathbb{R}^{n}$ with smooth boundary. Indeed, the Cheeger [13] and Buser-Ledoux [9, 28] inequalities state that for any absolutely-continuous, log-concave probability measure $\mu$ on $\mathbb{R}^{n}$,

$$
\frac{1}{9} \leqslant C_{P}(\mu) \cdot h^{2}(\mu) \leqslant 4
$$

A well-known conjecture by Kannan, Lovász and Simonovits (KLS) states that the Poincaré constant of a log-concave probability measure is equivalent, up to a multiplicative universal constant, to the operator norm of the covariance matrix of $\mu$. See the recent papers by Chen [14] and by Klartag and Lehec [25] for more background and for the best known results towards this conjecture.

Abbreviate $\gamma=\gamma_{1}$, the standard Gaussian measure in $\mathbb{R}^{n}$, whose Poincaré constant is $C_{P}(\gamma)=1$ (e.g., [2, Proposition 4.1.1]). It was proven by Cattiaux and Guillin [12, Theorem 9.4.3] that when $\mu$ is a log-concave probability measure,

$$
\begin{equation*}
C_{P}(\mu) \leqslant C_{P}(\mu * \gamma)+1 \tag{1.2}
\end{equation*}
$$

where $\mu * \gamma$ is the convolution of $\mu$ and $\gamma$. The reverse inequality $C_{P}(\mu) \geqslant$ $C_{P}(\mu * \gamma)-1$ is much easier to obtain and does not require log-concavity (see, e.g., [3, Proposition 1] or [15]). Our main result in this paper is an improvement upon (1.2):

THEOREM 1.1. - Let $\mu$ be a log-concave probability measure on $\mathbb{R}^{n}$. Then,

$$
\begin{equation*}
C_{p}(\mu) \leqslant C_{P}(\mu * \gamma) \tag{1.3}
\end{equation*}
$$

Moreover, assuming that $\mu$ admits a density that is smooth and positive in $\mathbb{R}^{n}$ and that $L_{\mu}$ has a discrete spectrum tending to infinity, we have

$$
\lambda_{k}^{(\mu * \gamma)} \leqslant \lambda_{k}^{(\mu)} \quad(k=1,2, \ldots)
$$

We remark that the reverse inequality $1 / \lambda_{k}^{(\mu * \gamma)} \leqslant 1 / \lambda_{k}^{(\mu)}+1$ is known to hold without any log-concavity assumptions, see e.g. [15]. Two proofs of Theorem 1.1 are presented here. One of these proofs utilizes a method from Kim and Milman [23] to construct a contraction transporting $\mu * \gamma$ to $\mu$. Recall that a map $T: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is a contraction if $|T(x)-T(y)| \leqslant|x-y|$ for all $x, y \in \mathbb{R}^{n}$.

THEOREM 1.2. - Let $\mu$ be a log-concave probability measure on $\mathbb{R}^{n}$. Then there exists a contraction $T: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ that pushes forward $\mu * \gamma$ to $\mu$.

This result is reminiscent of Caffarelli's theorem [10], which states that there is a contraction pushing forward $\gamma$ to $\mu$ in the case where the density of $\mu$ with respect to the measure $\gamma$ is log-concave. As is well-known, Theorem 1.2 implies that the Poincaré constant of $\mu$ is not larger than that of $\mu * \gamma$. Moreover, as explained e.g. in Ledoux [27, Proposition 1.2], it follows from Theorem 1.2 that when $\mu$ is an absolutely-continuous, log-concave probability measure on $\mathbb{R}^{n}$,

$$
\begin{equation*}
h(\mu) \geqslant h(\mu * \gamma) \tag{1.4}
\end{equation*}
$$

There is also a corresponding inequality between the log-Sobolev constants of $\mu$ and $\mu * \gamma$, or any other quantity involving a Rayleigh-type quotient, see Caffarelli [10, Corollary 8]. We explain the proof of Theorem 1.2 and its implications in Section 3.

We continue with a discussion of an additional proof of Theorem 1.1, which was chronologically the first proof that we found. For $s>0$ denote

$$
\mu_{s}=\mu * \gamma_{s}
$$

the evolution of the measure $\mu$ under the heat flow. The log-concavity of $\mu$ implies that $\mu_{s}$ is log-concave as well. We will show that $C_{P}\left(\mu_{s}\right)$ is nondecreasing in $s$. For a function $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$ we consider its evolution under the heat semigroup

$$
P_{s}(f)=f * \gamma_{s} \quad(s>0)
$$

whenever the integrals defining the convolution converge absolutely. Setting $P_{0}=$ Id we obtain the heat semigroup $\left(P_{s}\right)_{s \geqslant 0}$, which satisfies the Euclidean heat equation

$$
\begin{equation*}
\frac{\partial}{\partial s} P_{s} f=\frac{\Delta P_{s} f}{2} \quad(s>0) \tag{1.5}
\end{equation*}
$$

The operator $P_{s}: L^{2}\left(\mu_{s}\right) \rightarrow L^{2}(\mu)$ is a contraction operator with $P_{s}(1)=1$ since

$$
\left\|P_{s}(f)\right\|_{L^{2}(\mu)}^{2}=\int_{\mathbb{R}^{n}} P_{s}(f)^{2} \rho \leqslant \int_{\mathbb{R}^{n}} P_{s}\left(f^{2}\right) \rho=\int_{\mathbb{R}^{n}} f^{2} P_{s}(\rho)=\|f\|_{L^{2}\left(\mu_{s}\right)}^{2}
$$

Write $\rho_{s}=\rho * \gamma_{s}$ for the density of the probability measure $\mu_{s}$, which is a smooth positive function in $\mathbb{R}^{n}$. The adjoint operator $Q_{s}=P_{s}^{*}: L^{2}(\mu) \rightarrow$ $L^{2}\left(\mu_{s}\right)$ is defined by $Q_{0} \varphi=\varphi$ and

$$
\begin{equation*}
Q_{s} \varphi=\frac{P_{s}(\varphi \rho)}{\rho_{s}} \quad(s>0) \tag{1.6}
\end{equation*}
$$

and again it is a contraction operator with $Q_{s}(1)=1$. It follows from (1.5) and (1.6) that the evolution equation for $Q_{s}$ is the parabolic equation

$$
\begin{equation*}
\frac{\partial}{\partial s} Q_{s} \varphi=\frac{\Delta Q_{s} \varphi}{2}+\nabla \log \rho_{s} \cdot \nabla Q_{s} \varphi \tag{1.7}
\end{equation*}
$$

We are thus led to define the "box operator"

$$
\begin{equation*}
\square_{s} u=\frac{\Delta u}{2}+\nabla\left(\log \rho_{s}\right) \cdot \nabla u \tag{1.8}
\end{equation*}
$$

This operator resembles the Laplace operator $L_{s}:=L_{\mu_{s}}$. Indeed, we have

$$
\begin{equation*}
L_{s}=\square_{s}+\frac{\Delta}{2} . \tag{1.9}
\end{equation*}
$$

The $\square_{s}$ operator obeys a Bochner-type formula, which is unsurprising as $\square_{s}$ equals half of the Laplace operator associated with the log-concave probability measure whose density is proportional to $\rho_{s}^{2}$. Indeed, we compute that for smooth $u, v: \mathbb{R}^{n} \rightarrow \mathbb{R}$,

$$
\begin{equation*}
\Gamma_{2}(u, v):=\square_{s}\langle\nabla u, \nabla v\rangle-\left\langle\nabla \square_{s} u, \nabla v\right\rangle-\left\langle\nabla u, \nabla \square_{s} v\right\rangle \tag{1.10}
\end{equation*}
$$

satisfies

$$
\begin{equation*}
\Gamma_{2}(u, u)=\left\|\nabla^{2} u\right\|_{H S}^{2}-2 \nabla^{2}\left(\log \rho_{s}\right) \nabla u \cdot \nabla u \tag{1.11}
\end{equation*}
$$

where $\left\|\nabla^{2} u\right\|_{H S}$ is the Hilbert-Schmidt norm of the Hessian matrix $\nabla^{2} u$. The expression in (1.11) is similar to the Bochner-type formula of the operator $L_{s}$, the main difference being the factor 2 in front of the second summand in (1.11), which is the "curvature term." Moreover, setting $\Gamma_{0}(u, v)=u v$ and $\Gamma_{1}(u, v)=\nabla u \cdot \nabla v$, we have

$$
\frac{\mathrm{d}}{\mathrm{~d} s} \int_{\mathbb{R}^{n}} \Gamma_{i}\left(Q_{s} \varphi, Q_{s} \varphi\right) \mathrm{d} \mu_{s}=-\int_{\mathbb{R}^{n}} \Gamma_{i+1}\left(Q_{s} \varphi, Q_{s} \varphi\right) \mathrm{d} \mu_{s} \quad(i=0,1)
$$

under some regularity assumptions to be explained below. It follows that the Rayleigh quotient

$$
\frac{\int_{\mathbb{R}^{n}}\left|\nabla Q_{s} \varphi\right|^{2} \mathrm{~d} \mu_{s}}{\int_{\mathbb{R}^{n}}\left(Q_{s} \varphi\right)^{2} \mathrm{~d} \mu_{s}}
$$

is non-increasing in $s \in(0, \infty)$. This fact, formulated as Theorem 2.4 below, implies Theorem 1.1. More details, explanations and rigourous proofs are provided in Section 2.

In Section 4 we discuss conceptual aspects of the evolution $\left(Q_{s} \varphi\right)_{s \geqslant 0}$, and explain how it is equivalent to Eldan's stochastic localization [19, 30] and Föllmer's drift [32]. We also provide a Bayesian interpretation of this evolution, and explore various connections between these points of view.

## 2. A dynamic variant of $\Gamma$-calculus

In this section we prove Theorem 1.1. Consider the linear differential operator $\square_{s}$ defined by formula (1.8) above. Similarly to the formalism from [2],
for smooth functions $u, v: \mathbb{R}^{n} \rightarrow \mathbb{R}$ we define $\Gamma_{0}(u, v)=u v$, and for $i \geqslant 0$ and $s>0$,

$$
\begin{equation*}
\Gamma_{i+1}(u, v)=\square_{s} \Gamma_{i}(u, v)-\Gamma_{i}\left(u, \square_{s} v\right)-\Gamma_{i}\left(\square_{s} u, v\right)-\frac{\mathrm{d} \Gamma_{i}}{\mathrm{~d} s}(u, v) \tag{2.1}
\end{equation*}
$$

Thus $\Gamma_{1}(u, v)=\nabla u \cdot \nabla v$ and $\Gamma_{2}(u, v)$ coincides with definition (1.10) above. The rationale for definition (2.1) is that, under regularity assumptions stated below,

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{~d} s} \int_{\mathbb{R}^{n}} \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \mathrm{d} \mu_{s}=-\int_{\mathbb{R}^{n}} \Gamma_{i+1}\left(\varphi_{s}, \varphi_{s}\right) \mathrm{d} \mu_{s} \tag{2.2}
\end{equation*}
$$

where $\varphi_{s}=Q_{s} \varphi$. If we were allowed to ignore all regularity issues, (2.2) could be proven as follows: differentiating under the integral sign and applying (1.5) and (1.7),

$$
\begin{align*}
& \frac{\mathrm{d}}{\mathrm{~d} s} \int_{\mathbb{R}^{n}} \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \rho_{s} \\
& =2 \int_{\mathbb{R}^{n}} \Gamma_{i}\left(\square_{s} \varphi_{s}, \varphi_{s}\right) \rho_{s}+\int_{\mathbb{R}^{n}} \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \frac{\Delta \rho_{s}}{2}+\int_{\mathbb{R}^{n}} \frac{\mathrm{~d} \Gamma_{i}}{\mathrm{~d} s}\left(\varphi_{s}, \varphi_{s}\right) \rho_{s} \\
& =\int_{\mathbb{R}^{n}}\left[2 \Gamma_{i}\left(\square_{s} \varphi_{s}, \varphi_{s}\right)+\frac{\Delta \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right)}{2}\right] \mathrm{d} \mu_{s}+\int_{\mathbb{R}^{n}} \frac{\mathrm{~d} \Gamma_{i}}{\mathrm{~d} s}\left(\varphi_{s}, \varphi_{s}\right) \mathrm{d} \mu_{s} . \tag{2.3}
\end{align*}
$$

Next we use (1.9) and the fact that $\int_{\mathbb{R}^{n}}\left(L_{s} u\right) \mathrm{d} \mu_{s}=0$ under regularity assumptions (e.g., when $u$ is smooth and compactly supported). This yields

$$
\begin{align*}
& \frac{\mathrm{d}}{\mathrm{~d} s} \\
& \quad \int_{\mathbb{R}^{n}} \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \rho_{s} \\
& \quad=\int_{\mathbb{R}^{n}}\left[2 \Gamma_{i}\left(\square_{s} \varphi_{s}, \varphi_{s}\right)-\square_{s} \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right)\right] \mathrm{d} \mu_{s}+\int_{\mathbb{R}^{n}} \frac{\mathrm{~d} \Gamma_{i}}{\mathrm{~d} s}\left(\varphi_{s}, \varphi_{s}\right) \mathrm{d} \mu_{s}  \tag{2.4}\\
& \quad=-\int_{\mathbb{R}^{n}} \Gamma_{i+1}\left(\varphi_{s}, \varphi_{s}\right) \mathrm{d} \mu_{s}
\end{align*}
$$

This would be a rigorous proof for (2.2) had we worked in the context of a compact Riemannian manifold (which also has a heat kernel $P_{s}: L^{2}(\mu) \rightarrow$ $L^{2}\left(\mu_{s}\right)$ and a corresponding adjoint $\left.Q_{s}=P_{s}^{*}\right)$. However, in this paper we are interested in the non-compact situation of $\mathbb{R}^{n}$, since we rely on the fact that the heat flow preserves curvature conditions such as log-concavity, which is currently known to hold only for a Euclidean space [26]. Nevertheless, the operators $\square_{s}$ and $\left(Q_{s}\right)_{s \geqslant 0}$ seem rather natural also in the Riemannian setting. We remark that our formalism is rather similar to the inhomogeneous Bakry-Émery method recently introduced by Bodineau and Bauerschmidt [5] in a different context.

Our first task in this section is to rigorously justify (2.2) for a fairly large class of functions $\varphi$. To do this, we shall express $Q_{s}$ explicitly as an integral operator.

Recall that we work with an absolutely-continuous, log-concave probability measure $\mu$ on $\mathbb{R}^{n}$ having density $\rho$. As before, for $s>0$ we write $\mu_{s}=\mu * \gamma_{s}$ and $\rho_{s}=P_{s} \rho=\rho * \gamma_{s}$, while the operator $Q_{s}$ is defined via formula (1.6). For $s>0$ and $y \in \mathbb{R}^{n}$ we define the probability density

$$
\begin{equation*}
p_{s, y}(x)=e^{\frac{\langle y, x\rangle}{s}-\frac{|x|^{2}}{2 s}} \frac{\rho(x)}{Z_{s, y}} \tag{2.5}
\end{equation*}
$$

where

$$
\begin{equation*}
Z_{s, y}=\int_{\mathbb{R}^{n}} e^{\frac{\langle y, x\rangle}{s}-\frac{|x|^{2}}{2 s}} \rho(x) \mathrm{d} x=\frac{\rho_{s}(y)}{\gamma_{s}(y)} \tag{2.6}
\end{equation*}
$$

is a normalizing constant (the "partition function"). In the next lemma we express the value of $Q_{s} \varphi$ at the point $y$ as the average of $\varphi$ with respect to the density $p_{s, y}$.

Lemma 2.1. - Let $y \in \mathbb{R}^{n}, s>0$ and suppose that $\varphi \in L^{1}(\mu)$, or more generally, that $\varphi: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is such that $\varphi(x) e^{-t|x|^{2}} \in L^{1}(\mu)$ for some $t \in$ $(0, s)$. Then

$$
\begin{equation*}
Q_{s} \varphi(y)=\int_{\mathbb{R}^{n}} \varphi \cdot p_{s, y} \tag{2.7}
\end{equation*}
$$

Furthermore, $Q_{s} \varphi(y)$ is a smooth function of $y \in \mathbb{R}^{n}$ and $s>0$ which satisfies

$$
\begin{align*}
\nabla Q_{s} \varphi & =\frac{Q_{s}(x \varphi)-Q_{s}(\varphi) Q_{s}(x)}{s}  \tag{2.8}\\
\partial_{s} Q_{s} \varphi(y) & =-\frac{Q_{s}\left(f_{y} \varphi\right)-Q_{s}(\varphi) Q_{s}\left(f_{y}\right)}{s^{2}}=\square_{s} Q_{s} \varphi(y) \tag{2.9}
\end{align*}
$$

where $f_{y}(x)=\langle x, y\rangle-|x|^{2} / 2$.
Proof. - According to (2.5) and (2.6),

$$
\begin{align*}
& \rho_{s}(y) \cdot \int_{\mathbb{R}^{n}} \varphi \cdot p_{s, y} \\
& \quad= \int_{\mathbb{R}^{n}} \varphi(x) \cdot e^{\langle y, x\rangle / s-|x|^{2} /(2 s)} \rho(x) \mathrm{d} x \cdot \gamma_{s}(y) \\
& \quad=\int_{\mathbb{R}^{n}} \varphi(x) \rho(x) \cdot \gamma_{s}(y-x) \mathrm{d} x=\left[(\varphi \rho) * \gamma_{s}\right](y)=P_{s}(\varphi \rho) \tag{2.10}
\end{align*}
$$

Now (2.7) follows from (1.6) and (2.10). The smoothness of $Q_{s} \varphi$ and equations (2.8) and (2.9) follow by differentiating (2.6) and (2.7) under the integral sign. This is legitimate, since any partial derivative in the $(s, y)$-variables of the function $p_{s, y}(x) \varphi(x)$ is seen to be bounded by an integrable function, and the bound is locally uniform in $s$ and $y$.

By a multi-index $(k, \alpha)$ we mean a non-negative integer $k$ and a vector $\alpha=\left(\alpha_{1}, \ldots, \alpha_{n}\right)$ of nonnegative integers. For a multi-index $(k, \alpha)$ and for a smooth function $f(s, y)$ we abbreviate

$$
\partial_{s}^{k} \partial_{y}^{\alpha} f=\left(\frac{\partial}{\partial s}\right)^{k}\left(\frac{\partial}{\partial y_{1}}\right)^{\alpha_{1}} \cdots\left(\frac{\partial}{\partial y_{n}}\right)^{\alpha_{n}} f(s, y) \quad\left(s>0, y \in \mathbb{R}^{n}\right)
$$

We denote $|\alpha|=\sum_{i}\left|\alpha_{i}\right|$. We say that a measurable function $\varphi: \mathbb{R}^{n} \rightarrow \mathbb{R}$ has subexponential decay relative to $\rho$ if there exists $a, C>0$ such that $|\varphi(x)| \leqslant \frac{C}{\sqrt{\rho(x)}} e^{-a|x|}$ for all $x \in \mathbb{R}^{n}$ for which $\rho(x)>0$. Since $\rho$ is an integrable, log-concave function, there exist $\alpha, \beta>0$ such that

$$
\begin{equation*}
\rho(x) \leqslant \alpha e^{-\beta|x|} \quad \text { for all } x \in \mathbb{R}^{n} . \tag{2.11}
\end{equation*}
$$

(see e.g., [8, Lemma 2.2.1]). In particular, a function $\varphi$ that grows at most polynomially at infinity, has subexponential decay relative to $\rho$.

Lemma 2.2. - Fix $s>0$ and let $(k, \alpha)$ be a multi-index. Then,
(i) The function $\partial_{s}^{k} \partial_{y}^{\alpha} \log \rho_{s}(y)$ grows at most polynomially at infinity in $y \in \mathbb{R}^{n}$.
(ii) Let $\varphi$ have subexponential decay relative to $\rho$. Then the function $\partial_{s}^{k} \partial_{y}^{\alpha} Q_{s} \varphi(y)$ has subexponential decay relative to $\rho_{s}$.

Moreover, if $s$ varies in an interval $\left[s_{0}, s_{1}\right]$ with $s_{0}>0$, then the implied constants in these two assertions may be chosen not to depend on $s$.

Proof. - Using the heat equation, we can replace time derivatives of $\rho_{s}$ by space derivatives, so in (i) we only need consider space derivatives of $\log \rho_{s}$. By differentiating (2.6) with respect to $y$ we see that

$$
\nabla_{y} \log \rho_{s}(y)=-\frac{y}{s}+\nabla_{y} \log Z_{s, y}=\frac{Q_{s}(x)(y)-y}{s}
$$

Repeated differentiations show that conclusion (i) would follow once we prove the following claim: for any $d>0$, the function $Q_{s}\left(|x|^{d}\right)(y)$ grows at most polynomially at infinity as a function of $y \in \mathbb{R}^{n}$, with the implied constants not depending on $s \in\left[s_{0}, s_{1}\right]$.

Let us prove this claim. It follows from (2.11) that $C_{d}:=\int_{\mathbb{R}^{n}}(1+$ $|x|)^{d} \rho(x) \mathrm{d} x<\infty$. Therefore, for any $y \in \mathbb{R}^{n}$ and $s>0$,

$$
\frac{\int_{\mathbb{R}^{n}}|x-y|^{d} e^{-|x-y|^{2} /(2 s)} \rho(x) \mathrm{d} x}{\int_{\mathbb{R}^{n}} e^{-|x-y|^{2} /(2 s)} \rho(x) \mathrm{d} x} \leqslant \frac{\int_{\mathbb{R}^{n}}|x-y|^{d} \rho(x) \mathrm{d} x}{\int_{\mathbb{R}^{n}} \rho(x) \mathrm{d} x} \leqslant C_{d}(1+|y|)^{d}
$$

where the first inequality follows from the fact that $|x-y|^{d}$ is increasing in $|x-y|$ while $e^{-|x-y|^{2} /(2 s)}$ is decreasing in $|x-y|$, and the second inequality
uses $|x-y| \leqslant(1+|x|)(1+|y|)$. Consequently, as $|x|^{d} \leqslant 2^{d}\left(|y|^{d}+|x-y|^{d}\right)$,

$$
\begin{aligned}
Q_{s}\left(|x|^{d}\right)(y)=\int_{\mathbb{R}^{n}}|x|^{d} p_{s, y}(x) \mathrm{d} x=\frac{\int_{\mathbb{R}^{n}}|x|^{d} e^{-|x-y|^{2} /(2 s)} \rho(x) \mathrm{d} x}{\int_{\mathbb{R}^{n}} e^{-|x-y|^{2} /(2 s)} \rho(x) \mathrm{d} x} \\
\leqslant \widetilde{C}_{d}(1+|y|)^{d}
\end{aligned}
$$

for some coefficient $\widetilde{C}_{d}$ depending only on $\rho$ and on $d$. This shows that $Q_{s}\left(|x|^{d}\right)(y)$ grows at most polynomially, from which (i) follows.

We move on to the proof of (ii). Let $a, C>0$ be such that $\sqrt{\rho(x)} \cdot|\varphi(x)| \leqslant$ $C e^{-a|x|}$ for all $x$. From (1.6),

$$
\begin{align*}
&\left|\varphi_{s}\right| \leqslant \frac{C}{P_{s}(\rho)} P_{s}\left(\sqrt{\rho(x)} e^{-a|x|}\right) \leqslant \frac{C}{P_{s}(\rho)} \cdot \sqrt{P_{s}(\rho) P_{s}\left(e^{-2 a|x|}\right)} \\
&=\frac{C}{\sqrt{\rho_{s}}} P_{s}\left(e^{-2 a|x|}\right)^{\frac{1}{2}} \tag{2.12}
\end{align*}
$$

where we have used the Cauchy-Schwarz inequality for $P_{s}$. In order to conclude that $\varphi_{s}$ has subexponential decay relative to $\rho_{s}$, it remains only to note the following: since $P_{s}$ is convolution with a Gaussian of covariance $s \cdot$ Id, there exists $\widetilde{C}=\widetilde{C}_{a, s, n}>0$ such that $P_{s}\left(e^{-2 a|x|}\right)(y) \leqslant \widetilde{C} e^{-2 a|y|}$ for all $y \in \mathbb{R}^{n}$. (The constant $\widetilde{C}=\sup _{s \in\left[s_{0}, s_{1}\right]} \int_{\mathbb{R}^{n}} e^{2 a|x|} \gamma_{s}(x) \mathrm{d} x$ works for all $s \in\left[s_{0}, s_{1}\right]$.)

We still need to bound the partial derivatives of $\varphi_{s}(y)$ with respect to the $s$-variable and $y$-variables. The first-order derivatives are given by formulas (2.8) and (2.9), and higher-order derivatives may be computed by repeated applications of these two formulas. Thus $\partial_{s}^{k} \partial_{y}^{\alpha} Q_{s} \varphi(y)$ can be expressed as a sum with a fixed number of summands. Each of these summands is a product of a term of the form $\frac{1}{s^{m}} Q_{s}(f \varphi)$, where $f$ is a polynomial of degree bounded by $2 k+|\alpha|$, and terms of the form $Q_{s}(p)$ with $p$ a polynomial in the space variables. For any such $p$, the function $Q_{s}(p)$ grows at most polynomially because $Q_{s}\left(|x|^{d}\right)$ does for all $d$. In addition, $f \varphi$ has subexponential decay relative to $\rho$, so by the previous part of the proof, $Q_{s}(f \varphi)$ has subexponential decay relative to $\rho_{s}$. Consequently, each of the summands in $\partial_{s}^{k} \partial_{y}^{\alpha} Q_{s} \varphi(y)$ has subexponential decay relative to $\rho_{s}$, so $\partial_{s}^{k} \partial_{y}^{\alpha} Q_{s} \varphi(y)$ does as well.

Recall the definition (2.1) of $\Gamma_{i}(u, v)$. In the next proposition we rigorously justify the computations in (2.3) and (2.4). We discuss only the case $i=0,1$; while the extension to higher-order carrés des champs presents no particular difficulty, it has been omitted as it is unnecessary for our purposes.

Proposition 2.3. - Fix $s>0$ and $i=0,1$. Suppose that $\varphi$ has subexponential decay relative to $\rho$ and $\varphi_{s}=Q_{s}(\varphi)$. Then equation (2.2) holds for $\varphi_{s}$.

Proof.- Recall that $\Gamma_{0}(u, v)=u v, \Gamma_{1}(u, v)=\nabla u \cdot \nabla v$ and that $\varphi_{s}(y)$ is smooth in $(s, y)$. Therefore,

$$
\begin{equation*}
\partial_{s}\left[\Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \rho_{s}\right]=2 \Gamma_{i}\left(\partial_{s} \varphi_{s}, \varphi_{s}\right) \rho_{s}+\Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \rho_{s} \cdot \partial_{s}\left(\log \rho_{s}\right) \tag{2.13}
\end{equation*}
$$

According to Lemma 2.2 we may bound the expression in (2.13) by the integrable function $C e^{-a|y|}$ for some $C, a>0$, and the bound is locally uniform in $s$. This justifies interchanging differentiation and integration to obtain

$$
\begin{aligned}
\partial_{s} \int_{\mathbb{R}^{n}} \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \rho_{s} & =\int_{\mathbb{R}^{n}} \partial_{s}\left[\Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \rho_{s}\right] \\
& =2 \int_{\mathbb{R}^{n}} \Gamma_{i}\left(\square_{s} \varphi_{s}, \varphi_{s}\right) \rho_{s}+\int_{\mathbb{R}^{n}} \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \frac{\Delta \rho_{s}}{2}
\end{aligned}
$$

where we have used Lemma 2.1 and the heat equation (1.5). Next we need to carry out the integrations by parts of (2.3) and (2.4) and show that no boundary terms arise. When integrating the term $\Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \Delta \rho_{s} / 2$ by parts twice, we encounter the boundary integrands $\Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \nabla \rho_{s}$ and $\nabla \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right)$. $\rho_{s}$. Both of these decay exponentially at infinity, so the integration by parts over $\mathbb{R}^{n}$ introduces no boundary terms, verifying (2.3). In (2.4), we use the integration by parts formula

$$
\int_{\mathbb{R}^{n}} L_{s} \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \cdot \rho_{s}=\int_{\mathbb{R}^{n}} \operatorname{div}\left(\rho_{s} \nabla \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right)\right)=0
$$

which is again justified by the exponential decay of $\nabla \Gamma_{i}\left(\varphi_{s}, \varphi_{s}\right) \cdot \rho_{s}$ at infinity. This completes the proof of (2.2).

We write $H^{1}(\mu)$ for the space of all functions in $L^{2}(\mu)$ whose weak derivatives belong to $L^{2}(\mu)$, equipped with the norm

$$
\|f\|_{H^{1}(\mu)}=\sqrt{\int_{\mathbb{R}^{n}} f^{2} \mathrm{~d} \mu+\int_{\mathbb{R}^{n}}|\nabla f|^{2} \mathrm{~d} \mu}
$$

See e.g. the appendix of [4] and the references therein for information about weak derivatives, the Sobolev space $H^{1}(\mu)$, and for a proof of the fact that the space of smooth, compactly supported functions in $\mathbb{R}^{n}$ is dense in $H^{1}(\mu)$.

ThEOREM 2.4. - Let $\mu$ be an absolutely-continuous, log-concave probability measure on $\mathbb{R}^{n}$ and let $0 \not \equiv \varphi \in H^{1}(\mu)$. Then with $\varphi_{s}=Q_{s} \varphi$, the Rayleigh quotient

$$
\begin{equation*}
R_{\varphi}(s)=\frac{\int_{\mathbb{R}^{n}}\left|\nabla \varphi_{s}\right|^{2} \mathrm{~d} \mu_{s}}{\int_{\mathbb{R}^{n}} \varphi_{s}^{2} \mathrm{~d} \mu_{s}} \tag{2.14}
\end{equation*}
$$

is non-increasing in $s \in[0, \infty)$. Consequently, the function $\log \left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}$ is convex in $s \in[0, \infty)$.

For the proof of Theorem 2.4 we require the following technical lemma:
Lemma 2.5 .
(i) For any fixed $s \geqslant 0$, the quantities $R_{\varphi}(s)$ and $\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}$ depend continuously on $\varphi \in H^{1}(\mu) \backslash\{0\}$.
(ii) For any $0 \not \equiv \varphi \in H^{1}(\mu)$,

$$
\begin{equation*}
R_{\varphi}(0)=\lim _{s \rightarrow 0^{+}} R_{\varphi}(s) \quad \text { and } \quad\|\varphi\|_{L^{2}(\mu)}=\lim _{s \rightarrow 0^{+}}\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)} \tag{2.15}
\end{equation*}
$$

Proof. - We first prove part (i). Let $\varphi \in L^{2}(\mu) \backslash\{0\}$. By Cauchy-Schwarz, we have the pointwise bound

$$
\begin{equation*}
\varphi_{s}^{2} \rho_{s}=\frac{P_{s}^{2}(\rho \varphi)}{P_{s}(\rho)} \leqslant \frac{P_{s}\left(\varphi^{2} \rho\right) P_{s}(\rho)}{P_{s}(\rho)}=P_{s}\left(\varphi^{2} \rho\right) \tag{2.16}
\end{equation*}
$$

Moreover, the log-concavity of $\mu$ implies that whenever $\varphi \in H^{1}(\mu)$,

$$
\begin{equation*}
\left|\nabla \varphi_{s}\right|^{2} \rho_{s} \leqslant P_{s}\left(|\nabla \varphi|^{2} \rho\right) \tag{2.17}
\end{equation*}
$$

Indeed, the probability density $p_{s, y}$ from (2.5) is "more log-concave than $\gamma_{s}$," in the sense that $p_{s, y} / \gamma_{s}$ is log-concave. The Brascamp-Lieb inequality (see e.g., $[2, \S 4.9]$ ) thus implies that the Poincare constant of the probability density $p_{s, y}$ is at most $s$. That is, letting $X$ be a random vector with density $p_{s, y}$ and $f$ a weakly differentiable function with $\mathbb{E}|f(X)|^{2}<\infty$ and $\mathbb{E}|\nabla f(X)|^{2}<\infty$,

$$
\begin{equation*}
\operatorname{Var} f(X) \leqslant s \cdot \mathbb{E}|\nabla f(X)|^{2} \tag{2.18}
\end{equation*}
$$

Hence, by Lemma 2.1, for any $\theta \in S^{n-1}=\left\{x \in \mathbb{R}^{n} ;|x|=1\right\}$,

$$
\begin{aligned}
& \nabla \varphi_{s} \cdot \theta=\frac{\mathbb{E}(X \cdot \theta) \varphi(X)-\mathbb{E}(X \cdot \theta) \mathbb{E} \varphi(X)}{s} \\
& \leqslant \frac{\sqrt{\operatorname{Var}(X \cdot \theta) \operatorname{Var}(\varphi(X))}}{s} \leqslant \sqrt{\mathbb{E}|\nabla \varphi(X)|^{2}}
\end{aligned}
$$

which implies $(2.17)$ since $\mathbb{E}|\nabla \varphi(X)|^{2}=Q_{s}\left(|\nabla \varphi|^{2}\right)=P_{s}\left(|\nabla \varphi|^{2} \rho\right) / \rho_{s}$. By integrating over $\mathbb{R}^{n}$, the inequalities (2.16) and (2.17) imply that

$$
\begin{array}{ll}
\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)} \leqslant\|\varphi\|_{L^{2}(\mu)} & \text { for } \varphi \in L^{2}(\mu) \\
\left\|\nabla \varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}=\sqrt{\int_{\mathbb{R}^{n}}\left|\nabla \varphi_{s}\right|^{2} \mathrm{~d} \mu_{s}} \leqslant\|\varphi\|_{H^{1}(\mu)} & \text { for } \varphi \in H^{1}(\mu)
\end{array}
$$

Consequently, the functional $\varphi \mapsto\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}$ is 1-Lipschitz in $L^{2}(\mu)$, while the functional $\varphi \mapsto\left\|\nabla \varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}$ is 1-Lipschitz in $H^{1}(\mu)$. In particular, for any fixed $s \geqslant 0$, the quantities $R_{\varphi}(s)$ and $\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}$ depend continuously on $\varphi \in H^{1}(\mu) \backslash\{0\}$, proving (i).

For part (ii), note that $\varphi \mapsto R_{\varphi}(s)$ is locally uniformly continuous in $H^{1}(\mu) \backslash\{0\}$, being the quotient of two positive, 1-Lipschitz functions. Hence, it suffices to prove (2.15) for $\varphi$ in a dense subset of $H^{1}(\mu) \backslash\{0\}$. We may thus assume that $\varphi$ is smooth and compactly supported. We claim that for almost every $y \in \mathbb{R}^{n}$,

$$
\begin{equation*}
\varphi_{s}^{2}(y) \rho_{s}(y) \xrightarrow{s \rightarrow 0} \varphi^{2}(y) \rho(y) \quad \text { and } \quad\left|\nabla \varphi_{s}(y)\right|^{2} \rho_{s}(y) \xrightarrow{s \rightarrow \infty}|\nabla \varphi(y)|^{2} \rho(y) . \tag{2.19}
\end{equation*}
$$

Let $K=\left\{x \in \mathbb{R}^{n} ; \rho(x)>0\right\}$. In proving (2.19), we may thus assume that $y \notin \partial K$, since the boundary of the convex set $K$ has Lebesgue measure zero. If $y \notin \bar{K}$ then $\rho$ vanishes in a neighborhood of $y$, hence

$$
P_{s}\left(\varphi^{2} \rho\right)(y) \xrightarrow{s \rightarrow 0} 0 \quad \text { and } \quad P_{s}\left(|\nabla \varphi|^{2} \rho\right)(y) \xrightarrow{s \rightarrow 0} 0,
$$

and (2.19) follows from the bounds (2.16) and (2.17). As $\rho$ is log-concave, it is locally Lipschitz on $K$, so by the Rademacher theorem, $\rho$ is differentiable almost everywhere in the interior of $K$. It thus suffices to prove (2.19) for $y \in K$ such that $\rho$ is differentiable at $y$. Differentiating $\varphi_{s}$ yields

$$
\begin{equation*}
\nabla \varphi_{s}=\nabla\left(\frac{P_{s}(\varphi \rho)}{P_{s}(\rho)}\right)=\frac{\nabla P_{s}(\varphi \rho)}{P_{s}(\rho)}-\frac{P_{s}(\varphi \rho) \nabla \rho_{s}}{\rho_{s}^{2}} \quad(s>0) \tag{2.20}
\end{equation*}
$$

It is a property of the heat semigroup that if $f$ is a bounded measurable function differentiable at a point $y \in \mathbb{R}^{n}$, then $\nabla P_{s}(f)(y) \rightarrow \nabla f(y)$ as $s \rightarrow 0$; this is easily shown by writing $\nabla P_{s} f=f * \nabla \gamma_{s}$ and approximating $f$ by its first-order Taylor polynomial. Applying this to the functions $\rho$ and $\varphi \rho$ which are bounded in $\mathbb{R}^{n}$ and differentiable at $y$, we obtain $\nabla \rho_{s}(y) \rightarrow \nabla \rho(y)$ and $\nabla P_{s}(\varphi \rho)(y) \rightarrow \nabla(\varphi \rho)(y)$ as $s \rightarrow 0$. Moreover, $\varphi_{s}(y) \rightarrow \varphi(y), \rho_{s}(y) \rightarrow \rho(y)$ because $\rho$ and $\varphi \rho$ are continuous at $y$ and bounded in $\mathbb{R}^{n}$. It follows that $\nabla \varphi_{s}(y) \rightarrow \nabla \varphi(y)$, completing the proof of (2.19).

Finally, the functions $|\nabla \varphi|^{2} \rho$ and $\varphi^{2} \rho$ are bounded and compactly supported. Hence, for $s \in(0,1]$ the Gaussian convolutions $P_{s}\left(|\nabla \varphi|^{2} \rho\right)$ and $P_{s}\left(\varphi^{2} \rho\right)$ are bounded by $C e^{-|y|^{2} / 2}$ in $\mathbb{R}^{n}$ for some constant $C$ that does not depend on $s$. From (2.16), (2.17), the dominated convergence theorem, and (2.19), we obtain

$$
\int_{\mathbb{R}^{n}} \varphi_{s}^{2} \rho_{s} \xrightarrow{s \rightarrow 0} \int_{\mathbb{R}^{n}} \varphi^{2} \rho \quad \text { and } \quad \int_{\mathbb{R}^{n}}\left|\nabla \varphi_{s}\right|^{2} \rho_{s} \xrightarrow{s \rightarrow 0}|\nabla \varphi|^{2} \rho,
$$

completing the proof of (2.15).
Remark 2.6. - Inequality (2.17) states that $\left|\nabla \varphi_{s}\right|^{2} \leqslant Q_{s}\left(|\nabla \varphi|^{2}\right)$. Using the interpretation in [37, §3.1] and arguing as in [37, §3.2], one may prove the stronger gradient bound

$$
\left|\nabla \varphi_{s}\right| \leqslant Q_{s}(|\nabla \varphi|)
$$

which we do not need here.

It follows from (2.1) and a straightforward computation that

$$
\begin{align*}
\Gamma_{2}(u, u)=\square_{s}|\nabla u|^{2}-2\left\langle\nabla \square_{s} u\right. & , \nabla u\rangle \\
& =\left\|\nabla^{2} u\right\|_{H S}^{2}-2\left\langle\nabla^{2}\left(\log \rho_{s}\right) \nabla u, \nabla u\right\rangle \tag{2.21}
\end{align*}
$$

On the other hand, the Bochner formula for the differential operator $L_{s}=$ $L_{\mu_{s}}$ states that for any smooth, compactly supported function $u: \mathbb{R}^{n} \rightarrow \mathbb{R}$,

$$
\begin{align*}
\int_{\mathbb{R}^{n}}\left(-L_{s}\right)^{2} u \cdot u \mathrm{~d} \mu_{s} & =\int_{\mathbb{R}^{n}}\left(L_{s} u\right)^{2} \mathrm{~d} \mu_{s} \\
& =\int_{\mathbb{R}^{n}}\left[\left\|\nabla^{2} u\right\|_{H S}^{2}-\left\langle\nabla^{2}\left(\log \rho_{s}\right) \nabla u, \nabla u\right\rangle\right] \mathrm{d} \mu_{s} \tag{2.22}
\end{align*}
$$

See $[2, \S 1.16 .1]$ for a proof of (2.22). Formula (2.22) remains valid when $u$ and its partial derivatives are smooth functions with subexponential decay relative to $\rho_{s}$, since the integration by parts yield no boundary terms as in the proof of Proposition 2.3. Thanks to Lemma 2.2, we know that formula (2.22) is valid for $u=Q_{s} \varphi$ whenever $\varphi$ has subexponential decay relative to $\rho$.

The integrand on the right-hand side of (2.22) is almost identical to the expression in (2.21), the only difference is the coefficient 2 in front of the second summand.

Proof of Theorem 2.4. - When $u$ is a smooth function such that $u$ and its partial derivatives have subexponential decay relative to $\rho_{s}$, we write for $i=1,2$,

$$
\|u\|_{\dot{H}^{i}\left(\mu_{s}\right)}=\sqrt{\int_{\mathbb{R}^{n}}\left(-L_{s}\right)^{i} u \cdot u \mathrm{~d} \mu}
$$

Thus $\|u\|_{\dot{H}^{1}\left(\mu_{s}\right)}^{2}=\int_{\mathbb{R}^{n}}|\nabla u|^{2} \mathrm{~d} \mu_{s}$. The operator $L_{s}$ is initially defined by the formula $L_{s} u=\Delta u+\nabla \log \rho_{s} \cdot \nabla u$ assuming $u$ and its partial derivatives have subexponential decay relative to $\rho_{s}$. This operator is essentially self-adjoint and negative semi-definite in $L^{2}\left(\mu_{s}\right)$ (e.g., [2, Corollary 3.2.2]). Hence, by the spectral theorem and the Cauchy-Schwarz inequality,

$$
\begin{equation*}
\|u\|_{\dot{H}^{1}\left(\mu_{s}\right)}^{2} \leqslant\|u\|_{\dot{H}^{2}\left(\mu_{s}\right)} \cdot\|u\|_{L^{2}\left(\mu_{s}\right)} \tag{2.23}
\end{equation*}
$$

Consider first the case where $0 \not \equiv \varphi \in H^{1}(\mu)$ has subexponential decay relative to $\rho$ and $s>0$. Thanks to Proposition 2.3 we may apply (2.2) and compute that

$$
\frac{\mathrm{d}}{\mathrm{~d} s} R_{\varphi}(s)=\frac{\left\|\varphi_{s}\right\|_{\dot{H}^{1}\left(\mu_{s}\right)}^{4}-\int_{\mathbb{R}^{n}} \Gamma_{2}\left(\varphi_{s}, \varphi_{s}\right) \mathrm{d} \mu_{s} \cdot\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}^{2}}{\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}^{4}}
$$

Hence, from (2.21) and (2.22),

$$
\begin{aligned}
\frac{\mathrm{d}}{\mathrm{~d} s} R_{\varphi}(s)=\frac{\left\|\varphi_{s}\right\|_{\dot{H}^{1}\left(\mu_{s}\right)}^{4}-\left\|\varphi_{s}\right\|_{\dot{H}^{2}\left(\mu_{s}\right)}^{2}\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}^{2}}{\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}^{4}} & \\
& +\frac{\int_{\mathbb{R}^{n}}\left\langle\left(\nabla^{2} \log \rho_{s}\right) \nabla \varphi_{s}, \nabla \varphi_{s}\right\rangle \mathrm{d} \mu_{s}}{\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}^{2}} .
\end{aligned}
$$

By $\log$-concavity $\nabla^{2} \log \rho_{s} \leqslant 0$. Hence we conclude from (2.23) that

$$
\frac{\mathrm{d}}{\mathrm{~d} s} R_{\varphi}(s) \leqslant 0
$$

Therefore $R_{\varphi}(s)$ is non-increasing in $s \in(0, \infty)$. It follows from (2.2) that

$$
\partial_{s} \log \|\varphi\|_{L^{2}\left(\mu_{s}\right)}=-R_{\varphi}(s),
$$

and consequently $\log \|\varphi\|_{L^{2}\left(\mu_{s}\right)}$ is convex in $s \in(0, \infty)$. Lemma 2.5 now implies that $R_{\varphi}(s)$ is decreasing in $s \in[0, \infty)$ and $\log \left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}$ is convex in $s \in[0, \infty)$.

Finally, compactly supported smooth functions, which certainly have subexponential decay relative to $\rho$, are dense in $H^{1}(\mu)$. The Rayleigh quotient and $\left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}$ are continuous on $H^{1}(\mu) \backslash\{0\}$ by Lemma 2.5, hence we obtain that $R_{\varphi}(s)$ is non-increasing and $\log \left\|\varphi_{s}\right\|_{L^{2}\left(\mu_{s}\right)}$ is convex in $s \in[0, \infty)$ for any $0 \not \equiv \varphi \in H^{1}(\mu)$.

Using the min-max characterization of eigenvalues, we derive our main result as a corollary to Theorem 2.4.

Proof of Theorem 1.1. - We may set $s=1$, since $\mu_{s}=\mu * \gamma$ for $s=1$. We may assume that $\mu$ is absolutely continuous, as otherwise we may pass to a lower dimension thanks to the well-known fact that the Poincaré constant of a Cartesian product of two measures is the maximum of the Poincare constants of the factors. The Poincaré constant of $\mu$, which is finite and positive (see [6]), satisfies

$$
\frac{1}{C_{P}(\mu)}=\inf \left\{R_{\varphi}(0) ; 0 \not \equiv \varphi \in H^{1}(\mu), \int_{\mathbb{R}^{n}} \varphi \mathrm{~d} \mu=0\right\}
$$

and similarly for $\mu_{s}$. For any $\varepsilon>0$ there exists $0 \not \equiv \varphi \in H^{1}(\mu)$ with $\int \varphi \mathrm{d} \mu=$ 0 such that $R_{\varphi}(0)<C_{P}(\mu)^{-1}+\varepsilon$. Since $\int \varphi_{s} \mathrm{~d} \mu_{s}=\int \varphi \mathrm{d} \mu=0$, we deduce from Theorem 2.4 that,

$$
\frac{1}{C_{P}\left(\mu_{s}\right)} \leqslant R_{\varphi}(s) \leqslant R_{\varphi}(0)<\frac{1}{C_{P}(\mu)}+\varepsilon
$$

As $\varepsilon>0$ was arbitrary, inequality (1.3) is proven.
Next, assume that $L_{\mu}$ has discrete spectrum, and let $k \geqslant 1$. There exists a $(k+1)$-dimensional subspace $E \subseteq H^{1}(\mu)$ such that $R_{\varphi}(0) \leqslant \lambda_{k}^{(\mu)}$ for any
$0 \not \equiv \varphi \in E$. For $s>0$ the linear operator $Q_{s}$ defined in (1.6) is one-to-one in $L^{1}(\mu)$. (Indeed, given $P_{s}(\varphi \rho)$ we may recover the Fourier transform of $\varphi \rho \in L^{1}\left(\mathbb{R}^{n}\right)$ which determines $\varphi \in L^{1}(\mu)$.) Hence

$$
E_{s}=\left\{Q_{s} \varphi ; \varphi \in E\right\}
$$

is a $(k+1)$-dimensional subspace, and $R_{\varphi}(s) \leqslant R_{\varphi}(0) \leqslant \lambda_{k}^{(\mu)}$ for all $\varphi \in E$. In other words, there exists a $(k+1)$-dimensional subspace $E_{s} \subseteq H^{1}\left(\mu_{s}\right)$ on which the Rayleigh quotient is at most $\lambda_{k}^{(\mu)}$. By the min-max characterization of eigenvalues,

$$
\lambda_{k}^{\left(\mu_{s}\right)} \leqslant \lambda_{k}^{(\mu)}
$$

completing the proof.
The proof of Theorem 1.1 clearly shows that $C_{P}\left(\mu * \gamma_{s}\right) \geqslant C_{P}(\mu)$ for all $s>0$, so by the semigroup property $s \mapsto C_{P}\left(\mu * \gamma_{s}\right)$ is non-decreasing in $s \in[0, \infty)$.

Remark 2.7. - Let $\mu$ be a log-concave probability measure in $\mathbb{R}^{n}$ with density $\rho=e^{-W}$, where $W$ is a smooth function such that

$$
\lim _{x \rightarrow \infty} \frac{|\nabla W(x)|^{2}}{2}-\Delta W(x)=\infty
$$

In this case, we have the strict inequality

$$
\begin{equation*}
\lambda_{k}^{(\mu * \gamma)}<\lambda_{k}^{(\mu)} \quad(k=1,2, \ldots) \tag{2.24}
\end{equation*}
$$

In order to prove (2.24), we first observe that $\nabla^{2} \log \rho_{s}(y)<0$ for all $y \in$ $\mathbb{R}^{n}$ as follows from the equality case of the Brascamp-Lieb inequality or from $[16,18]$. Arguing as in the proof of Theorem 2.4 and using the fact that $\nabla \varphi_{s} \not \equiv 0$ as $\varphi_{s}$ is non-constant, we conclude that $\mathrm{d} R_{\varphi}(s) / \mathrm{d} s<0$ whenever $0 \not \equiv \varphi \in H^{1}(\mu)$ has subexponential decay relative to $\rho$.

Therefore (2.24) would follow from Theorem 2.4, as in the proof of Theorem 1.1 above, had we known that any eigenfunction $\varphi$ of $L_{\mu}$ has subexponential decay relative to $\rho$.

Indeed, let $A: L^{2}\left(\mathbb{R}^{n}\right) \xrightarrow{\sim} L^{2}(\mu)$ be the isometry given by $A(g)=e^{\frac{W}{2}} g$. It is well-known and easy to verify that $A^{-1} L_{\mu} A$ is the Schrödinger operator

$$
-\Delta+\frac{|\nabla W|^{2}}{4}-\frac{\Delta W}{2}
$$

which is of the form $-\Delta+V$ with $V \geqslant 0$ and $V \rightarrow \infty$ as $x \rightarrow \infty$. By results on the decay of eigenfunctions of Schrödinger operators [36, Theorem XIII.70], the function $A^{-1} \varphi$ has subexponential decay at infinity, and hence $\varphi$ has subexponential decay relative to $\rho$.

## 3. A contraction transporting $\mu * \gamma$ to $\mu$

In this section we prove Theorem 1.2 using the arguments of Kim and Milman [23]. To begin with, we work with a log-concave probability measure $\mu$ with a smooth, strictly positive density $\rho$ on $\mathbb{R}^{n}$. We furthermore make the regularity assumption that there exists $\varepsilon>0$ such that

$$
\begin{equation*}
-\nabla^{2} \log \rho(y) \leqslant \frac{1}{\varepsilon} \cdot \operatorname{Id} \quad\left(y \in \mathbb{R}^{n}\right) \tag{3.1}
\end{equation*}
$$

We shall later remove these assumptions on $\rho$. As above, for $s \geqslant 0$ we write $\mu_{s}=\mu * \gamma_{s}$ and $\rho_{s}$ is the density of $\mu_{s}$. Thus $\rho_{s}$ is smooth, positive and log-concave in $\mathbb{R}^{n}$. For $s \geqslant 0$ consider the advection field

$$
\begin{equation*}
W_{s}(y)=-\frac{1}{2} \nabla \log \rho_{s}(y) \tag{3.2}
\end{equation*}
$$

The "physical" interpretation of this vector field is as follows. One of the derivations of the heat equation is based on Fourier's law, according to which the flux of heat across a tiny surface in a short time interval is proportional to the temperature gradient across the surface. If we think of the heat as carried by a fluid of particles with density $\rho(x, t)$, this means that the current of heat is proportional to $-\nabla \rho$ (we take $\frac{1}{2}$ to be the constant of proportionality); since the current of heat is simply $\rho v$, where $v(x, t)$ is the bulk velocity of the fluid, we obtain $v=-\frac{1}{2} \frac{\nabla \rho}{\rho}$, which is (3.2). For more details, see [38, §5.4].

With this point of view, the trajectory of a particle located at time $s=0$ at the point $y \in \mathbb{R}^{n}$ is the curve $s \mapsto T_{s}(y)$ where

$$
\left\{\begin{align*}
\frac{\mathrm{d}}{\mathrm{~d} s} T_{s}(y) & =W_{s}\left(T_{s}(y)\right), \quad s \geqslant 0  \tag{3.3}\\
T_{0}(y) & =y
\end{align*}\right.
$$

Lemma 3.1. - Under the above assumptions on $\rho$, the ordinary differential equation (3.3) determines the family of maps $\left(T_{s}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}\right)_{s \geqslant 0}$. These maps are all diffeomorphisms, and $T_{s}(y)$ is smooth in $(s, y) \in[0, \infty) \times \mathbb{R}^{n}$.

Proof. - Since $\rho_{s}=\rho * \gamma_{s}$, the function $\rho_{s}(y)$ is smooth and positive in $(s, y) \in[0, \infty) \times \mathbb{R}^{n}$. Therefore $W_{s}(y)$ is smooth in $(s, y) \in[0, \infty) \times \mathbb{R}^{n}$ as well. It remains to show that $W_{s}$ is $1 /(2 \varepsilon)$-Lipschitz on $\mathbb{R}^{n}$ for any $s \geqslant 0$. Once this is shown, the standard theory of ordinary differential equations implies the existence and uniqueness of solutions to (3.3) and their smooth dependence on initial conditions (e.g., [22, Chapter V]). The fact that the $T_{s}$ are diffeomorphisms follows from the theory of flows of time-dependent vector fields on manifolds (e.g., [29, Chapter 17]).

We need to compute the derivative of $W_{s}$. As in the beginning of the proof of Lemma 2.2 above, by differentiating (2.6) we see that for any $s>0$
and $y \in \mathbb{R}^{n}$,

$$
\begin{equation*}
D W_{s}(y)=-\frac{1}{2} \nabla^{2} \log \rho_{s}(y)=\frac{s \cdot \operatorname{Id}-\operatorname{Cov}\left(p_{s, y}\right)}{2 s^{2}} \tag{3.4}
\end{equation*}
$$

where $\operatorname{Cov}\left(p_{s, y}\right) \in \mathbb{R}^{n \times n}$ is the covariance matrix of the probability density $p_{s, y}$. Since $\rho_{s}$ is log-concave, the differential $D W_{s}$ is a symmetric positive semidefinite matrix. From (2.5) and the regularity assumption (3.1) we see that for $s \geqslant 0$ and $x \in \mathbb{R}^{n}$,

$$
\begin{equation*}
-\nabla^{2} \log p_{s, y}(x) \leqslant\left(\frac{1}{\varepsilon}+\frac{1}{s}\right) \cdot \operatorname{Id} \tag{3.5}
\end{equation*}
$$

in the sense of symmetric matrices. It is well-known (see [7, Theorem 5.4]) that (3.5) implies that

$$
\begin{equation*}
\left(\frac{1}{\varepsilon}+\frac{1}{s}\right)^{-1} \cdot \operatorname{Id} \leqslant \operatorname{Cov}\left(p_{s, y}\right) \tag{3.6}
\end{equation*}
$$

From (3.4) and (3.6) we deduce the pointwise bound

$$
\left\|D W_{s}(y)\right\|_{o p} \leqslant \frac{1}{2(s+\varepsilon)} \quad\left(y \in \mathbb{R}^{n}\right)
$$

where $\|\cdot\|_{o p}$ is the operator norm. This bound clearly applies also for $s=0$. Therefore $W_{s}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is $1 /(2 \varepsilon)$-Lipschitz for any $s \geqslant 0$, completing the proof.

As explained in Kim and Milman [23], the diffeomorphism $T_{s}$ is an expansion, i.e., $\left|T_{s}(x)-T_{s}(y)\right| \geqslant|x-y|$ for all $x, y$ and $s$. In order to prove this, we show that everywhere in $\mathbb{R}^{n}$,

$$
\begin{equation*}
\left(D T_{s}\right)^{*}\left(D T_{s}\right) \geqslant \mathrm{Id} \tag{3.7}
\end{equation*}
$$

Inequality (3.7) is certainly true when $s=0$, while the fact that $D W_{s}$ is positive semidefinite implies that

$$
\frac{\partial}{\partial s}\left(D T_{s}\right)^{*}\left(D T_{s}\right)=2\left(D T_{s}\right)^{*}\left(D W_{s}\right) D T_{s} \geqslant 0
$$

Therefore (3.7) holds true. This implies that $\left\|D\left(T_{s}^{-1}\right)\right\|_{o p} \leqslant 1$, hence $T_{s}^{-1}$ is a contraction and $T_{s}$ is an expansion. Next, from (3.2) and the heat equation $\partial \rho_{s} / \partial s=\Delta \rho_{s} / 2$ we obtain the linear transport equation (also known as the continuity equation),

$$
\frac{\partial \rho_{s}}{\partial s}+\operatorname{div}\left(\rho_{s} W_{s}\right)=0 \quad\left(s \geqslant 0, y \in \mathbb{R}^{n}\right)
$$

The continuity equation implies that $\rho_{s}$ is the density of the pushforward of $\mu$ under the diffeomorphism $T_{s}$ (see e.g. [38, Theorem 5.34]). Consequently, the map $T_{s}^{-1}$ is a contraction that pushes forward $\mu_{s}$ to $\mu$.

Proof of Theorem 1.2. - Set $s=1$ so that $\mu_{s}=\mu * \gamma$. We have just established the existence of a contraction transporting $\mu_{s}$ to $\mu$ under the additional requirement that $\mu$ admits a smooth, positive density satisfying the regularity assumption (3.1).

Consider now the case where $\mu$ is an arbitrary absolutely-continuous, log-concave probability measure in $\mathbb{R}^{n}$. For any $\varepsilon>0$, the measure $\mu_{\varepsilon}=$ $\mu * \gamma_{\varepsilon}$ has a smooth, positive, log-concave density satisfying the regularity assumption (3.1), as follows from the computation in (3.4) above. Hence there exists a contraction transporting $\mu_{\varepsilon} * \gamma$ to $\mu_{\varepsilon}$. By [23, Lemma 3.3], in order to show that there exists a contraction from $\mu * \gamma$ to $\mu$, it suffices to show that

$$
\begin{equation*}
\mu_{\varepsilon} \xrightarrow{\varepsilon \rightarrow 0^{+}} \mu \tag{3.8}
\end{equation*}
$$

in the total variation metric, and that $\mu_{\varepsilon} * \gamma \longrightarrow \mu * \gamma$ as $\varepsilon \rightarrow 0$ in the weak topology. Since $\mu_{\varepsilon} * \gamma=(\mu * \gamma)_{\varepsilon}$ and since convergence in total variation implies convergence in the weak topology, it suffices to prove (3.8). Thus we need to show that

$$
\begin{equation*}
\int_{\mathbb{R}^{n}}\left|\rho_{\varepsilon}(x)-\rho(x)\right| \mathrm{d} x \xrightarrow{\varepsilon \rightarrow 0^{+}} 0 . \tag{3.9}
\end{equation*}
$$

Arguing as in (2.11) and the paragraph following (2.12) above, we know that there exist $a, b>0$ such that $\rho_{\varepsilon}(x) \leqslant a e^{-b|x|}$ for all $x \in \mathbb{R}^{n}$ and $0 \leqslant \varepsilon \leqslant 1$, with $\rho_{0}=\rho$. Since $\rho$ is continuous almost everywhere in $\mathbb{R}^{n}$, the integrand in (3.9) converges to zero almost everywhere, and (3.9) follows from the dominated convergence theorem.

Thus the conclusion of the theorem is valid when $\mu$ is an absolutely continuous, log-concave probability measure. Finally, if $\mu$ is not absolutely continuous, then we may project to a lower dimension using an orthogonal projection, which is a contraction, and reduce matters to the absolutely continuous case.

Theorem 1.2 implies that for $\nu=\mu * \gamma$ and $0 \not \equiv \varphi \in H^{1}(\nu)$ we have the following inequality between Rayleigh quotients:

$$
\begin{equation*}
\frac{\int_{\mathbb{R}^{n}}|\nabla(\varphi \circ T)|^{2} \mathrm{~d} \nu}{\int_{\mathbb{R}^{n}}(\varphi \circ T)^{2} \mathrm{~d} \nu} \leqslant \frac{\int_{\mathbb{R}^{n}}|(\nabla \varphi) \circ T|^{2} \mathrm{~d} \nu}{\int_{\mathbb{R}^{n}}(\varphi \circ T)^{2} \mathrm{~d} \nu}=\frac{\int_{\mathbb{R}^{n}}|\nabla \varphi|^{2} \mathrm{~d} \mu}{\int_{\mathbb{R}^{n}} \varphi^{2} \mathrm{~d} \mu} . \tag{3.10}
\end{equation*}
$$

We may now repeat the proof of Theorem 1.1 from Section 2, with the linear map $\varphi \mapsto \varphi \circ T$ playing the role of the linear map $\varphi \mapsto Q_{1} \varphi$. This yields another proof of Theorem 1.1, relying on (3.10) in place of Theorem 2.4.

## 4. A Bayesian interpretation of Eldan's stochastic localization

Eldan's stochastic localization technique was introduced by Eldan in [19] and developed since then by several authors in different settings [14, 20, $24,30]$. The method has turned out to be useful in particular for the study of log-concave measures, culminating thus far in the breakthrough result of Chen [14] showing that the isotropic constant grows more slowly than any power of the dimension. In this section, we give a "Bayesian" interpretation of Eldan's stochastic localization relating it to the heat flow and to the operator $Q_{s}$ introduced above, as well as to the Föllmer drift in the theory of Wiener space. It was this line of development which led us to the results announced in the introduction; however, this section may be read independently.

We refer to [35] for background on stochastic processes. Let $\mu$ be an absolutely continuous probability measure on $\mathbb{R}^{n}$ with density $p_{0}$ and with finite second moments. Let $\left(W_{t}\right)_{t \geqslant 0}$ be a standard Brownian motion on $\mathbb{R}^{n}$ with $W_{0}=0$.

The stochastic localization process, in the version introduced by [30], is a density-valued stochastic process $p_{t}$ driven by $W_{t}$, defined as follows: for every $x \in \mathbb{R}^{n}$, the process $\left(p_{t}(x)\right)_{t \geqslant 0}$ is the solution to the stochastic differential equation

$$
\begin{equation*}
\mathrm{d} p_{t}(x)=p_{t}(x)\left\langle x-a_{t}, \mathrm{~d} W_{t}\right\rangle \tag{4.1}
\end{equation*}
$$

with initial condition $p_{0}$, where $a_{t}=\int_{\mathbb{R}^{n}} x \cdot p_{t}(x) \mathrm{d} x$ is the barycenter of $p_{t}$. As this equation has no drift term, $p_{t}(x)$ is a martingale, and $p_{t}$ is almost surely a probability density. In particular, $\mathbb{E}\left[p_{t}(x)\right]=p_{0}(x)$, and for any test function $\varphi$, we have $\mathbb{E}_{X \sim p_{0}}[\varphi(X)]=\mathbb{E}\left[\mathbb{E}_{X \sim p_{t}}[\varphi(X)]\right]$.

The process $\left(p_{t}\right)_{t \geqslant 0}$ has another description, as a stochastic "tilt" of $p_{0}$. In this section, for $t \geqslant 0$ and $\theta \in \mathbb{R}^{n}$ let $p_{t, \theta}$ denote the probability density given by

$$
\begin{equation*}
p_{t, \theta}(x)=\frac{1}{Z(t, \theta)} e^{\langle\theta, x\rangle-\frac{t|x|^{2}}{2}} p_{0}(x), \tag{4.2}
\end{equation*}
$$

where $Z(t, \theta)=\int_{\mathbb{R}^{n}} e^{\langle\theta, x\rangle-\frac{t|x|^{2}}{2}} p_{0}(x) \mathrm{d} x$ is a normalization constant. Let $a(t, \theta)$ denote the barycenter of $p_{t, \theta}$, and define the stochastic process $\theta_{t}$ via the differential equation

$$
\begin{equation*}
\mathrm{d} \theta_{t}=a\left(t, \theta_{t}\right) \mathrm{d} t+\mathrm{d} W_{t}, \quad \theta_{0}=0 \tag{4.3}
\end{equation*}
$$

It turns out that when $\theta_{t}$ and $p_{t}$ are driven by the same Brownian motion, $p_{t}$ is precisely equal to $p_{t, \theta_{t}}$. For proofs of these and other formulas relating to the stochastic localization process, and for the application to the KLS conjecture, see [30, 31] or [14].

The Bayesian interpretation of the Eldan process is quite simple: let $X$ be a random vector distributed according to $\mu$, independent of the Brownian motion $\left(W_{t}\right)_{t \geqslant 0}$. Denote

$$
\begin{equation*}
\widetilde{\theta}_{t}=t X+W_{t} \quad(t \geqslant 0) \tag{4.4}
\end{equation*}
$$

Our main observations are the following two claims:
(i) The process $\left(\widetilde{\theta}_{t}\right)_{t \geqslant 0}$ coincides in law with the process $\left(\theta_{t}\right)_{t \geqslant 0}$ which solves (4.3) above.
(ii) For any fixed $t>0$ and $\theta \in \mathbb{R}^{n}$, the probability density $p_{t, \theta}$ on $\mathbb{R}^{n}$ is precisely the conditional probability distribution of $X$ given that $\widetilde{\theta}_{t}=\theta$.

Thus, when we observe the tilt process $\left(\theta_{t}\right)_{t \geqslant 0}$, we actually see a Brownian motion with a constant drift $X$ which is unknown, but whose prior distribution is known to us. Moreover, the posterior probability density for the unknown drift $X$ given the observation of the process $\left(\theta_{s}\right)_{0 \leqslant s \leqslant t}$ until time $t$ depends only on the state of the process at time $t$, and is equal to $p_{t, \theta_{t}}$.

In the following proposition we prove these two claims. For $T>0$ let $\mathcal{V}_{T}=C_{0}\left([0, T], \mathbb{R}^{n}\right)$ be the Wiener space of $\mathbb{R}^{n}$-valued continuous functions $\left(W_{t}\right)_{0 \leqslant t \leqslant T}$ with $W_{0}=0$. Slightly abusing notation, we write $\gamma_{T}$ for the Wiener probability measure on $\mathcal{V}_{T}$ and $\left\{\mathcal{F}_{t}\right\}_{0 \leqslant t \leqslant T}$ for the natural filtration, i.e., $\mathcal{F}_{t}$ is the $\sigma$-algebra generated by $\left(W_{s}\right)_{0 \leqslant s \leqslant t}$.

Proposition 4.1. - Let $\mu$ be a probability measure on $\mathbb{R}^{n}$ which is absolutely continuous with respect to the Lebesgue measure $\lambda$. Fix $T>0$, and consider the space $\Omega=\mathbb{R}^{n} \times \mathcal{V}_{T}$ and the transformation $\tau: \Omega \rightarrow \Omega$ given by

$$
\tau\left(x,\left(W_{t}\right)_{0 \leqslant t \leqslant T}\right)=\left(x,\left(W_{t}+t x\right)_{0 \leqslant t \leqslant T}\right) .
$$

Write $\nu=\tau_{*}\left(\mu \otimes \gamma_{T}\right)$. Then,
(i) The stochastic process $\left(\widetilde{\theta}_{t}\right)_{t \geqslant 0}$ described in (4.4) coincides in law with the Itô process $\left(\theta_{t}\right)_{t \geqslant 0}$ defined as the solution to the stochastic differential equation (4.3).
(ii) The measure $\nu$ is absolutely continuous with respect to $\lambda \otimes \gamma_{T}$ on $\Omega$ with density

$$
\frac{\mathrm{d} \nu}{\mathrm{~d}\left(\lambda \otimes \gamma_{T}\right)}(x, \widetilde{\theta})=p_{0}(x) e^{\left\langle\tilde{\theta}_{T}, x\right\rangle-\frac{T|x|^{2}}{2}}
$$

for $x \in \mathbb{R}^{n}$ and $\tilde{\theta}=\left(\widetilde{\theta}_{t}\right)_{0 \leqslant t \leqslant T} \in \mathcal{V}_{T}$. Consequently, when $\left(X,\left(\widetilde{\theta}_{t}\right)_{0 \leqslant t \leqslant T}\right)$ is the stochastic process described in (4.4), the conditional distribution of $X$ with respect to $\widetilde{\theta}=\left(\widetilde{\theta}_{t}\right)_{0 \leqslant t \leqslant T}$ is given by
the probability density

$$
\begin{equation*}
q_{T}(x \mid \widetilde{\theta})=\frac{p_{0}(x) e^{\left\langle\tilde{\theta}_{T}, x\right\rangle-\frac{T|x|^{2}}{2}}}{\int_{\mathbb{R}^{n}} p_{0}(y) e^{\left\langle\tilde{\theta}_{T}, y\right\rangle-\frac{T|y|^{2}}{2}} \mathrm{~d} y}=p_{T, \tilde{\theta}_{T}}(x) \quad\left(x \in \mathbb{R}^{n}\right) \tag{4.5}
\end{equation*}
$$

Proof. - We first prove (ii). For $x \in \mathbb{R}^{n}$, let $\tau_{x}: \mathcal{V}_{T} \rightarrow \mathcal{V}_{T}$ be defined by $\tau_{x}\left(\left(W_{t}\right)_{t \leqslant T}\right)=\left(W_{t}+t x\right)_{t \leqslant T}$ so that $\tau(x, \omega)=\left(x, \tau_{x}(\omega)\right)$. By Fubini's theorem,

$$
\begin{equation*}
\nu=\tau_{*}\left(\mu \otimes \gamma_{T}\right)=\int_{\mathbb{R}^{n}}\left(x, \tau_{x}\right)_{*} \gamma_{T} \mathrm{~d} \mu(x) \tag{4.6}
\end{equation*}
$$

This means that for any test function $g$,

$$
\int_{\Omega} g \mathrm{~d} \nu=\int_{\mathbb{R}^{n}}\left(\int_{\mathcal{V}_{T}} g(x, \widetilde{\theta}) \mathrm{d}\left(\left(\tau_{x}\right)_{*} \gamma_{T}\right)(\widetilde{\theta})\right) \mathrm{d} \mu(x)
$$

Since $\tau_{x}$ is just a translation in Wiener space by the deterministic function $f_{x}(t)=t x$, the Cameron-Martin theorem [11] yields that the density of $\left(\tau_{x}\right)_{*} \gamma_{T}$ with respect to $\gamma_{T}$ at the point $\left(\widetilde{\theta}_{t}\right)_{0 \leqslant t \leqslant T} \in \mathcal{V}_{T}$ equals

$$
\begin{align*}
\frac{\mathrm{d}\left(\tau_{x}\right)_{*} \gamma_{T}}{\mathrm{~d} \gamma_{T}}\left(\left(\widetilde{\theta}_{t}\right)_{t \leqslant T}\right) & =\exp \left(\int_{0}^{T}\left\langle f_{x}^{\prime}(t), \mathrm{d} \widetilde{\theta}_{t}\right\rangle-\frac{1}{2} \int_{0}^{T}\left|f_{x}^{\prime}(t)\right|^{2} \mathrm{~d} t\right) \\
& =\exp \left(\int_{0}^{T} x \mathrm{~d} \widetilde{\theta}_{t}-\frac{1}{2} \int_{0}^{T}|x|^{2} \mathrm{~d} t\right)=e^{\left\langle\tilde{\theta}_{T}, x\right\rangle-\frac{T|x|^{2}}{2}} \tag{4.7}
\end{align*}
$$

It follows from (4.6) and (4.7) that

$$
\begin{equation*}
\frac{\mathrm{d} \nu}{\mathrm{~d}\left(\lambda \otimes \gamma_{T}\right)}(x, \widetilde{\theta})=p_{0}(x) \cdot \frac{\mathrm{d} \nu}{\mathrm{~d}\left(\mu \otimes \gamma_{T}\right)}(x, \widetilde{\theta})=p_{0}(x) e^{\left\langle\tilde{\theta}_{T}, x\right\rangle-T \frac{|x|^{2}}{2}} \tag{4.8}
\end{equation*}
$$

The probability measure $\nu$ is the joint distribution of the stochastic process

$$
\left(X,\left(\widetilde{\theta}_{t}\right)_{0 \leqslant t \leqslant T}\right)
$$

described in (4.4). Therefore, when conditioning on the entire stochastic process $\left(\widetilde{\theta}_{t}\right)_{0 \leqslant t \leqslant T}$, it follows from (4.8) that the probability density function of $X$ is proportional to $x \mapsto p_{0}(x) e^{\left\langle\tilde{\theta}_{T}, x\right\rangle-T \frac{|x|^{2}}{2}}$ in $\mathbb{R}^{n}$. This completes the proof of (ii).

We move on to the proof of (i). We endow $\Omega$ with the probability measure $\mu \otimes \gamma_{T}$, and assume that $\left(X,\left(W_{t}\right)_{t \geqslant 0}\right)$ is distributed according to this measure, while $\widetilde{\theta}_{t}=t X+W_{t}$. Thus,

$$
\begin{equation*}
\mathrm{d} \widetilde{\theta}_{t}=X \mathrm{~d} t+\mathrm{d} W_{t} \tag{4.9}
\end{equation*}
$$

Write $\mathcal{N}_{t}$ for the $\sigma$-algebra generated by $\left(\widetilde{\theta}_{s}\right)_{0 \leqslant s \leqslant t}$. Abbreviate $\mathbb{E}[X \mid \widetilde{\theta}]=$ $\mathbb{E}\left[X \mid \mathcal{N}_{t}\right](\widetilde{\theta})$ for the conditional expectation of $X$ with respect to $\mathcal{N}_{t}$, which
is a function of $\left(\widetilde{\theta}_{s}\right)_{0 \leqslant s \leqslant t}$. According to (4.9) and [35, Theorem 8.4.3], the process $\left(\widetilde{\theta}_{t}\right)_{0 \leqslant t \leqslant T}$ coincides in law with the process $\left(\theta_{t}\right)_{0 \leqslant t \leqslant T}$ defined by the initial condition $\theta_{0}=\widetilde{\theta}_{0}=0$ and the stochastic differential equation

$$
\mathrm{d} \theta_{t}=b\left(t, \theta_{t}\right) \mathrm{d} t+\mathrm{d} W_{t}
$$

if the function $b(t, x)$ defined for $0<t \leqslant T$ and $x \in \mathbb{R}^{n}$ satisfies

$$
\begin{equation*}
b\left(t, \tilde{\theta}_{t}\right)=\mathbb{E}[X \mid \widetilde{\theta}] \quad \forall \tilde{\theta} \in \mathcal{V}_{t} \tag{4.10}
\end{equation*}
$$

(To be precise, the statement in [35, Theorem 8.4.3] only treats time-independent diffusions, but the proofs generalize almost verbatim to the timedependent case which we need.) The random variable $\mathbb{E}[X \mid \widetilde{\theta}]$, viewed as an $\mathcal{N}_{t}$-measurable function on $\Omega$, is the conditional expectation of $X$ given $\widetilde{\theta}=\left(\widetilde{\theta}_{s}\right)_{0 \leqslant s \leqslant t}$. According to (ii), the conditional distribution of $X$ given $\widetilde{\theta}$ is given by the probability density $q_{t}(x \mid \widetilde{\theta})$ from (4.5). Hence for any $0<t<T$ and $\widetilde{\theta} \in \mathcal{V}_{t}$,

$$
\mathbb{E}[X \mid \widetilde{\theta}]=\int_{\mathbb{R}^{n}} x \cdot q_{t}(x \mid \widetilde{\theta}) \mathrm{d} x=\int_{\mathbb{R}^{n}} x \cdot p_{t, \tilde{\theta}_{t}}(x) \mathrm{d} x=a\left(t, \widetilde{\theta}_{t}\right)
$$

We have thus verified condition (4.10) with $b(t, x)=a(t, x)$, completing the proof of (i).

To reiterate, we have interpreted Eldan's stochastic localization for the measure $\mu$ as the following procedure: a value $x$ is sampled from the distribution $\mu$, and a Brownian motion with constant drift $x$, namely $\theta_{t}=t x+W_{t}$, is shown to an observer who knows the distribution $\mu$, but not the value of $x$. From the observer's perspective, $\theta_{t}$ satisfies the stochastic differential equation (4.3), and at time $t$, the observer's posterior probability distribution for the hidden drift coefficient $x$ is precisely $p_{t, \theta_{t}}$. The fact that $p_{t, \theta_{t}}$ is a martingale now follows immediately from the law of total probability: for $s<t$,

$$
\begin{align*}
& \mathbb{E}\left[p_{t, \theta_{t}}(x) \mid \theta_{s}\right] \\
& \quad=\mathbb{E}\left[p\left(X=x \mid \theta_{t}\right) \mid \theta_{s}\right]=\int_{\mathbb{R}^{n}} p\left(X=x \mid \theta_{t}=\theta\right) p\left(\theta_{t}=\theta \mid \theta_{s}\right) \mathrm{d} \theta \\
& \quad=\int_{\mathbb{R}^{n}} p\left(X=x, \theta_{t}=\theta \mid \theta_{s}\right) \mathrm{d} \theta=p\left(X=x \mid \theta_{s}\right)=p_{s, \theta_{s}}(x) \tag{4.11}
\end{align*}
$$

In Bayesian terms, this simply means that if we continually obtain information about an unknown random variable $X$ and update our posteriors for $X$ accordingly, our expectation at time $s$ for our estimate of $X$, or any function of $X$, at time $t$ must coincide with our current estimate of $X$.

## Remark 4.2.

(i) A curious property of Eldan's stochastic localization, in the Bayesian interpretation, is that the posterior distribution at time $t$ depends only on $\theta_{t}$ : the full path $\left(\theta_{s}\right)_{0 \leqslant s \leqslant t}$ contains no more information about $X$ than $\theta_{t}$ alone. This is a limiting case of an amusing exercise in linear algebra and statistics which we now describe. Suppose that we are given $N$ noisy observations of an unknown quantity $X$, of the form

$$
\begin{equation*}
X+Z_{1}, X+\frac{Z_{1}+Z_{2}}{2}, \ldots, X+\frac{Z_{1}+\ldots+Z_{N}}{N} \tag{4.12}
\end{equation*}
$$

where $Z_{1}, \ldots, Z_{N}$ are independent, standard Gaussian random variables. Assume that the apriori distribution of $X$ is known to us. What is the posterior distribution of $X$ given the $N$ observations in (4.12)? As it turns out, the posterior distribution depends only on the last of these $N$ observations, for which the Gaussian noise is of the smallest variance. The first $N-1$ observations are completely useless in this context.
(ii) A suitably generalized version of this interpretation applies to the general stochastic localization process with a control matrix $C_{t}=$ $C\left(t, \theta_{t}\right)$, as defined by [31]: a random variable $X$ is drawn from $\mu$ as above, but instead of a Brownian motion with drift $X$, what the observer sees is an Itô process defined by the SDE $\mathrm{d} \theta_{t}=C\left(t, \theta_{t}\right) X \mathrm{~d} t+$ $C\left(t, \theta_{t}\right)^{\frac{1}{2}} \mathrm{~d} W_{t}$. Again, $p_{t}$ represents the observer's posterior distribution for $X$ given the observation of $\theta_{t}$ up to time $t$. To prove that this description corresponds to the definition of the process in [31] one repeats the above argument using Girsanov's theorem, rather than the Cameron-Martin theorem. The "path-independence" property of the posteriors from the previous remark does not hold in this case.

### 4.1. Time inversion

Let us now explain the relationship between the tilt process in its Bayesian interpretation and our work in Section 2. A well-known identity for Brownian motion is the time-inversion property: suppose that $\left(W_{t}\right)_{t \in[0, \infty)}$ is a standard Brownian motion in $\mathbb{R}^{n}$ with $W_{0}=0$. Define $\left(\widetilde{W}_{s}\right)_{s \in[0, \infty)}$ by

$$
\widetilde{W}_{s}=s W_{\frac{1}{s}}
$$

and $\widetilde{W}_{0}=0$. Then $\left(\widetilde{W}_{s}\right)_{s \in[0, \infty)}$ is a standard Brownian motion as well. Consequently, from (4.4) we see that

$$
\begin{equation*}
Y_{s}:=s \widetilde{\theta}_{1 / s}=X+\widetilde{W}_{s} \tag{4.13}
\end{equation*}
$$

Recalling that the tilt process $\left(\theta_{t}\right)_{t \geqslant 0}$ coincides in law with $\left(\widetilde{\theta}_{t}\right)_{t \geqslant 0}$, we conclude from (4.13) that the tilt process coincides in law with the time inversion of a Brownian motion with a starting point drawn from the distribution $\mu$.

Applying this time inversion, we treat the time-inverted tilt process $\left(Y_{s}\right)_{s \geqslant 0}$ as just a Brownian motion with a random starting point. Working with it requires nothing more than the explicit expression for the Euclidean heat kernel; for instance, the distribution of $Y_{s}$ is given by the probability density function $\rho_{s}=P_{s} \rho$ with $\rho=p_{0}$. Given a function $\varphi$ on $\mathbb{R}^{n}$ and $t>0$, the random variable

$$
\int_{\mathbb{R}^{n}} \varphi p_{t}
$$

associated to the tilt process coincides in law with the distribution of $Q_{s} \varphi$ under the measure $\mu_{s}$, for $s=1 / t$. Moreover,

$$
\begin{equation*}
Q_{s} \varphi(y)=\mathbb{E}\left[\varphi(X) \mid \theta_{t}=\theta\right] \quad \text { for } s=1 / t>0, \theta=t y \in \mathbb{R}^{n} \tag{4.14}
\end{equation*}
$$

It is this elementary, "functional analytic" perspective on the measures $p_{t, \theta}$ (or, in the new variables, $p_{s, y}$ ) that is taken in Section 2, which makes no explicit use of stochastic localization, pathwise analysis, martingales or stochastic calculus at all.

### 4.2. Föllmer drift as a "time-compressed" version of stochastic localization

Föllmer drift is a well-known stochastic process which couples between an absolutely continuous measure $\mu$ and Wiener measure on path space over a finite time interval, without loss of generality $[0,1]$. It is the same process referred to as the " $h$-process" in Cattiaux and Guillin [12], because of its relation to Doob's $h$-transform.

In brief, the Föllmer drift associated to $\mu$ is a Brownian motion conditioned to have law $\mu$ at time $t=1$. The measure $\mathcal{P}^{\mu}$ on $\mathcal{V}_{1}=C_{0}[0,1]$ defining the Föllmer drift of $\mu$ is defined as the measure having Radon-Nikodym derivative

$$
\begin{equation*}
\frac{\mathrm{d} \mathcal{P}^{\mu}}{\gamma_{1}}(W)=\frac{\mathrm{d} \mu}{\mathrm{~d} \gamma}\left(W_{1}\right) \quad\left(W=\left(W_{t}\right)_{0 \leqslant t \leqslant 1} \in \mathcal{V}_{1}\right) \tag{4.15}
\end{equation*}
$$

where $\gamma_{1}$ on the left side of (4.15) is the Wiener measure on $C_{0}[0,1]$, while $\gamma$ on the right side of (4.15) is the standard Gaussian measure in $\mathbb{R}^{n}$. The Föllmer drift $\mathcal{P}^{\mu}$ turns out to have a certain energy-minimizing property, and
its energy is precisely twice the relative entropy $H(\mu \mid \gamma)$, properties which make it quite useful for proving functional inequalities; see, e.g., [21, 32].

We can interpret Föllmer drift in a manner completely analogous to the tilt process: let $X$ be a random variable drawn from $\mu$ and let $\left(B_{t}\right)_{0 \leqslant t \leqslant 1}$ denote an independent standard Brownian bridge on $[0,1]$. Then the law of the process

$$
\begin{equation*}
X_{t}=t X+B_{t} \quad(0 \leqslant t \leqslant 1) \tag{4.16}
\end{equation*}
$$

is precisely that of the Föllmer drift associated to $\mu$. Moreover, just as above, one may consider an observer who sees $X_{t}$ but not $X$ and define posterior probability distributions for $X$ given $\left(X_{s}\right)_{s \leqslant t}$. These posterior probability distributions are the random measures $\mu_{t}$ in [21, §3], in a slightly different normalization (in [21] the measure $\mu_{t}$ is the posterior probability distribution of $\left(X-X_{t}\right) / \sqrt{1-t}$ given $X_{t}$, rather than the posterior probability distribution of $X$ itself given $X_{t}$ ).

In fact, there is an even closer relationship between Föllmer drift $X_{t}$ and the tilt process $\theta_{t}$ of Eldan's stochastic localization, which manifests in two separate ways. First of all, for $t \leqslant 1$, we may write

$$
\begin{equation*}
\theta_{t}=t X+W_{t}=t\left(X+W_{1}\right)+\left(W_{t}-t W_{1}\right) \tag{4.17}
\end{equation*}
$$

Note that $W_{1}$ is independent of $W_{t}-t W_{1}$ as these are jointly Gaussian, centered and $\mathbb{E} W_{1}\left(W_{t}-t W_{1}\right)=0$. Recall that $B_{t}=W_{t}-t W_{1}$ is one way to construct a Brownian bridge. From (4.16) and (4.17) we see that the tilt process $\left(\theta_{t}\right)_{t \in[0,1]}$ has the law of the Föllmer drift for the measure associated to $X+W_{1}$, namely $\mu * \gamma$. In the same fashion, one sees that for any $T>0$, the process $\left(\theta_{t}\right)_{t \in[0, T]}$ is identical in law to the Föllmer drift of $\mu * \gamma_{T}$, with the time interval rescaled to $[0, T]$.

Another way to construct a standard Brownian bridge from a standard Brownian motion is by "time compression": the process

$$
\widetilde{B}_{t}=(1-t) W_{\frac{t}{1-t}} \quad(0 \leqslant t \leqslant 1)
$$

coincides in law with the standard Brownian bridge. By inverting this operation, we can construct a Brownian motion from a Brownian bridge: $W_{t}=$ $(1+t) \widetilde{B}_{\frac{t}{1+t}}$. Hence, the Föllmer drift for $\mu$ and the tilt process of Eldan's stochastic localization satisfy the reciprocal relations

$$
X_{t} \simeq(1-t) \theta_{\frac{t}{1-t}} \quad \text { and } \quad \theta_{t} \simeq(1+t) X_{\frac{t}{1+t}}
$$

where $\simeq$ means "coincides in law". This follows from the defining formulas $\theta_{t}=t X+W_{t}, X_{t}=t X+B_{t}$ and the corresponding relations for $B_{t}$ and $W_{t}$. Thus, Föllmer drift is simply a time-compressed version of the tilt process.

## Note added in revision

El Alaoui and Montanari [1] have independently arrived at many of the results of Section 4, almost simultaneously with the appearance of our paper on arXiv. Furthermore, in a recent paper by Klartag and Lehec [25], the functional analytic framework suggested above, and its relation to Eldan's Stochastic localization, were instrumental in showing that the KLS conjecture holds true up to a factor that is polylogarithmic in the dimension.

## Appendix A. Discreteness of the spectrum for rapidly decreasing log-concave densities

The goal of this appendix is to prove the following proposition:
Proposition A.1. - Let $\mu$ be a log-concave probability measure on $\mathbb{R}^{n}$ with smooth, positive density $\rho=e^{-V}$ such that $\frac{V(x)}{|x|} \rightarrow \infty$ as $x \rightarrow \infty$. Then the spectrum of $L_{\mu}$ is discrete.

For $A \subseteq \mathbb{R}^{n}$ we write $C_{c}^{\infty}(A)$ for the class of smooth, compactly supported functions in $\mathbb{R}^{n}$ that are supported in the set $A$. As explained in [2, §4.10], in order to prove Proposition A it suffices to show the following:
$(*)$ For any $a>0$ there exists $r>0$ such that for any $f \in C_{c}^{\infty}\left(\mathbb{R}^{n} \backslash B_{r}\right)$,

$$
\int_{\mathbb{R}^{n}}|\nabla f|^{2} \mathrm{~d} \mu \geqslant a \cdot \int_{\mathbb{R}^{n}} f^{2} \mathrm{~d} \mu .
$$

Here $B_{r}=\left\{x \in \mathbb{R}^{n} ;|x| \leqslant r\right\}$.
Consider first the one-dimensional case in which $\mu$ is supported on a half-line. Thus $\mathrm{d} \mu=\rho(x) \mathrm{d} x=e^{-W(x)} \mathrm{d} x$ is a measure on $[0, \infty)$ with $W$ : $[0, \infty) \rightarrow \mathbb{R}$ smooth and convex. We will apply the Muckenhoupt criterion ([34]; see also [2, §4.5.1]), which we state as the following lemma:

Lemma A.2. - Let $\rho:[0, \infty) \rightarrow(0, \infty)$ be such that

$$
C:=\sup _{r>0} \int_{r}^{\infty} \rho \int_{0}^{r} \frac{1}{\rho}<\infty .
$$

Then for every smooth, compactly supported function $f:[0, \infty) \rightarrow \mathbb{R}$,

$$
\int_{0}^{\infty} f^{2} \rho \leqslant 4 C \int_{0}^{\infty}\left(f^{\prime}\right)^{2} \rho
$$

Let $x_{0}>0$ be such that $a:=W^{\prime}\left(x_{0}\right)>0$. Then $W^{\prime}(x) \geqslant a$ for all $x>x_{0}$ by convexity. Hence for any $r, x>x_{0}$,

$$
\begin{array}{ll}
\rho(r) \leqslant e^{-a(r-x)} \rho(x) & \text { for } x<r \\
\rho(x) \leqslant e^{-a(x-r)} \rho(r) & \text { for } x>r
\end{array}
$$

Therefore, for any $r>x_{0}$,

$$
\begin{align*}
& \int_{x_{0}}^{r} \frac{1}{\rho(x)} \mathrm{d} x \leqslant \frac{1}{\rho(r)} \int_{x_{0}}^{r} e^{-a(r-x)} \mathrm{d} x \leqslant \frac{1}{\rho(r)} \cdot \frac{1}{a}  \tag{A.1}\\
& \int_{r}^{\infty} \rho(x) \mathrm{d} x \leqslant \rho(r) \int_{r}^{\infty} e^{-a(x-r)} \mathrm{d} x \leqslant \rho(r) \cdot \frac{1}{a} \tag{A.2}
\end{align*}
$$

Thus we obtain $\sup _{r>x_{0}} \int_{r}^{\infty} \rho \int_{x_{0}}^{r} \frac{1}{\rho} \leqslant \frac{1}{a^{2}}$, so Muckenhoupt's criterion yields that for any $f \in C_{c}^{\infty}\left(\left[x_{0}, \infty\right)\right)$,

$$
\begin{equation*}
\int_{x_{0}}^{\infty}\left(f^{\prime}\right)^{2} \mathrm{~d} \mu \geqslant \frac{a^{2}}{4} \int_{x_{0}}^{\infty} f^{2} \mathrm{~d} \mu \tag{A.3}
\end{equation*}
$$

whenever $a=W^{\prime}\left(x_{0}\right)>0$.
So much for the one-dimensional case. Now let $\mathrm{d} \mu=e^{-V} \mathrm{~d} x$ be an $n$ dimensional log-concave measure, and consider the family of functions $f_{R}$ : $S^{n-1} \rightarrow \mathbb{R}$ defined by $f_{R}(u)=\frac{V(R u)-V(0)}{R}$. By convexity, $f_{R}$ is monotone increasing in $R$, and by assumption $f_{R}$ converges pointwise to infinity. Hence, applying Dini's theorem, we see that $f_{R}$ converges uniformly to $\infty$. Denote $V_{u}(r)=V(r u)$ for $u \in S^{n-1}$ and $r \geqslant 0$. By convexity,

$$
V_{u}^{\prime}(r) \geqslant f_{R}(u)
$$

Hence for every $a>0$ there exists $\widetilde{R}>0$ such that $V_{u}^{\prime}(r) \geqslant a$ for all $r \geqslant \widetilde{R}$ and $u \in S^{n-1}$. Denoting $W_{u}(r)=V_{u}(r)-(n-1) \log r$, we see we see that $W_{u}$ is convex in $(0, \infty)$ and that for any $a>0$ there exists $R=R(a)>0$ such that $W_{u}^{\prime}(r) \geqslant a$ for all $r \geqslant R$ and $u \in S^{n-1}$. By integrating in polar coordinates and using (A.3) we conclude that for any $a>0$ and $f \in C_{c}^{\infty}\left(\mathbb{R}^{n} \backslash B_{R}\right)$, denoting $f_{u}(r)=f(r u)$,

$$
\begin{aligned}
\|\nabla f\|_{L^{2}(\mu)}^{2} & \geqslant \int_{S^{n-1}} \int_{0}^{\infty} r^{n-1}\left(f_{u}^{\prime}(r)\right)^{2} e^{-V_{u}(r)} \mathrm{d} r \mathrm{~d} u \\
& \geqslant \int_{S^{n-1}}\left(\int_{R}^{\infty}\left(f_{u}^{\prime}(r)\right)^{2} e^{-W_{u}(r)} \mathrm{d} r\right) \mathrm{d} u \\
& \geqslant \frac{a^{2}}{4} \int_{S^{n-1}}\left(\int_{R}^{\infty} f_{u}(r)^{2} e^{-W_{u}(r)} \mathrm{d} r\right) \mathrm{d} u=\frac{a^{2}}{4}\|f\|_{L^{2}(\mu)}^{2}
\end{aligned}
$$

We have thus verified condition $(*)$ above, completing the proof of Proposition A.1.

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