# On the Generalised Langevin Equation for Simulated Annealing 

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#### Abstract

In this paper, we consider the generalised (higher order) Langevin equation for the purpose of simulated annealing and optimisation of nonconvex functions. Our approach modifies the underdamped Langevin equation by replacing the Brownian noise with an appropriate Ornstein-Uhlenbeck process to account for memory in the system. Under reasonable conditions on the loss function and the annealing schedule, we establish convergence of the continuous time dynamics to a global minimum. In addition, we investigate the performance numerically and show better performance and higher exploration of the state space compared to the underdamped and overdamped Langevin dynamics with the same annealing schedule.


## 1 Introduction

Algorithms for optimisation have received significant interest in recent years due to applications in machine learning, data science and molecular dynamics. Models in machine learning are formulated to have some loss function and parameters with respect to which it is to be minimised, where use of optimisation techniques is heavily relied upon. We refer to $[7,62]$ for related discussions. Many models, for instance neural networks, use parameters that vary over a continuous space, where gradient-based optimisation methods can be used to find good parameters that generate effective predictive ability. As such, the design and analysis of such algorithms for global optimisation has been the subject of considerable research [60] and it has proved useful to study algorithms for global optimisation using tools from the theory of stochastic processes and dynamical systems. A paradigm of the use of stochastic dynamics for the design of algorithms for global optimisation is one of simulated annealing, where overdamped Langevin dynamics with a time dependent temperature (1.1) that decreases with an appropriate cooling schedule is used to guarantee the global minimum of a nonconvex loss function $U: \mathbb{R}^{n} \rightarrow \mathbb{R}$ :

$$
\begin{equation*}
d X_{t}=-\nabla U\left(X_{t}\right) d t+\sqrt{2 T_{t}} d W_{t} \tag{1.1}
\end{equation*}
$$

Here $W_{t}$ is a standard $n$-dimensional Wiener process and $T_{t}: \mathbb{R}_{+} \rightarrow \mathbb{R}$ is an appropriate decreasing determinstic function of time often referred to as the annealing or cooling schedule. For fixed $T_{t}=T>0$, this is the dynamics used for the related problem of sampling from a possibly high dimensional probability measure, for example in the unadjusted Langevin algorithm [18]. Gradually decreasing $T_{t}$ to zero balances the exploration-exploitation trade-off by allowing at early times larger noise to drive $X_{t}$ and hence sufficient mixing to escape local minima. Designing an appropriate annealing schedule is well-understood. We briefly mention classical references $[15,24,25,27,28,31$, $32,35]$, as well as the more recent [34, 41, 57], where one can find details and convergence results. In this paper we aim to consider generalised versions of (1.1) for the same purpose.

Using dynamics such as (1.1) has clear connections with sampling. When $T_{t}=T$ is a constant function, the invariant distribution of $X$ is proportional to $\exp \left(-\frac{U(x)}{T}\right) d x$. In addition, when $T_{t}$ decreases with time, the probability measure $\nu_{t}(d x) \propto \exp \left(-\frac{U(x)}{T_{t}}\right) d x$ converges weakly to the set of global minima based on the Laplace principle [33]. One can expect that if one replaces (1.1) with a stochastic process that mixes faster and maintains the same invariant distribution for constant temperatures, then the superior speed of convergence should improve
performance in optimisation due to the increased exploration of the state space. Indeed, it is well known that many different dynamics can be used in order to sample from a given probability distribution, or for finding the minima of a function when the dynamics is combined with an appropriate cooling schedule for the temperature. Different kinds of dynamics have already been considered for sampling, e.g. nonreversible dynamics, preconditioned unadjusted Langevin dynamics [2, 4, 39, 54], as well as for optimisation, e.g. interacting Langevin dynamics [65], consensus based optimisation $[8,9,58]$, etc.

A natural candidate in this direction is to use the underdamped or kinetic Langevin dynamics:

$$
\begin{align*}
d X_{t} & =Y_{t} d t  \tag{1.2a}\\
d Y_{t} & =-\nabla U\left(X_{t}\right) d t-T_{t}^{-1} \mu Y_{t} d t+\sqrt{2 \mu} d W_{t} \tag{1.2b}
\end{align*}
$$

Here the reversibility property of (1.1) has been lost; the improvement from breaking reversibility in both the context of sampling and that of optimisation ${ }^{1}$ is investigated in $[16,38]$ and $[21]$ respectively. When $T_{t}=T$, (1.2) can converge faster than (1.1) to its invariant distribution

$$
\rho(d x, d y) \propto \exp \left(-\frac{1}{T}\left(U(x)+\frac{|y|^{2}}{2}\right)\right) d x d y
$$

see [19] or Section 6.3 of [56] for particular comparisons and also [5, 6] for more applications using variants of (1.2). In the context of simulated annealing, using this set of dynamics has recently been studied rigorously in [47], where the author established convergence to global minima using the generalised $\Gamma$-calculus [48] framework that is based on Bakry-Emery theory. Note that (1.2b) uses the temperature in the drift rather than the diffusion constant in the noise as in (1.1). Both formulations admit the same invariant measure when $T_{t}=T$. In the remainder of the paper, we adopt this formulation to be closer to [47].

In this paper we will consider an extension of the kinetic Langevin equation by adding an additional auxiliary variable that accounts for the memory in the system. To the best of the authors' knowledge, this has not been attempted before in the context of simulated annealing and global optimisation. In particular we consider the Markovian approximation to the generalised Langevin equation:

$$
\begin{align*}
d X_{t} & =Y_{t} d t  \tag{1.3a}\\
d Y_{t} & =-\nabla U\left(X_{t}\right) d t+\lambda^{\top} Z_{t} d t  \tag{1.3b}\\
d Z_{t} & =-\lambda Y_{t} d t-T_{t}^{-1} A Z_{t} d t+\Sigma d W_{t} \tag{1.3c}
\end{align*}
$$

with $A \in \mathbb{R}^{m \times m}$ being positive definite and $\Sigma \in \mathbb{R}^{m \times m}$ restricted to satisfying

$$
\Sigma \Sigma^{\top}=A+A^{\top}
$$

Here $X_{t}, Y_{t} \in \mathbb{R}^{n}$ and $Z_{t} \in \mathbb{R}^{m}$ (with $m \geq n$ ), $M^{\top}$ denotes the transpose of a matrix $M, \lambda \in \mathbb{R}^{m \times n}$ is a rank $n$ matrix with a left inverse $\lambda^{-1} \in \mathbb{R}^{n \times m}$.

Our aim is to establish convergence using similar techniques as [47] and investigate the improvements in performance. Equation (1.3) is related to the generalised Langevin equation, where memory is added to (1.2) by an integrating over past velocities with a kernel $\Gamma: \mathbb{R}_{+} \rightarrow \mathbb{R}^{n \times n}$ :

$$
\begin{equation*}
\ddot{x}=-\nabla U(x)-\int_{0}^{t} \Gamma(t-s) \dot{x}(s) d s+F_{t} \tag{1.4}
\end{equation*}
$$

with $F_{t}$ being a zero mean stationary Gaussian process with an autocorrelation matrix given by the fluctuationdissipation theorem

$$
\mathbb{E}\left(F_{t} F_{s}^{\top}\right)=T_{t} \Gamma(t-s)
$$

When $T_{t}=T$, (1.4) is equivalent to (1.3) when setting

$$
\begin{equation*}
\Gamma(t)=\lambda^{\top} e^{-A t} \lambda, \tag{1.5}
\end{equation*}
$$

[^0]and the invariant distribution becomes
$$
\rho(d x, d y, d z) \propto \exp \left(-\frac{1}{T}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)\right) d x d y d z
$$
see Section 8.2 of [56] for details ${ }^{2}$. In the spirit of adding a momentum variable in (1.1) to get (1.2), (1.3) adds an additional auxiliary variable to the Langevin system whilst preserving the invariant distribution in the $x$ marginal. In the constant temperature context, (1.4) is natural from the point of statistical mechanics and has already been considered as a sampling tool in $[10,11,12,50]$ with considerable success. We will demonstrate numerically that the additional tuning parameters can improve performance; see also [49] for recent work demonstrating advantages of using (1.4) compared to using (1.2) when sampling from a log concave density. A detailed study of the Markovian approximation (1.3) of the generalised Langevin dynamics (1.4) can be found in [52].

To motivate the use of (1.3), consider the quadratic case where $U=\alpha x^{2}$ and $0<\alpha<1$. This case allows for explicit or numerical calculation of the spectral gaps of the generators in (1.1)-(1.3) in order to compare the rate of convergence to equilibrium; see $[53,44]$ for details. For a given $T$, it is possible to choose $\lambda$ and $A$, such that the spectral gap of the generator of (1.3) is much larger than that of (1.2) with the best choice of $\mu$ being used. The latter is already larger than that of the overdamped dynamics in (1.1). We will later demonstrate numerically that this will translate to better exploration in simulated annealing (when $T_{t}$ is decreasing in time).

Use of (1.4) is also motivated by parallels with accelerated gradient descent algorithms. When the noise is removed from (1.2), the second order differential equation can be loosely considered as a continuous time version of Nesterov's algorithm [64]. The latter is commonly preferred to discretising the first order differential equation given by the noiseless version of (1.1), because in the high dimensional and low iterations setting it achieves the optimal rate of convergence for convex optimisation; see Chapter 2 in [51] and also [26] for a nonconvex setting. Here we would like to investigate the effect of adding another auxiliary variable, which would correspond to a third order differential equation when noise is removed. When noise is added for the fixed temperature case, [20] has studied the long time behaviour and stability for different choices of a memory kernel as in (1.4). Finally, we note that generalised Langevin dynamics in (1.4) have additionally been studied in related areas such as sampling problems in molecular dynamics from chemical modelling [1, 10, 11, 12, 50, 68], see also [36] for work determining the kernel $\Gamma$ in the generalised system (1.4) from data.

Our theoretical results will focus only on the continuous time dynamics and follow closely the approach in [47]. The main requirement in terms of assumptions are quadratic upper and lower bounds on $U$ and bounded second derivatives. This is different to classical references such as [25], [27] or [32]. These works also rely on the Poincaré inequality, an approach which will be mirrored here (and in [47] for the underdamped case) using a log-Sobolev inequality; see also [31] for the relationship between such functional inequalities and the annealing schedule in the finite state space case. We will also present detailed numerical results for different choices of $U$. There are many possibilities for the method of discretisation of (1.3), we will use a time discretisation scheme that appeared in [3], but will not present theoretical results on the time discretised dynamics; this is beyond the scope of this article. We refer instead the interested reader to [61] for a study on discretisation schemes for the system (1.3), [14] for recent results on (1.2) and its time-discretisation and [22,23] for linking discrete time Markov chains with the overdamped Langevin system in (1.1).

### 1.1 Contributions and organisation of the paper

- To the best of the authors' knowledge, neither of the generalised Langevin systems (1.3) and (1.4) have been considered along with simulated annealing to solve a global optimisation problem. The main theoretical contribution consists of Theorem 2.4 that establishes convergence in probability of $X_{t}$ in the higher order Markovian dynamics (1.3) to a global minimiser of $U$. For the optimal cooling schedule $T_{t}$, the rate of convergence is as the known rate for the Langevin system (1.2) presented in [47].
- On a more technical level, the assumptions and proofs here closely parallel those of [47] bar a number of differences. Due to the different dynamics, we use a different form of the distorted entropy, stated formally in

[^1](C.8). In addition, we use different truncation arguments for establishing dissipation of this distorted entropy. We provide more details on differences in Remarks 2.1 and C.4. Also we make an effort to emphasise the role of the critical factor of the cooling schedule in the rate of convergence in Theorem 2.4. This can be seen in our assumptions for $T_{t}$ and $U$ below.

- Our results cover also convergence to equilibrium for the constant temperature case, which is relevant for sampling problems. In Proposition 2.5 we establish exponential convergence to equilibrium for (1.3) with $T_{t}=T>0$, see also Remark 2.4. This is not surprising, see [52] for similar results.
- Detailed numerical experiments are provided to illustrate the performance of our approach. We also discuss thoroughly tuning issues. In particular, we investigate the role of matrix $A$ and how it can be chosen to increase exploration of the state space. For the leapfrog time discretisation of [3], our results suggest that exploration of the state space is increased considerably compared to using an Euler discretisation of (1.2).

The paper is organised as follows. Section 2 will present the assumptions and main theoretical results. Detailed proofs can be found in the appendices. Section 3 presents numerical results demonstrating the effectiveness of our approach both in terms of reaching the global minimum and the exploration of the state space. In Section 4, we provide some concluding remarks.

## 2 Main Result

Let $L_{t}$ denote the infinitesimal generator of the associated semigroup to (1.3) at $t>0$ and temperature $T_{t}$. This is given by

$$
\begin{equation*}
L_{t}=\left(y \cdot \nabla_{x}-\nabla_{x} U(x) \cdot \nabla_{y}\right)+\left(z^{\top} \lambda \nabla_{y}-y^{\top} \lambda^{\top} \nabla_{z}\right)-T_{t}^{-1} z^{\top} A \nabla_{z}+A: D_{z}^{2} \tag{2.1}
\end{equation*}
$$

where we denote the gradient vector as $\nabla_{x}=\left(\partial_{x_{1}}, \ldots, \partial_{x_{n}}\right)^{\top}$, the Hessian with $D_{x}^{2}$ and similarly for the $y$ and $z$ variables. For matrices $M, N \in \mathbb{R}^{r \times r}$ we denote $M: N=\sum_{i, j} M_{i j} N_{i j}$ for all $1 \leq i, j \leq r$ and the operator norm $|M|=\sup \left\{\frac{|M v|}{|v|}: v \in \mathbb{R}^{r}\right.$ with $\left.x \neq 0\right\}$. We will also use $|v|$ to denote Euclidean distance for a vector $v$.

Let $m_{t}$ be the law of $\left(X_{t}, Y_{t}, Z_{t}\right)$ in (1.3) and, with slight abuse of notation, we will also denote as $m_{t}$ the corresponding Lebesgue density. Similarly we define $\mu_{T_{t}}$ be the instantaneous invariant law of the process

$$
\begin{equation*}
\mu_{T_{t}}(d x, d y, d z)=\frac{1}{Z_{T_{t}}} \exp \left(-\frac{1}{T_{t}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)\right) d x d y d z \tag{2.2}
\end{equation*}
$$

with $Z_{T_{t}}=\int \exp \left(-\frac{1}{T_{t}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)\right) d x d y d z$. Finally, define the density between the two laws:

$$
\begin{equation*}
h_{t}:=\frac{d m_{t}}{d \mu_{T_{t}}} \tag{2.3}
\end{equation*}
$$

We proceed by stating our assumptions.

## Assumption 1.

1. The potential $U$ is smooth with bounded second derivatives

$$
\begin{equation*}
\left|D_{x}^{2} U\right|_{\infty}:=\sup _{x \in \mathbb{R}^{n}}\left|D_{x}^{2} U(x)\right|<\infty \tag{2.4}
\end{equation*}
$$

and satisfies

$$
\begin{gather*}
|\bar{a} \circ x|^{2}+U_{m} \leq U(x) \leq|\bar{a} \circ x|^{2}+U_{M}  \tag{2.5}\\
\nabla_{x} U(x) \cdot x \geq r_{1}|x|^{2}-U_{g}  \tag{2.6}\\
\left|\nabla_{x} U(x)\right|^{2} \leq r_{2}|x|^{2}+U_{g} \tag{2.7}
\end{gather*}
$$

for some constants $\bar{a} \in \mathbb{R}_{+}^{n}, r_{1}, r_{2}, U_{g}>0$, and $U_{m}, U_{M} \in \mathbb{R}$, where o denotes the Hadamard product. In the rest of the paper, the smallest element of $\bar{a}$ is denoted as

$$
a_{m}:=\min _{i} \bar{a}_{i},
$$

where $\bar{a}=\left(\bar{a}_{1}, \ldots, \bar{a}_{n}\right)$.
2. The temperature $T_{t}: \mathbb{R}_{+} \rightarrow \mathbb{R}_{+}$is continuously differentiable and there exists some constant $t_{0}>0$ such that $T_{t}$ satisfies for all $t>t_{0}$ :
(i) $\lim _{t \rightarrow \infty} T_{t}=0$,
(ii) $T_{t} \geq E(\ln t)^{-1}$ for some constant $E>U_{M}-U_{m} \geq 0$ and $\widetilde{T}>0$,
(iii) $-\widetilde{T} t^{-1} \leq T_{t}^{\prime} \leq 0$ for some constant $\widetilde{T}>0$. In addition, there exists a small $\delta>0$ such that $T_{t}$ is constant on $[0, \delta]$.
3. The initial density $m_{0}$ satisfies:
(i) $m_{0}$ is smooth,
(ii) $\int \frac{\left|\nabla m_{0}\right|^{2}}{m_{0}} d x d y d z<\infty$,
(iii) $\int\left(|x|^{p}+|y|^{p}+|z|^{p}\right) m_{0} d x d y d z<0$ for $2 \leq p \leq \bar{p}, p \in \mathbb{N}$ and some $\bar{p} \in \mathbb{N}$.

Remark 2.1. Note that (2.5) and (2.7) deviate from [47]. The modification is useful for providing a clear characterisation of the annealing schedule (2. above) and the log-Sobolev constant in (C.17) found in the appendices. The relationship (C.17) between the log-Sobolev constant here and the critical value $U_{M}-U_{m}$ for $E$ mirrors that between the spectral gap of the overdamped Langevin generator (of (1.1)) and the same critical value appearing in the annealing schedule in (1.1) as shown in [34] and [59]. For the overdamped case, more recent extenstions such as the Eyring-Kramers formula for the spectral gap and the log-Sobolev constant can be found in [43]. Future work could consider extension of these ideas for the underdamped and generalised Langevin case.

We present two key propositions.
Proposition 2.1. For all $t>0$, denote by $\left(X^{T_{t}}, Y^{T_{t}}, Z^{T_{t}}\right)$ a r.v. with distribution $\mu_{T_{t}}$. For any $\delta, \alpha>0$, there exists a constant $\hat{A}>0$ such that

$$
\mathbb{P}\left(U\left(X^{T_{t}}\right)>\min U+\delta\right) \leq \hat{A} e^{-\frac{\delta-\alpha}{T_{t}}}
$$

Proposition 2.2. Under Assumption 1, for all $t>0, X_{t}, Y_{t}, Z_{t}$ are well defined, $\mathbb{E}\left[\left|X_{t}\right|^{2}+\left|Y_{t}\right|^{2}+\left|Z_{t}\right|^{2}\right]<\infty$ and $m_{t} \in \mathcal{C}_{+}^{\infty}=\left\{m \in \mathcal{C}^{\infty}: m>0\right\}$.

Propositions 2.1 and 2.2 follow by modifications on the proofs of Lemma 3, Proposition 4 and Proposition 5 of [47]. The statements are restated as Propositions B. 1 and B. 2 along with proofs in the Appendices. The proof of Proposition 2.1 uses probabilistic arguments and the one for Proposition 2.2 Hörmander's condition and controllability.
Remark 2.2. Proposition 2.1 can be thought of as a Laplace principle; Proposition 2.2 says roughly that the process (1.3) does not blow up in finite time and and the noise in the dynamics (1.3c) for $Z_{t}$ spreads throughout the system, that is to $X_{t}$ and $Y_{t}$.

Proposition 2.3. Under Assumption 1, for any $\alpha>0$, there exists some constant $B>0$ and $t_{h}>0$, both depending on $\left|D_{x}^{2} U\right|_{\infty}, A, \lambda, T_{t}, \alpha, U_{M}, U_{m}, U_{g}, E, r_{2}$ and $a_{m}$, such that for all $t>t_{h}$,

$$
\begin{equation*}
\int h_{t} \ln h_{t} d \mu_{T_{t}} \leq B\left(\frac{1}{t}\right)^{1-\frac{U_{M}-U_{m}}{E}-2 \alpha} \tag{2.8}
\end{equation*}
$$

The full proof is the contained in the appendices and follows directly from Proposition C.8. Therein a similar statement is proved for the distorted entropy that has the following form:

$$
H(t):=\int\left(\frac{\left\langle S \nabla h_{t}, \nabla h_{t}\right\rangle}{h_{t}}+\beta\left(T_{t}^{-1}\right) h_{t} \ln \left(h_{t}\right)\right) d \mu_{T_{t}}
$$

where $S$ being a well chosen matrix (so that (C.14) holds) and $\beta(\cdot)$ is a particular polynomial (see (C.8) for the precise form of $H(t)$ and (C.11) for $\beta(\cdot))$. This construction of $H$ compared to a standard definition of entropy compensates for the fact that the diffusion is degenerate (see [67] for a general discussion). The proof requires use of time derivatives of $H$, which is rigorously established using a truncation argument, whereby a sequence of compact functions with specific properties are multiplied onto the integrand. Then the problem is split into the partial time and partial temperature derivatives where, amongst other tools, (C.14) and a log-Sobolev inequality are used as in [47] to arrive at a bound that allows a Grönwall-type argument.
Remark 2.3. Proposition C. 8 is a statement about the distorted entropy $H(t)$, which bounds the entropy $\int h_{t} \ln h_{t} d \mu_{T_{t}}$. In fact this is achieved in such a way that the bound becomes less sharp as $t$ becomes large but without consequences for Theorem 2.4.

Remark 2.4. Part of the analysis used in the proof of Proposition C. 8 can be used for the sampling case and $T_{t}=T$, i.e. working only with the partial time derivatives mentioned above for the invariant distribution. Proposition 2.5 below shows exponential convergence to equilibrium for the generalised Langevin equation (1.3) with constant temperature.

We proceed with the statement of our main result.
Theorem 2.4. Under Assumption 1, for any $\delta>0$, as $t \rightarrow \infty$,

$$
\mathbb{P}\left(U\left(X_{t}\right) \leq \min U+\delta\right) \rightarrow 1
$$

If in addition $T_{t}=E(\ln t)^{-1}$, then for any $\alpha>0$, there exists a constant $C>0$ such that for all $t>0$,

$$
\mathbb{P}\left(U\left(X_{t}\right) \leq \min U+\delta\right) \leq C\left(\frac{1}{t}\right)^{r(E)}
$$

where the exponential rate $r:\left(U_{M}-U_{m}, \infty\right) \rightarrow \mathbb{R}$ is defined by

$$
\begin{aligned}
r(E) & :=\min \left(\frac{1-\frac{U_{M}-U_{m}}{E}-\alpha}{2}, \frac{\delta-\alpha}{E}\right) \\
& = \begin{cases}\frac{1}{2}\left(1-\frac{U_{M}-U_{m}}{E}-\alpha\right) & \text { if } E<\frac{U_{M}-U_{m}+2(\delta-\alpha)}{1-\alpha} \\
\frac{\delta-\alpha}{E} & \text { otherwise. }\end{cases}
\end{aligned}
$$

Proof. For all $t>0$, denote by $\left(X^{T_{t}}, Y^{T_{t}}, Z^{T_{t}}\right)$ a random variable with distribution $\mu_{T_{t}}$. For all $\delta>0$, with the definition (2.3) of $h_{t}$ and triangle inequality,

$$
\mathbb{P}\left(U\left(X_{t}\right)>\min U+\delta\right) \leq \mathbb{P}\left(U\left(X^{T_{t}}\right)>\min U+\delta\right)+\int\left|h_{t}-1\right| d \mu_{T_{t}}
$$

Pinsker's inequality gives

$$
\begin{equation*}
\int\left|h_{t}-1\right| d \mu_{T_{t}} \leq\left(2 \int h_{t} \ln h_{t} d \mu_{T_{t}}\right)^{\frac{1}{2}} \tag{2.9}
\end{equation*}
$$

which, by Proposition 2.3, together with Proposition 2.1 gives the result.
Remark 2.5. The cooling schedule $T_{t}=E(\ln t)^{-1}$ is optimal with respect to the method of proof for Proposition C.8; see Proposition D. 2 in the appendices. This is a consistent with most works in simulated annealing, e.g. $[15,24,25,27,28,31,32,35]$.

Remark 2.6. The 'mountain-like' shape of $r$ indicates the bottleneck for the rate of convergence at low and high values of $E$ : a small $E$ means convergence to the instantaneous equilibrium $\mu_{T_{t}}$ is slow and a large $E$ means the convergence of $\mu_{T_{t}}$ to the global minima of $U$ is slow.

Proposition 2.5. Let 1. and 3. of Assumption 1 hold and let $T_{t}=T$ for all $t$ for some constant $T>0$. It holds that

$$
\int\left|h_{t}-1\right| d \mu_{T} \leq \sqrt{\frac{2 H(0)}{\beta(T)}} e^{-\frac{C_{*}^{-1}}{2} t}
$$

where $C_{*}$ is the log-Sobolev constant (C.17)

$$
\begin{equation*}
C_{*}=A_{*}+\beta\left(T^{-1}\right) e^{\left(U_{M}-U_{m}\right) T^{-1}} \frac{T}{4} \max \left(2, a_{m}^{-2}\right) \tag{2.10}
\end{equation*}
$$

for all $t>0$, $H$ is the distorted entropy defined in (C.8), $A_{*}=A_{*}\left(\left|D_{x}^{2} U\right|_{\infty}, \lambda\right)>0$ is a constant and $\beta$ is a second order polynomial (C.11) depending on $\left|D_{x}^{2} U\right|_{\infty}, A$ and $\lambda$.

Proof. See proof of Proposition D. 1 in the Appendices.

## 3 Numerical results

Here we investigate the numerical performance of (1.3) in terms of convergence to a global optimum and exploration capabilities and compare with (1.2). In Section 3.1, we will present the discretisation we use for both sets of dynamics and some details related to the annealing schedule and parameters. In Section 3.2 and 3.3, for different parameters and cost functions, we present results for the probability of convergence to the global minimum and transition rates between different regions of the state space. We will investigate thoroughly the effect of $E$ appearing in the annealing schedule as well as the parameters in the dynamics (1.2) and (1.3).

### 3.1 Time discretisation

In order to simulate from (1.3), we will use the following time discretisation:

$$
\begin{align*}
Y_{n+\frac{1}{2}} & =Y_{n}-\frac{\Delta t}{2} \gamma \nabla U\left(X_{n}\right)+\frac{\Delta t}{2} \lambda^{\top} Z_{n}  \tag{3.1a}\\
X_{n+1} & =X_{n}+\Delta t \gamma Y_{n+\frac{1}{2}}  \tag{3.1b}\\
Z_{n+1} & =Z_{n}-\theta \lambda Y_{n+\frac{1}{2}}-\theta A Z_{n}+\alpha \sqrt{T_{k}} \Sigma \xi_{n}  \tag{3.1c}\\
Y_{n+1} & =Y_{n+\frac{1}{2}}-\frac{\Delta t}{2} \gamma \nabla U\left(X_{n+1}\right)+\frac{\Delta t}{2} \lambda^{\top} Z_{n+1} \tag{3.1d}
\end{align*}
$$

where $\Delta t$ denotes the time incremements in the discretisation, $\xi_{n}$ are i.i.d. standard $m$ dimensional normal random variables with unit variance and $\theta=1-\exp (-\Delta t)$, and $\alpha=\sqrt{1-\theta^{2}}$. Specifically this is method 2 of [3] applied on a slight modification of (1.3), where $\gamma Y_{t} d t$ and $\gamma \nabla U d t$ is used instead in the r.h.s. of (1.3a) and (1.3b). Tuning $\gamma$ can improve numerical perfomance especially in high dimensional problems, but we note that this has no effect in terms of the instantaneous invariant density in (2.2); $\gamma$ will not appear in (2.2) similar to $\lambda$ and $A$. Unless stated otherwise, in the remainder we will use $\gamma=1$.

As we will see below the choices for $A$ is important. To illustrate this we will use different choices of the form $A=\mu A_{i} ; i$ here is an index for different designs of $A$. The first choice will be to set $m=n$ and set $A_{1}=I_{n}$ where $I_{n}$ is $n \times n$ identity matrix. For the rest, we will use $m=2 n$ and

$$
A_{2}=\left(\begin{array}{cc}
0 & -I_{n} \\
I_{n} & I_{n}
\end{array}\right), \quad A_{3}=\left(\begin{array}{cc}
I_{n} & -I_{n} \\
I_{n} & I_{n}
\end{array}\right), \quad A_{4}=\left(\begin{array}{ccc}
1 & \cdots & \\
-1 & \ddots & \\
\vdots & \ddots & \\
-1 & \cdots & -1
\end{array}\right)
$$

Similarly we will use in each case $\lambda=\bar{\lambda} \lambda_{i}$ with $\lambda_{1}=I_{n}$ and

$$
\lambda_{i}=\binom{I_{n}}{0}
$$

for $i=2,3,4$. As a result $\bar{\lambda}, \mu>0$ are the main tuning constants for (3.1) that do not involve the annealing schedule.

The Langevin system (1.2) will be approximated with the following Euler-Maruyama scheme,

$$
\begin{align*}
X_{k+1} & =X_{k}+\Delta t \gamma Y_{k}  \tag{3.2a}\\
Y_{k+1} & =Y_{k}-\Delta t \gamma \nabla U\left(X_{k}\right)-\Delta t \mu Y_{k}+\sqrt{\Delta t \mu T_{k}} \xi_{k} \tag{3.2b}
\end{align*}
$$

To make valid comparisons, both (3.1) and (3.2) will use the same noise realisation $\xi_{k}$ (or the first common $n$ elements) and same step size $\Delta t$. Similarly, $\gamma=1$ unless stated otherwise. Finally for both cases we will use following annealing schedule:

$$
T_{k}=\left(\frac{1}{5}+\frac{\ln (1+k \Delta t)}{E}\right)^{-1}
$$

where $E$ is an additional tuning parameter.

### 3.2 Sample path properties

Our first set of simulations focus on illustrating some properties of the sample paths generated by (3.1) and (3.2). We will use the following bivariate potential function

$$
\begin{gather*}
U\left(x_{1}, x_{2}\right)=\frac{x_{1}^{2}}{5}+\frac{x_{2}^{2}}{10}+5 e^{-x_{1}^{2}}-7 e^{-\left(x_{1}+5\right)^{2}-\left(x_{2}-3\right)^{2}}-6 e^{-\left(x_{1}-5\right)^{2}-\left(x_{2}+2\right)^{2}} \\
 \tag{3.3}\\
+\frac{5 x_{1}^{2} e^{-\frac{x_{1}^{2}}{9}} \cos \left(x_{1}+2 x_{2}\right) \cos \left(2 x_{1}-x_{2}\right)}{1+\frac{x_{2}^{2}}{9}}
\end{gather*}
$$

The global minimum is located at $(-5,3)$, but there are plenty of local minima where the process can get trapped. In addition, there is a barrier along the vertical line $\left\{x_{1}=0\right\}$ that makes crossing from each half plane less likely. Here we set $\Delta t=0.1, E=5$ and each sample is initialised at $(4,2)$. As a result, it is harder to cross $\left\{x_{1}=0\right\}$ to reach the global minimum and it is quite common to get stuck in other local minima such as near $(5,-2)$.

To illustrate this, in Figure 3.1 we present contour plots of $U$ together with typical realisation of sample paths (in the left panels) for (3.2) and (3.1) for the different choices of $A_{i}$. As expected, (3.1) generates smoother paths than those of (3.2). We also employ independent runs of each stochastic process for the same initialisation. The results are presented in the right panels of Figure 3.1, where we show heat maps for two dimensional histograms representing the frequency of visiting each $\left(x_{1}, x_{2}\right)$ location over 15 independent realisations of each process. The heat maps in Figure 3.1 do not directly depict time dependence in the paths and only illustrate which areas are visited more frequently. Of course converging at the global minumum or the local one at $(5,-2)$ will result in more visits at these areas. The aim here is to investigate the exploration of the state space. A careful examination of the plots shows more visits for (3.1) near $\left\{x_{1}=0\right\}$. The increased number of crossings of the vertical line $\left\{x_{1}=0\right\}$ are also demonstrated in Table 3.1 for more independent runs.

### 3.3 Performance and tuning

As expected, the tuning parameters, $E, \bar{\lambda}$ and $\mu$ play significant roles in the performance of (3.1) and (3.2). As $E$ is common to both, we wish to demontrate that the additional tuning variable for (3.1) will can improve performance.

We first comment on relative scaling of $\bar{\lambda}$ and $\mu$ based on earlier work for quadratic $U$ and $T_{t}=T$ being constant. A quadratic $U$ satisfies the bounds in Assumption 1 and is of particular interest because analytical calculations are possible for the spectral gap of $L_{t}$, which in turn gives the (exponential) rate of convergence to the equilibrium

| Method equation | Number of transitions across $x=0$ |
| :---: | :---: |
| $(3.2)$ | 20609 |
| $(3.1)$ with $A=A_{1}$ | 21532 |
| $(3.1)$ with $A=A_{2}$ | 38804 |
| $(3.1)$ with $A=A_{3}$ | 32745 |
| $(3.1)$ with $A=A_{4}$ | 38948 |

Table 3.1: Number of crossings across the vertical line $\left\{x_{1}=0\right\}$ for $U$ defined in (3.3). The results are for $k=10^{5}$ iterations over $10^{4}$ independent runs.
distribution. It is observed numerically in [53] that in this case, (1.3) has a spectral gap that is approximately a function of $\frac{\bar{\lambda}^{2}}{\mu}$. On the other hand, the spectral gap of (1.2) with quadratic $U$ is a function of $\mu$ thanks to Theorem 3.1 in [44]. For the rest of the comparison, we will use $\frac{\bar{\lambda}^{2}}{\mu}$ and $\mu$ as variables for (3.1) and (3.2) respectively as these quantities appear to have a distinct effect on the mixing in each case.

We will consider three different cost functions $U$ and set $\Delta t=0.02$. As before we will initialise at a point well separated from the global minimum and consider each method to be successful if it convergences at a particular tolerance region near the global minumum. The details are presented in Table 3.2. We choose the popular Alpine function in 12 dimensions ( $\nabla U_{1}$ here is a subgradient) and two variants of (3.3). $U_{2}$ is modified to have the same quadratic confinement in $x_{1}$ and $x_{2}$ direction and there are several additional local minima due to the last term in the sum. More importantly, compared to (3.3) (and $U_{3}$ ) it has a narrow region near the origin that allows easier passage through $\left\{x_{1}=0\right\}$. On the other hand $U_{3}$ similar to (3.3) except that the well near the global minimum (and the dominant local minimum at $(5,-2)$ ) are elongated in the direction of $x_{2}$ (and $x_{1}$ respectively).

| Cost function | Initial condition | Tolerance sets |
| :---: | :---: | :---: |
| $U_{1}(x)=\frac{1}{2} \sum_{i=1}^{12}\left\|x_{i} \sin \left(x_{i}\right)+0.1 x_{i}\right\|$. | $x_{j}=6 \forall j$ | $x_{j} \in[-2,2] \forall j$ |
| $\begin{gathered} U_{2}\left(x_{1}, x_{2}\right)=\frac{x_{1}^{2}}{7}+\frac{x_{2}^{2}}{7}+5\left(1-e^{-9 x_{2}^{2}}\right) e^{-x_{1}^{2}} \\ -7 e^{-\left(x_{1}+5\right)^{2}-\left(x_{2}-3\right)^{2}}-6 e^{-\left(x_{1}-5\right)^{2}-\left(x_{2}+2\right)^{2}} \\ \quad+\frac{5 x_{1}^{2} e^{-\frac{x_{1}^{2}}{9}} \cos \left(x_{1}+2 x_{2}\right) \cos \left(2 x_{1}-x_{2}\right)}{1+\frac{x_{2}^{2}}{9}} \end{gathered}$ | $x_{1}=4, x_{2}=2$ | $x_{1} \in[-6.5,-4.5], x_{2} \in[1.5,4.5]$ |
| $\begin{gathered} U_{3}\left(x_{1}, x_{2}\right)=\frac{x_{1}^{2}}{5}+\frac{x_{2}^{2}}{10}+5 e^{-x_{1}^{2}}-7 e^{-2\left(x_{1}+5\right)^{2}-\frac{\left(x_{2}-3\right)^{2}}{5}} \\ -6 e^{-\frac{\left(x_{1}-5\right)^{2}}{5}}-2\left(x_{2}+2\right)^{2} \end{gathered}$ | $x_{1}=4, x_{2}=2$ | $x_{1} \in[-6.5,-4.5], x_{2} \in[1.5,4.5]$ |

Table 3.2: Details of three different cost functions, initialisation and tolerance regions corresponding to region of attraction of global minimum.

In Figure 3.2 we present proportions of simulations converging at the region near the global minimum for $U=U_{1}$ depending on $E$ and $\mu$ for (3.2) and on $E$ and $\frac{\bar{\lambda}^{2}}{\mu}$ for (3.1) based on discussion above. To produce the figures related to (3.1) after setting $E, \frac{\bar{\lambda}^{2}}{\mu}$ we pick a random value of $\mu$ from a grid. The aim of this procedure is to ease visualisation, reduce computational cost and to emphasise that it is $\frac{\bar{\lambda}^{2}}{\mu}$ that is crucial for mixing and the performance here is not a product of a tedious tuning for $\mu$. In addition, we only look at $A=A_{1}, A_{2}, A_{3} ; A_{4}$ is omitted due to numerical instabilities when implementing (3.1) for such high dimensional dynamics. The right panels of Figure 3.2 are based on final state and the left on a time average over the last 5000 iterations. In this example it is clear (3.1) results to higher probability of reaching the global minumum. Another interesting observation is that for the generalised Langevin dynamics good performance is more robust to the chosen value of $E$. In this example, this means that adding an additional tuning variable and scaling $\mu$ proportional to $\bar{\lambda}^{2}$, makes it easier to find a configuration of the parameters $E, \mu, \bar{\lambda}$ that leads to good perfomance, compared to using (3.2) and tuning $E, \mu$. We believe this is linked with the increased exploration demonstrated earlier in Figure 3.1.

In Figures 3.3 and 3.4 we present similar results for $U_{2}$ and $U_{3}$. The left panels show proportions of reaching near the correct global minimum calculated using time averages near the final point and the right panels present the number of jumps across $\left\{x_{1}=0\right\}$. All results are averaged over 20 independent runs. The aim here is to measure the extent of exploration of each process similar to Table 3.1. We observe that in both cases using (3.1) leads to higher number of jumps, and this registers as a marginal improvement in the probabilities of reaching the global minumum. We believe the benefit of the higher order dynamics here are the robustness of performance for different values of $E$ and $\frac{\bar{\lambda}^{2}}{\mu}$. This is especially for using $A_{3}$ and $A_{4}$. Finally we note that despite similarities between $U_{2}$ and $U_{3}$ there are significant features that are different: the sharpness in the confinement, the shape and number of attracting wells and the shape of barriers that obstruct crossing regions in the state space. This will have a direct effect in performance, which can explain the difference in performance when comparing Figures 3.3 and $3.4 ; U_{3}$ is a harder cost function to minimise. The generalised Langevin dynamics can improve performance in both cases and Figures 3.3 and 3.4 show that this is possible for a wide region in the tuning variables.

## 4 Conclusions

We explored the possibility of using the generalised Langevin equations in the context of simulated annealing. Our main purpose was to establish convergence as for the underdamped Langevin equation and provide a proof of concept in terms of performance improvement. Although the theoretical results hold for any scaling matrix $A$, we saw in our numerical results that its choice has great impact on the performance. In Section 3, $A_{2}, A_{3}$ or $A_{4}$ seemed to improve the exploration on the state space and the success proportion of the algorithm. There is plenty of work still required in terms of providing a more complete methodology for choosing $A$. This is left as future work and is also closely linked with time discretisation issues as a poor choice for $A$ could lead to numerical integration stiffness. This motivates the development and study of improved numerical integration schemes, in particular, the extension of the conception and analysis on numerical schemes such as BAOAB [37] for the Langevin equation for (1.3) and the extension of the work in [49] for non-identity matrices $\lambda$ and $A$.

In addition, the system in (1.3) is not the only way to add an auxiliary variable to the underdamped Langevin equations in (1.2) whilst retaining the appropriate equilibrium distribution. Our choice was motivated by a clear connection to the generalised Langevin equation (1.4) and link with accelerated gradient descent, but it could be the case that a different third or higher order equations could be used with possibly improved performance. Along these lines, one could consider adding skew-symmetric terms as in [17]. As regards to theory, an interesting extension could involve establishing how the results here can be extended to establish a comparison of optimisation and sampling in a nonconvex setting for an arbitrary number of dimensions similar to [40]. We leave for future work finding optimal constants in the convergence results, investigating dependance on parameters and how the limits of these parameters and constants relate to existing results for the Langevin equation in (1.2) in [47, 55]. Finally, one could also aim to extend large deviation results in $[34,41,59]$ for the overdamped Langevin dynamics to the underdamped and generalised case.

## Acknowledgements

The authors would like to thank Tony Lelievre, Gabriel Stoltz and Urbain Vaes for their helpful remarks. M.C. was funded under a EPSRC studentship. G.A.P. was partially supported by the EPSRC through grants EP/P031587/1, EP/L024926/1, and EP/L020564/1. N.K. and G.A.P. were funded in part by JPMorgan Chase \& Co under a J.P. Morgan A.I. Research Awards 2019. Any views or opinions expressed herein are solely those of the authors listed, and may differ from the views and opinions expressed by JPMorgan Chase \& Co. or its affiliates. This material is not a product of the Research Department of J.P. Morgan Securities LLC. This material does not constitute a solicitation or offer in any jurisdiction.

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Figure 3.1: Dynamics in order from top: (3.2), (3.1) with $A=A_{1}, \ldots, A_{4}$. Left: One instance of noise realisation. Right: Log histogram of 20 independent runs.


Figure 3.2: Proportion of simulations satisfying optimality tolerance for $U=U_{1}$. Panels from top to bottom: (3.2), (3.1) with $A=A_{1}, A_{2}, A_{3}$ ( $A_{4}$ is omitted due to numerical instabilities when implementing (3.1)). Left: Final position, right: time-average of last 5000 iterations. We use $\gamma=3$ for improving visualisation, the results and improvement in using (3.1) are similar for the case of $\gamma=1$. Results here are for 20 independent runs and $k \leq 5 \cdot 10^{4}$-iterations.


Figure 3.3: Both proportion of success and numerical transition rates for $U=U_{2}$. Panels from top to bottom: (3.2), (3.1) with $A=A_{1}, A_{2}, A_{3}, A_{4}$. Left: Proportion satisfying the optimality tolerance for time-average of last 5000 iterations. Right: Average number of crossings across $\left\{x_{1}=0\right\}$ for each independent run. The remaining details are as in caption of Figure 3.2.


Figure 3.4: Results for $U=U_{3}$. Details are as in caption of Figure 3.3.

## Appendices

## Appendix A Notation and preliminaries

For any $\phi \in \mathcal{C}^{\infty}$ and $f: \mathbb{R}^{2 n+m} \rightarrow \mathbb{R}$ smooth enough,

$$
\begin{equation*}
L_{t}(\phi(f))=\phi^{\prime}(f) L_{t}(f)+\phi^{\prime \prime}(f) \Gamma_{t}(f) \tag{A.1}
\end{equation*}
$$

where $\Gamma_{t}$ is the carré du champ operator for $L_{t}$ given by

$$
\begin{equation*}
\Gamma_{t}(f)=\frac{1}{2} L_{t}\left(f^{2}\right)-f L_{t}(f)=\nabla_{z} f \cdot\left(A \nabla_{z} f\right) \tag{A.2}
\end{equation*}
$$

The $L^{2}\left(\mu_{T_{t}}\right)$ adjoint $L_{t}^{*}$ of $L_{t}$ is

$$
L_{t}^{*}=-\left(y \cdot \nabla_{x}-\nabla_{x} U(x) \cdot \nabla_{y}\right)-\left(z^{\top} \lambda \nabla_{y}-y^{\top} \lambda^{\top} \nabla_{z}\right)-T_{t}^{-1} z^{\top} A^{\top} \nabla_{z}+A: D_{z}^{2}
$$

Recall from Monmarché [47, 48] for $\Phi: \mathcal{A}_{+} \rightarrow \mathcal{A}$, where $\mathcal{A}$ and $\mathcal{A}_{+}$are appropriate spaces (namely $\mathcal{A}$ is assumed to be an algebra contained in the domain $\mathcal{D}\left(L_{t}^{*}\right)$ of $L_{t}^{*}$ fixed by $L_{t}^{*}$ and $\mathcal{A}_{+}=\{f \in \mathcal{A}: f \geq 0\}$ ), differentiable in the sense that for any $f, g \in \mathcal{A}_{+}$,

$$
(d \Phi(f) \cdot g)(x) \quad:=\lim _{s \rightarrow 0} \frac{(\Phi(f+s g))(x)-(\Phi(f))(x)}{s}
$$

exists for all $x \in \mathbb{R}^{2 n+m}$, the $\Gamma_{\Phi}$ operator for $L_{t}^{*}$ is defined by

$$
\begin{equation*}
\Gamma_{L_{t}^{*}, \Phi}(h):=\frac{1}{2}\left(L_{t}^{*} \Phi(h)-d \Phi(h) \cdot\left(L_{t}^{*} h\right)\right) \tag{A.3}
\end{equation*}
$$

It will be helpful to keep in mind that $L_{t}^{*}$ is only a term away from satisfying the standard chain and product rules:

$$
\begin{align*}
L_{t}^{*}(\psi(f)) & =\psi^{\prime}(f) L_{t}^{*} f+\psi^{\prime \prime}(f) \nabla_{z} f \cdot\left(A \nabla_{z} f\right)  \tag{A.4}\\
L_{t}^{*}(f g) & =f L_{t}^{*}(g)+g L_{t}^{*}(f)+\frac{1}{2} \nabla_{z} f \cdot\left(\left(A+A^{\top}\right) \nabla_{z} g\right) \tag{A.5}
\end{align*}
$$

for all $f, g \in \mathcal{A}$ and $\psi \in \mathcal{C}^{\infty} \cdot \nabla_{z} f \cdot\left(A \nabla_{z} f\right)$ and $\frac{1}{2} \nabla_{z} f \cdot\left(\left(A+A^{\top}\right) \nabla_{z} g\right)$ are respectively the carré du champ and its symmetric bilinear operator via polarisation for $L_{t}^{*}$.
In addition, square brackets on a scalar-valued $D_{1}$ and a vector-valued operator $D_{2}$ both acting on scalar-valued functions denote the commutator bracket as follows:

$$
\begin{equation*}
\left[D_{1}, D_{2}\right] h=\left(D_{1}\left(D_{2} h\right)_{1}-\left(D_{2} D_{1} h\right)_{1}, \ldots\right) \tag{A.6}
\end{equation*}
$$

for $h$ smooth enough; this will be used as in (C.12).

## Appendix B Auxiliary results

Proposition B.1. For all $t>0$, denote by $\left(X^{T_{t}}, Y^{T_{t}}, Z^{T_{t}}\right)$ a r.v. with distribution $\mu_{T_{t}}$. For any $\delta, \alpha>0$, there exists a constant $\hat{A}>0$ such that

$$
\mathbb{P}\left(U\left(X^{T_{t}}\right)>\min U+\delta\right) \leq \hat{A} e^{-\frac{\delta-\alpha}{T_{t}}}
$$

Proof. The result follows exactly as in Lemma 3 in [47].
Proposition B.2. Under Assumption 1, for all $t>0, X_{t}, Y_{t}, Z_{t}$ are well defined, $\mathbb{E}\left[\left|X_{t}\right|^{2}+\left|Y_{t}\right|^{2}+\left|Z_{t}\right|^{2}\right]<\infty$ and $m_{t} \in \mathcal{C}_{+}^{\infty}=\left\{m \in \mathcal{C}^{\infty}: m>0\right\}$.

Proof. Nonexplosiveness and finiteness of second moments follow as in Proposition 4 in [47] with the modification

$$
G(x, y, t)=\frac{1}{T_{t}}\left(U(x)-\min _{\mathbb{R}^{n}} U+\frac{|y|}{2}+\frac{|z|}{2}\right)
$$

One can establish that $\left(\partial_{t}+L_{t}\right) G \leq C G$ for some constant $C$ depending on $t, T_{t}$. Nonexplosiveness and $\mathbb{E}\left[G\left(X_{t}, Y_{t}, Z_{t}\right]\right]<$ $\infty$ follow by Markov's inequality and Ito's formula, which implies finite second moments for the process.

For smoothness of the law $m_{t}$ of $\left(X_{t}, Y_{t}, Z_{t}\right)$, Theorem 1.2 in [13] can be used. ${ }^{3}$
For positivity of $m_{t}$, the steps in Lemma 3.4 of [42] can be followed: for simplicity, consider the case $m=n$, $\lambda=A=I_{n}$ and $T_{t}=1$, where the associated control problem becomes

$$
\begin{equation*}
\dot{X}_{t}+\ddot{X}_{t}+\dddot{X}_{t}+\nabla U\left(X_{t}\right)+D_{x}^{2} U\left(X_{t}\right) \dot{X}_{t}=\dot{V}_{t}, \quad X_{0}=x_{0} \in \mathbb{R}^{n} \tag{B.1}
\end{equation*}
$$

where $V: \mathbb{R}_{+} \rightarrow \mathbb{R}^{n}$ is a time-varying control and dots indicate partial time derivatives. Given an arbitrary point $X^{*} \in \mathbb{R}^{n}$, set for some fixed $T>0$ the control

$$
V_{t}=\int\left(\frac{X^{*}-x_{0}}{T}+\nabla U\left(\frac{t X^{*}+(T-t) x_{0}}{T}\right)+D_{x}^{2} U\left(\frac{t X^{*}+(T-t) x_{0}}{T}\right) \frac{X^{*}-x_{0}}{T}\right) d t
$$

By the boundedness assumption on the second derivatives of $U$, the unique solution to the control problem (B.1) is

$$
\begin{equation*}
X_{t}=\frac{t X^{*}+(T-t) x_{0}}{T} \tag{B.2}
\end{equation*}
$$

Now since with non-zero probability Brownian motion stays within an $\epsilon$-neighbourhood of any continuously differentiable path, and in particular of $V_{t}$, then positivity of $m_{t}$ follows by the support theorem of Stroock and Varadhan (Theorem 5.2 in [63]).
For the general case the initial and final values for $Y$. and $Z$. are not as above, that is, when $Y_{0}=\dot{X}_{0}, Z_{0}-\nabla U\left(X_{0}\right)=$ $\ddot{X}_{0}, Y_{T}=\dot{X}_{T}$ and $Z_{T}-\nabla U\left(X_{T}\right)=\ddot{X}_{T}$ are arbitrary values. Then it is easy to initialise and finalise $\dot{X}_{t}$ and $\ddot{X}_{t}$ at some small $\hat{t}>0$ and $T-\hat{t}$ respectively and extend the previous argument using a piecewise definition of $V_{t}$, e.g. see Lemma 4.2 and Appendix of [20].

## Appendix C Proof of Proposition 2.3

The following effort up to Proposition C. 8 is towards showing dissipation of a distorted entropy as required in the proof of Theorem 2.4.

## C. 1 Lyapunov function

Lemma C.1. Let $\delta<0$ be a small enough constant and $R: \mathbb{R}^{2 n+m+1} \rightarrow \mathbb{R}$ be defined as

$$
\begin{equation*}
R\left(x, y, z, T_{t}\right) \quad:=\quad U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}+\delta T_{t}\left(y^{\top} \lambda^{-1} z+\frac{1}{2} x \cdot y\right) \tag{C.1}
\end{equation*}
$$

Then there exist constants $a, b, c, d>0$ such that

$$
\begin{equation*}
a\left(|x|^{2}+|y|^{2}+|z|^{2}\right)-d \leq R\left(x, y, z, T_{t}\right) \leq b\left(|x|^{2}+|y|^{2}+|z|^{2}\right)+d \tag{C.2}
\end{equation*}
$$

and

$$
\begin{equation*}
L_{t}(R) \leq-c T_{t} R+d \tag{C.3}
\end{equation*}
$$

[^2]Proof. The first statement is clear by the quadratic assumption (2.5) on $U$ for small enough $\delta$. For the second statement, fix $\delta$ to be small enough for the first statement and additionally to satisfy

$$
\begin{equation*}
\delta \leq \frac{A_{c}}{2}\left[\left(\frac{|\lambda|^{2}}{2 r_{1}}+1+\frac{r_{2}}{r_{1}}\left|\lambda^{-1}\right|^{2}\right)\left(\max _{t \geq 0} T_{t}\right)^{2}+2|A|^{2}\left|\lambda^{-1}\right|^{2}\right]^{-1} \tag{C.4}
\end{equation*}
$$

where $|\cdot|$ is the operator norm and $A_{c}>0$ is the coercivity constant of the positive definite matrix $A$. Consider each of the terms of $L_{t}(R)$ seperately,

$$
\begin{equation*}
L_{t}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)=-\frac{1}{T_{t}} z^{\top} A z+\operatorname{Tr} A \tag{C.5}
\end{equation*}
$$

Using the quadratic bound (2.7) on $\nabla_{x} U$,

$$
\begin{align*}
L_{t}\left(y^{\top} \lambda^{-1} z\right)= & -\nabla_{x} U \lambda^{-1} z+|z|^{2}-|y|^{2}-T_{t}^{-1} z^{\top} A\left(\lambda^{-1}\right)^{\top} y \\
\leq & \frac{r_{1}}{4 r_{2}}\left|\nabla_{x} U\right|^{2}+\frac{r_{2}}{r_{1}}\left|\lambda^{-1}\right|^{2}|z|^{2}+|z|^{2}-|y|^{2}+\frac{|y|^{2}}{4} \\
& +T_{t}^{-2}|A|^{2}\left|\lambda^{-1}\right|^{2}|z|^{2} \\
\leq & \frac{r_{1}}{4}|x|^{2}+\frac{r_{1}}{4 r_{2}} U_{g}-\frac{3}{4}|y|^{2} \\
& +\left(1+\frac{r_{2}}{r_{1}}\left|\lambda^{-1}\right|^{2}+T_{t}^{-2}|A|^{2}\left|\lambda^{-1}\right|^{2}\right)|z|^{2} \tag{C.6}
\end{align*}
$$

Then also with the bound (2.6) for $\nabla_{x} U \cdot x$,

$$
\begin{align*}
L_{t}(x \cdot y) & =|y|^{2}-\nabla_{x} U \cdot x+z^{\top} \lambda x \\
& \leq|y|^{2}-r_{1}|x|^{2}+U_{g}+\frac{|\lambda|^{2}}{r_{1}}|z|^{2}+\frac{r_{1}}{4}|x|^{2} \tag{C.7}
\end{align*}
$$

Combining (C.5), (C.6), (C.7), given a large enough $C>0$,

$$
\begin{aligned}
L_{t}\left(R\left(x, y, z, T_{t}\right)\right)= & L_{t}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)+\delta T_{t} L_{t}\left(y^{\top} \lambda z\right)+\frac{\delta T_{t}}{2} L_{t}(x \cdot y) \\
\leq & -\delta T_{t} \frac{r_{1}}{8}|x|^{2}-\delta T_{t} \frac{1}{4}|y|^{2}-\frac{1}{T_{t}} z^{\top} A z+C \\
& +\delta T_{t}\left[\frac{|\lambda|^{2}}{2 r_{1}}+\left(1+\frac{r_{2}}{r_{1}}\left|\lambda^{-1}\right|^{2}+T_{t}^{-2}|A|^{2}\left|\lambda^{-1}\right|^{2}\right)\right]|z|^{2}
\end{aligned}
$$

Therefore for $\delta$ satisfying the bound (C.4), the $|z|^{2}$ term can be bounded,

$$
\begin{aligned}
L_{t}\left(R\left(x, y, z, T_{t}\right)\right) & \leq-\delta T_{t} \frac{r_{1}}{8}|x|^{2}-\delta T_{t} \frac{1}{4}|y|^{2}-\frac{A_{c}}{2 \max _{t \geq 0} T_{t}}|z|^{2}+C \\
& \leq-\frac{D T_{t}}{b} R+C^{\prime}
\end{aligned}
$$

where $D>0$ is small enough, $C^{\prime}>0$ is large enough and the right inequality of (C.2) has been used.

Lemma C.2. For $2 \leq p \leq \bar{p}$ with $p, \bar{p} \in \mathbb{N}$ from Assumption 1 on $m_{0}, \frac{\mathbb{E}\left[R\left(X_{t}, Y_{t}, Z_{t}, T_{t}\right)^{p}\right]}{(\ln (e+t))^{p}}$ is bounded uniformly in time.
Proof. It is equivalent to prove the result for $R+d>0$ in place of $R$ for any $p \leq \bar{p}$. Use induction w.r.t. $p$. The case for $p=0$ is obvious. Let

$$
R_{t}:=R\left(X_{t}, Y_{t}, Z_{t}, T_{t}\right)
$$

Consider the following terms separately

$$
\frac{d}{d t} \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right]=\partial_{t} \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right]+T_{t}^{\prime} \partial_{T_{t}} \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right]
$$

Firstly, by definition (C.1) of $R$ and the left bound of (C.2),

$$
\begin{aligned}
T_{t}^{\prime} \partial_{T_{t}} \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right] & =T_{t}^{\prime} \mathbb{E}\left[p\left(R_{t}+d\right)^{p-1} \delta\left(y^{\top} \lambda^{-1} z+\frac{1}{2} x \cdot y\right)\right] \\
& \leq\left|T_{t}^{\prime}\right| \mathbb{E}\left[p\left(R_{t}+d\right)^{p-1} \delta\left|y^{\top} \lambda^{-1} z+\frac{1}{2} x \cdot y\right|\right] \\
& \leq \frac{B_{p}}{t} \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right]
\end{aligned}
$$

for a constant $B_{p} \geq 0$.
With (A.1) and (A.2), for $p \geq 1$, using property (C.3) from Lemma C. 1 and again the left bound of (C.2), consider the auxiliary expression

$$
\begin{aligned}
\partial_{t} \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right]= & \mathbb{E}\left[\left(L_{t}\left((R+d)^{p}\right)\right)\left(X_{t}, Y_{t}, Z_{t}, T_{t}\right)\right] \\
= & \mathbb{E}\left[p\left(R_{t}+d\right)^{p-1}\left(L_{t}(R)\right)\left(X_{t}, Y_{t}, Z_{t}, T_{t}\right)\right) \\
& \left.+p(p-1)\left(R_{t}+d\right)^{p-2}\left(\Gamma_{t}(R)\right)\left(X_{t}, Y_{t}, Z_{t}, T_{t}\right)\right] \\
\leq & \mathbb{E}\left[p\left(R_{t}+d\right)^{p-1}\left(-c T_{t} R_{t}+d\right)\right. \\
& \left.+p(p-1)\left(R_{t}+d\right)^{p-2}\left(Z_{t}+\delta T_{t}\left(\lambda^{-1}\right)^{\top} Y_{t}\right) \cdot\left(A\left(Z_{t}+\delta T_{t}\left(\lambda^{-1}\right)^{\top} Y_{t}\right)\right)\right] \\
\leq & -c p T_{t} \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right]+A_{p} \mathbb{E}\left[\left(R_{t}+d\right)^{p-1}\right]
\end{aligned}
$$

for a constant $A_{p} \geq 0$. Then with the induction assumption, the slow-decay assumption on $T_{t}$, for large enough $t$,

$$
\begin{aligned}
& \frac{d}{d t}\left(e^{\frac{c}{2} p \int_{0}^{t} T_{s} d s} \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right]\right) \leq A_{p} e^{\frac{c}{2} p \int_{0}^{t} T_{s} d s}(\ln (e+t))^{p-1} \\
& \mathbb{E}\left[\left(R_{t}+d\right)^{p}\right] \leq \mathbb{E}\left[\left(R_{0}+d\right)^{p}\right] e^{-\frac{c}{2} p \int_{0}^{t} T_{s} d s}+\int_{0}^{t} A_{p} e^{-\frac{c}{2} p \int_{s}^{t} T_{u} d u}(\ln (e+s))^{p-1} d s \\
& \leq \mathbb{E}\left[\left(R_{0}+d\right)^{p}\right]+A_{p}(\ln (e+t))^{p-1} \int_{0}^{t} e^{-\frac{c}{2} p T_{t}(t-s)} d s \\
& \leq \mathbb{E}\left[\left(R_{0}+d\right)^{p}\right]+A_{p}(\ln (e+t))^{p-1} \frac{1-e^{-\frac{c}{2} p T_{t} t}}{\frac{c}{2} p T_{t}}
\end{aligned}
$$

Using the assumptions on $m_{0}$ and again the assumption $T_{t} \geq E(\ln t)^{-1}$, the result follows.

Corollary C.3. For any $2 \leq p \leq \bar{p}, \frac{\mathbb{E}\left[\left(\left|X_{t}\right|^{2}+\left|Y_{t}\right|^{2}+\left|Z_{t}\right|^{2}\right)^{p}\right]}{(\ln (e+t))^{p}}$ is bounded uniformly in time.
Proof. By the lower bound on $R$ in (C.2),

$$
\mathbb{E}\left[\left(\left|X_{t}\right|^{2}+\left|Y_{t}\right|^{2}+\left|Z_{t}\right|^{2}\right)^{p}\right] \leq \mathbb{E}\left[\left(\frac{R\left(X_{t}, Y_{t}, Z_{t}, T_{t}\right)+d}{a}\right)^{p}\right]
$$

which concludes by Lemma C. 2 after expanding.

## C. 2 Form of Distorted Entropy

Let $H(t)$ be the distorted entropy

$$
\begin{align*}
H(t):= & \int\left(\frac{\left.\left|2 \nabla_{x} h_{t}+8 S_{0}\left(\nabla_{y} h_{t}+\lambda^{-1} \nabla_{z} h_{t}\right)\right|^{2}\right)}{h_{t}}+\frac{\left|\nabla_{y} h_{t}+S_{1} \lambda^{-1} \nabla_{z} h_{t}\right|^{2}}{h_{t}}\right. \\
& \left.+\beta\left(T_{t}^{-1}\right) h_{t} \ln \left(h_{t}\right)\right) d \mu_{T_{t}}, \tag{C.8}
\end{align*}
$$

where $S_{0}, S_{1}>0$ are the constants

$$
\begin{align*}
& S_{0}:=\left(1+\left|D_{x}^{2} U\right|_{\infty}^{2}\right)^{\frac{1}{2}}  \tag{C.9}\\
& S_{1}:=2+156 S_{0}^{2}+1024 S_{0}^{4} \tag{C.10}
\end{align*}
$$

and $\beta$ is a second order polynomial

$$
\begin{equation*}
\beta\left(T_{t}^{-1}\right):=1+\beta_{0}+\beta_{1} T_{t}^{-1}+\beta_{2} T_{t}^{-2} \tag{C.11}
\end{equation*}
$$

with large enough coefficients $\beta_{0}, \beta_{1}, \beta_{2}>0$ depending on $\left|D_{x}^{2} U\right|_{\infty},\left|A^{\top}\right|$ and

$$
\hat{\lambda}^{2}:=\max \left(|\lambda|^{2},\left|\lambda^{\top}\right|^{2},\left|\lambda^{-1}\right|^{2},\left|\lambda^{-1}\right|\left|\lambda^{\top}\right|\right)
$$

Remark C.1. This particular expression for $H$ is not necessarily the best choice and it is quite possible to have only $\nabla_{x}$ and $\nabla_{y}$ and no $\nabla_{z}$ appearing in the first integrand for instance. However the above is a working expression and optimality of this is left as future work.

First, an auxiliary result which can be found as Lemma 12 of [47] is stated along with its proofs from [47] since the proof is not too long. Notice the $\Phi^{*}$ appearing in Lemma C. 4 appears in the first two terms of $H(t)$.

Lemma C.4. For

$$
\Phi^{*}(h)=\frac{|M \nabla h|^{2}}{h}
$$

where $M$ is matrix-valued, we have

$$
\Gamma_{L_{t}^{*}, \Phi^{*}}(h) \geq \frac{(M \nabla h) \cdot\left[L_{t}^{*}, M \nabla\right] h}{h}
$$

for all $0<h \in \mathcal{A}_{+}$, where the square bracket denotes the commutator vector (A.6), i.e.

$$
\begin{equation*}
\left[L_{t}^{*}, M \nabla\right] h=\left(L_{t}^{*}(M \nabla h)_{1}-\left(M \nabla L_{t}^{*} h\right)_{1}, \ldots\right) \tag{C.12}
\end{equation*}
$$

Proof. The second term in definition (A.3) of $\Gamma_{L_{t}^{*}, \Phi}(h)$ can be calculated to be

$$
\begin{equation*}
-d \Phi^{*}(h) \cdot L_{t}^{*} h=-\frac{2}{h}(M \nabla h) \cdot\left(M \nabla L_{t}^{*} h\right)+\frac{L_{t}^{*} h}{h^{2}}|M \nabla h|^{2} \tag{C.13}
\end{equation*}
$$

Using the adjusted product and chain rules (A.5) and (A.4) for $L_{t}^{*}$, the first term in the definition (A.3) of $\Gamma_{L_{t}^{*}, \Phi}$ can be calculated to be

$$
\begin{aligned}
L_{t}^{*}\left(\Phi^{*}(h)\right)= & \frac{1}{h} L_{t}^{*}\left(|M \nabla h|^{2}\right)+|M \nabla h|^{2} L_{t}^{*}\left(\frac{1}{h}\right) \\
& -\sum_{i} \frac{2}{h^{2}}(M \nabla h)_{i} \nabla_{z} h \cdot\left(\left(A+A^{\top}\right) \nabla_{z}(M \nabla h)_{i}\right) \\
= & \frac{2}{h}\left((M \nabla h) \cdot L_{t}^{*} M \nabla h+\sum_{i} \nabla_{z}(M \nabla h)_{i} \cdot\left(A \nabla_{z}(M \nabla h)_{i}\right)\right) \\
& +|M \nabla h|^{2} L_{t}^{*}\left(\frac{1}{h}\right)-\sum_{i} \frac{2}{h^{2}}(M \nabla h)_{i} \nabla_{z} h \cdot\left(\left(A+A^{\top}\right) \nabla_{z}(M \nabla h)_{i}\right)
\end{aligned}
$$

where the last summands can be bounded below with the positive definite property of $A$,

$$
\begin{aligned}
-\frac{2}{h^{2}} & (M \nabla h)_{i} \nabla_{z} h \cdot\left(\left(A+A^{\top}\right) \nabla_{z}(M \nabla h)_{i}\right) \\
& =-\frac{2}{h^{2}}(M \nabla h)_{i}\left(\nabla_{z} h\right)^{\top} A\left(\nabla_{z}(M \nabla h)_{i}\right)-\frac{1}{h^{2}}\left(\nabla_{z}(M \nabla h)_{i}\right)^{\top} A\left((M \nabla h)_{i} \nabla_{z} h\right) \\
& \geq-\frac{2}{h^{3}}(M \nabla h)_{i}^{2}\left(\nabla_{z} h\right)^{\top} A \nabla_{z} h-\frac{2}{h}\left(\nabla_{z}(M \nabla h)_{i}\right)^{\top} A \nabla_{z}(M \nabla h)_{i},
\end{aligned}
$$

which produces the bound

$$
L_{t}^{*}\left(\Phi^{*}(h)\right) \geq \frac{2}{h}(M \nabla h) \cdot L_{t}^{*} M \nabla h+|M \nabla h|^{2} L_{t}^{*}\left(\frac{1}{h}\right)-\frac{2}{h^{3}}|M \nabla h|^{2} \nabla_{z} h \cdot\left(A \nabla_{z} h\right) .
$$

Combining this with (C.13) then using the adjusted chain rule (A.4),

$$
\begin{aligned}
\Gamma_{L_{t}^{*}, \Phi^{*}}(h) \geq & \frac{1}{h}(M \nabla h) \cdot\left[L_{t}^{*}, M \nabla\right] h+\frac{1}{2}|M \nabla h|^{2}\left(L_{t}^{*}\left(\frac{1}{h}\right)+\frac{L_{t}^{*} h}{h^{2}}\right) \\
& -\frac{1}{h^{3}}|M \nabla h|^{2} \nabla_{z} h \cdot\left(A \nabla_{z} h\right) \\
= & \frac{1}{h}(M \nabla h) \cdot\left[L_{t}^{*}, M \nabla\right] h .
\end{aligned}
$$

With Lemma C.4, the distorted entropy (C.8) can be shown to be a correct one with the following proposition. Proposition C.5. Let $\Psi_{T_{t}}$ be the operator appearing in the integrand of the distorted entropy $H$, that is

$$
\begin{aligned}
& \Psi_{T_{t}}(h):=\frac{\left|2 \nabla_{x} h+8 S_{0}\left(\nabla_{y} h+\lambda^{-1} \nabla_{z} h\right)\right|^{2}}{h}+\frac{\left|\nabla_{y} h+S_{1} \lambda^{-1} \nabla_{z} h\right|^{2}}{h} \\
& +\beta\left(T_{t}^{-1}\right) h \ln (h)
\end{aligned}
$$

for $h \in \mathcal{A}_{+}$. It holds that

$$
\begin{equation*}
\Gamma_{L_{t}^{*}, \Psi_{T_{t}}}(h) \geq \frac{|\nabla h|^{2}}{h} \tag{C.14}
\end{equation*}
$$

Remark C.2. $H$ satisfying this property is crucial for proving dissipation in Proposition C. 8 and was the main consideration when making remark C.1.

Proof. Let $\Phi_{1}, \Phi_{2}, \Phi_{3}$ be defined by

$$
\Psi_{T_{t}}=: \Phi_{1}+\Phi_{2}+\beta\left(T_{t}^{-1}\right) \Phi_{3}
$$

Note that the $\Gamma_{\Phi}$ operator is linear in the second operator argument by linearity of $L_{t}^{*}$, so that (C.14) is

$$
\Gamma_{L_{t}^{*}, \Phi_{1}}(h)+\Gamma_{L_{t}^{*}, \Phi_{2}}(h)+\beta\left(T_{t}^{-1}\right) \Gamma_{L_{t}^{*}, \Phi_{3}}(h) \geq \frac{|\nabla h|^{2}}{h} .
$$

Consider $\Gamma_{L_{t}^{*}, \Phi_{3}}$ first. Using the definition (A.3) of $\Gamma_{L_{t}^{*}, \Phi}$ and the product and chain rule (A.5) and (A.4) for $L_{t}^{*}$,

$$
\begin{align*}
\Gamma_{L_{t}^{*}, \Phi_{3}}(h) & =\frac{1}{2}\left(\ln h L_{t}^{*} h+h L_{t}^{*} \ln h+\frac{2\left|\nabla_{z} h\right|^{2}}{h}-(1+\ln h) L_{t}^{*} h\right) \\
& =\frac{1}{2}\left(\ln h L_{t}^{*} h+L_{t}^{*} h+\frac{\left|\nabla_{z} h\right|^{2}}{h}-(1+\ln h) L_{t}^{*} h\right) \\
& =\frac{\left|\nabla_{z} h\right|^{2}}{2 h} . \tag{C.15}
\end{align*}
$$

Since the goal is to show (C.14), the availability of the term (C.15) means that regardless of how negative of a contribution $\Gamma_{L_{t}^{*}, \Phi_{1}}$ and $\Gamma_{L_{t}^{*}, \Phi_{2}}$ makes in the $z$-derivative term in (C.15), it is not a concern; this is reflected in the factor $\beta$.
For $\Gamma_{L_{t}^{*}, \Phi_{1}}$ and $\Gamma_{L_{t}^{*}, \Phi_{2}}$ we use $S_{0}, S_{1}>0$ defined as before in (C.9)-(C.10).
Lemma C. 4 gives

$$
\begin{aligned}
h \Gamma_{L_{t}^{*}, \Phi_{1}}(h) \geq & \left(2 \nabla_{x}+8 S_{0}\left(\nabla_{y}+\lambda^{-1} \nabla_{z}\right)\right) h \cdot\left[L_{t}^{*}, 2 \nabla_{x}+8 S_{0}\left(\nabla_{y}+\lambda^{-1} \nabla_{z}\right)\right] h \\
= & \left(2 \nabla_{x}+8 S_{0}\left(\nabla_{y}+\lambda^{-1} \nabla_{z}\right)\right) h \cdot\left(-2\left(D_{x}^{2} U\right) \nabla_{y}+8 S_{0}\left(\nabla_{x}-\lambda^{\top} \nabla_{z}\right.\right. \\
& \left.+\nabla_{y}+T_{t}^{-1} \lambda^{-1} A^{\top} \nabla_{z}\right) h \\
= & 16 \\
\quad & S_{0}\left|\nabla_{x} h\right|^{2}+2 \nabla_{x} h \cdot\left(\left(-2 D_{x}^{2} U+8 S_{0}\right) \nabla_{y} h\right) \\
& \quad+\nabla_{x} h \cdot\left(8 S_{0}\left(-\lambda^{\top}+T_{t}^{-1} \lambda^{-1} A^{\top}\right) \nabla_{z} h\right)+64 S_{0}^{2} \nabla_{x} h \cdot \nabla_{y} h \\
& +8 S_{0} \nabla_{y} h \cdot\left(\left(-2 D_{x}^{2} U+8 S_{0}\right) \nabla_{y} h\right) \\
& +8 S_{0} \nabla_{y} h \cdot\left(8 S_{0}\left(-\lambda^{\top}+T_{t}^{-1} \lambda^{-1} A^{\top}\right) \nabla_{z} h\right)+64 S_{0}^{2} \nabla_{x} h \cdot\left(\lambda^{-1} \nabla_{z} h\right) \\
& +\left(\left(-2 D_{x}^{2} U+8 S_{0}\right) \nabla_{y} h\right) \cdot\left(8 S_{0} \lambda^{-1} \nabla_{z} h\right) \\
& +8 S_{0}\left(\lambda^{-1} \nabla_{z} h\right) \cdot\left(8 S_{0}\left(-\lambda^{\top}+T_{t}^{-1} \lambda^{-1} A^{\top}\right) \nabla_{z} h\right) .
\end{aligned}
$$

In order to get a bound in terms of $\left(\partial_{i} h\right)^{2}$ terms rather than $\partial_{i} h \partial_{j} h$ terms, bounding $\partial_{i} h \partial_{j} h$ terms with some care and using the boundedness assumption (2.4) on the second derivatives of $U$ yield

$$
\begin{aligned}
h \Gamma_{L_{t}^{*}, \Phi_{1}}(h) \geq & 16 S_{0}\left|\nabla_{x} h\right|^{2}-\left(2\left|\nabla_{x} h\right|^{2}+2\left|D_{x}^{2} U\right|_{\infty}^{2}\left|\nabla_{y} h\right|^{2}+8\left|\nabla_{x} h\right|^{2}+8 S_{0}^{2}\left|\nabla_{y} h\right|^{2}\right) \\
& -\left(2\left|\nabla_{x} h\right|^{2}+8 S_{0}^{2} \hat{\lambda}^{2}\left(1+T_{t}^{-2}\left|A^{\top}\right|^{2}\right)\left|\nabla_{z} h\right|^{2}\right) \\
& -\left(\left|\nabla_{x} h\right|^{2}+1024 S_{0}^{4}\left|\nabla_{y} h\right|^{2}\right)-\left(16 S_{0}\left|D_{x}^{2} U\right|_{\infty}\left|\nabla_{y} h\right|^{2}\right. \\
& \left.+64 S_{0}^{2}\left|\nabla_{y} h\right|^{2}\right)-\left(32 S_{0}^{2}\left|\nabla_{y} h\right|^{2}+32 S_{0}^{2} \hat{\lambda}^{2}\left(1+T_{t}^{-2}\left|A^{\top}\right|^{2}\right)\left|\nabla_{z} h\right|^{2}\right) \\
& -\left(\left|\nabla_{x} h\right|^{2}+1024 S_{0}^{2} \hat{\lambda}^{2}\left|\nabla_{z} h\right|^{2}\right)-\left(\left(2\left|D_{x}^{2} U\right|_{\infty}^{2}+32 S_{0}^{2}\right)\left|\nabla_{y} h\right|^{2}\right. \\
& \left.+32 S_{0}^{2} \hat{\lambda}^{2}\left|\nabla_{z} h\right|^{2}\right)-64 S_{0}^{2} \hat{\lambda}^{2}\left(1+T_{t}^{-2}\left|A^{\top}\right|^{2}\right)\left|\nabla_{z} h\right|^{2} \\
\geq & 2\left|\nabla_{x} h\right|^{2}+S_{0}^{2}\left(-156-1024 S_{0}^{2}\right)\left|\nabla_{y} h\right|^{2} \\
& +S_{0}^{2} \hat{\lambda}^{2}\left(-1160-104 T_{t}^{-2}\left|A^{\top}\right|^{2}\right)\left|\nabla_{z} h\right|^{2} .
\end{aligned}
$$

$\Gamma_{L_{t}^{*}, \Phi_{2}}$ compensates for the negative $y$ derivatives:

$$
\begin{aligned}
h \Gamma_{L_{t}^{*}, \Phi_{2}}(h) \geq & \left(\nabla_{y}+S_{1} \lambda^{-1} \nabla_{z}\right) h \cdot\left[L_{t}^{*}, \nabla_{y}+S_{1} \lambda^{-1} \nabla_{z}\right] h \\
= & \left(\nabla_{y}+S_{1} \lambda^{-1} \nabla_{z}\right) h \cdot\left(\nabla_{x}-\lambda^{\top} \nabla_{z}+S_{1} \nabla_{y}+S_{1} T_{t}^{-1} \lambda^{-1} A^{\top} \nabla_{z}\right) h \\
= & \nabla_{x} h \cdot \nabla_{y} h-\nabla_{y} h \cdot\left(\lambda^{\top} \nabla_{z} h\right)+S_{1}\left|\nabla_{y} h\right|^{2}+S_{1} T_{t}^{-1} \nabla_{y} h \cdot\left(\lambda^{-1} A^{\top} \nabla_{z} h\right) \\
& +S_{1} \nabla_{x} h \cdot\left(\lambda^{-1} \nabla_{z} h\right)-S_{1}\left(\lambda^{-1} \nabla_{z} h\right) \cdot\left(\lambda^{\top} \nabla_{z} h\right) \\
& +S_{1}^{2} \nabla_{y} h \cdot\left(\lambda^{-1} \nabla_{z} h\right)+S_{1}^{2} T_{t}^{-1}\left(\lambda^{-1} \nabla_{z} h\right) \cdot\left(\lambda^{-1} A^{\top} \nabla_{z} h\right),
\end{aligned}
$$

where using the inequalities

$$
\begin{aligned}
\nabla_{x} h \cdot \nabla_{y} h & \geq-\frac{1}{2}\left|\nabla_{x} h\right|^{2}-\frac{1}{2}\left|\nabla_{y} h\right|^{2} \\
-\nabla_{y} h \cdot\left(\lambda^{\top} \nabla_{z} h\right) & \geq-\frac{1}{6}\left|\nabla_{y} h\right|^{2}-\frac{3}{2}\left|\lambda^{\top}\right|^{2}\left|\nabla_{z} h\right|^{2} \\
S_{1} T_{t}^{-1} \nabla_{y} h \cdot\left(\lambda^{-1} A^{\top} \nabla_{z} h\right) & \geq-\frac{1}{6}\left|\nabla_{y} h\right|^{2}-\frac{3}{2} S_{1}^{2} T_{t}^{-2}\left|\lambda^{-1}\right|^{2}\left|A^{\top}\right|^{2}\left|\nabla_{z} h\right|^{2} \\
S_{1} \nabla_{x} h \cdot\left(\lambda^{-1} \nabla_{z} h\right) & \geq-\frac{1}{2}\left|\nabla_{x} h\right|^{2}-\frac{1}{2} S_{1}^{2}\left|\lambda^{-1}\right|^{2}\left|\nabla_{z} h\right|^{2} \\
S_{1}^{2} \nabla_{y} h \cdot\left(\lambda^{-1} \nabla_{z} h\right) & \geq-\frac{1}{6}\left|\nabla_{y} h\right|^{2}-\frac{3}{2} S_{1}^{4}\left|\lambda^{-1}\right|^{2}
\end{aligned}
$$

gives the bound

$$
\begin{aligned}
h \Gamma_{L_{t}^{*}, \Phi_{2}}(h) \geq & -\left|\nabla_{x} h\right|^{2}+\left(1+156 S_{0}^{2}+1024 S_{0}^{2}\right)\left|\nabla_{y} h\right|^{2}-\frac{1}{2} \hat{\lambda}^{2}\left(3+S_{1}+S_{1}^{2}\right. \\
& \left.+3 S_{1}^{4}+S_{1}^{2} T_{t}^{-1}\left|A^{\top}\right|+3 S_{1}^{2} T_{t}^{-2}\left|A^{\top}\right|^{2}\right)\left|\nabla_{z} h\right|^{2}
\end{aligned}
$$

Putting together the bounds gives (C.14) given the coefficents (C.11) in $\beta$ are large enough.

## C. 3 Log-Sobolev Inequality

Proposition C.6. The distorted entropy (C.8) satisfies

$$
\begin{equation*}
H\left(h_{t}\right) \leq C_{t} \int \frac{\left|\nabla h_{t}\right|^{2}}{h_{t}} d \mu_{T_{t}} \tag{C.16}
\end{equation*}
$$

where

$$
\begin{equation*}
C_{t}=A_{*}+\beta\left(T_{t}^{-1}\right) e^{\left(U_{M}-U_{m}\right) T_{t}^{-1}} \frac{T_{t}}{4} \max \left(2, a_{m}^{-2}\right) \tag{C.17}
\end{equation*}
$$

for all $t>0$ and some constant $A_{*}>0$ depending on $\left|D_{x}^{2} U\right|_{\infty}$ and $\lambda$.
Proof. The first two terms in the integrand of $H(t)$ after expanding lead directly to the inequality result for $A_{*}$. For the last term of $H(t)$, the standard log-Sobolev inequality for a Gaussian measure [29] alongside the properties that log-Sobolev inequalities tensorises and are stable under perturbations, which can be found as Theorem 4.4 and Property 4.6 in [30] respectively, yields the result. Since the proof of Property 4.6 in [30] is not too long, it is repeated for $U$ satisfying the quadratic assumption (2.5) in order to get the precise form of $C_{t}$ :

$$
\begin{aligned}
\int h_{t} \ln h_{t} d \mu_{T_{t}} & =\int\left(h_{t} \ln h_{t}-h_{t}+1\right) d \mu_{T_{t}} \\
& \leq \int\left(h_{t} \ln h_{t}-h_{t}+1\right) Z_{T_{t}}^{-1} e^{-\frac{U_{m}}{T_{t}}} e^{-\frac{1}{T_{t}}\left(|\bar{a} \circ x|^{2}+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)} d x d y d z \\
& =e^{-\frac{U_{m}}{T_{t}}} Z_{T_{t}}^{-1} \int h_{t} \ln h_{t} e^{-\frac{1}{T_{t}}\left(|\bar{a} \circ x|^{2}+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)} d x d y d z \\
& \leq e^{-\frac{U_{m}}{T_{t}}} \max \left(\frac{T_{t}}{2}, \max _{i} \frac{T_{t}}{4 \bar{a}_{i}^{2}}\right) Z_{T_{t}}^{-1} \int \frac{\left|\nabla h_{t}\right|^{2}}{h_{t}} e^{-\frac{1}{T_{t}}\left(|\bar{a} \circ x|^{2}+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)} d x d y d z \\
& \leq e^{\frac{U_{M}-U_{m}}{T_{t}}} \max \left(\frac{T_{t}}{2}, \frac{T_{t}}{4 a_{m}^{2}}\right) \int \frac{\left|\nabla h_{t}\right|^{2}}{h_{t}} d \mu_{T_{t}}
\end{aligned}
$$

where the first inequality follows by (2.5) since $x \ln x-x+1 \geq 0$ for all $x \geq 0$.

## C. 4 Proof of Dissipation

To see that $H(t)$ dissipates over time, it will helpful to be able to pass a time derivative under the integral sign in $H(t)$, which would be straightforward for compactly supported smooth integrands but is not so for $H(t)$. Lemma C. 7 constructs compactly supported functions that when multiplied with the integrand in $H(t)$ gives sufficient properties for retrieving a bound on $\partial_{t} H(t)$ after passing the derviative under the integral sign.

The key nontrivial sufficient property turns out to be (C.18) below.
Let $\varphi: \mathbb{R} \rightarrow \mathbb{R}$ be the mollifier

$$
\begin{gathered}
\varphi(x):= \begin{cases}e^{\frac{1}{x^{2}-1}}\left(\int_{-1}^{1} e^{\frac{1}{y^{2}-1}} d y\right)^{-1} & \text { if }-1<x \leq 1 \\
0 & \text { otherwise },\end{cases} \\
\varphi_{m}(x):=\frac{1}{m} \varphi\left(\frac{x}{m}\right)
\end{gathered}
$$

and

$$
\nu_{m}:=\varphi_{m} * \mathbb{1}_{\left(-\infty, m^{2}\right]} \leq 1
$$

for $m>0$ where $\mathbb{1}_{\left(-\infty, m^{2}\right]}$ is the indicator function on $\left(-\infty, m^{2}\right]$.
Lemma C.7. The smooth functions $\eta_{m}: \mathbb{R}^{2 n+m+1} \rightarrow \mathbb{R}^{n}$

$$
\eta_{m}=\nu_{m}(\ln (R+2 d)),
$$

where $d>0$ is the negative of the lower bound of $R$ as in (C.2),

1. are compactly supported,
2. converge to 1 pointwise,
3. satisfy for some constant $C>0$ independent of $m$ and $t$

$$
\begin{equation*}
L_{t} \eta_{m} \leq \frac{C}{m} \tag{C.18}
\end{equation*}
$$

Remark C.3. Lemma C. 7 is different to Lemma 16 in [47]. We believe the few first equations in the proof of Lemma 16 [47] is incorrect and contradicts with equation (4) in the same paper. As a result we require proving (C.18) instead of Lemma 17 of [47].

Proof. By the quadratic assumption (2.5) on $U$ and the bound (C.2) on $R, R$ grows quadratically and in particular is bounded below by an arbitrarily large constant at infinity; along with the support of $\nu_{m}$ being bounded above, the first statement is clear. The second statement is also easy to check.
With the chain rule (A.1) and (A.2) for $L_{t}$,

$$
L_{t} \eta_{m}=\nu_{m}^{\prime}(\ln (R+2 d)) L_{t} \ln (R+2 d)+\nu_{m}^{\prime \prime}(\ln (R+2 d))\left(\nabla_{z} \ln (R+2 d)\right)^{\top} A \nabla_{z} \ln (R+2 d)
$$

It can be seen that $\nu_{m}^{\prime}$ and $\nu_{m}^{\prime \prime}$ are of at most order $m^{-1}$, explicitly:

$$
\begin{aligned}
& \nu_{m}(x)=\int_{-\infty}^{m^{2}} \varphi_{m}(x-y) d y=\int_{-\infty}^{x+m^{2}} \varphi_{m}(z) d z \\
& \nu_{m}^{\prime}(x)=\varphi_{m}\left(x+m^{2}\right) \leq m^{-1} \max \varphi \\
& \nu_{m}^{\prime \prime}(x)=\varphi_{m}^{\prime}\left(x+m^{2}\right) \leq m^{-2} \max \varphi^{\prime}
\end{aligned}
$$

Therefore for a constant $\bar{C}>0$,

$$
L_{t} \eta_{m} \leq \bar{C}\left(m^{-1} L_{t} \ln (R+2 d)+m^{-2}\left(\nabla_{z} \ln (R+2 d)\right)^{\top} A \nabla_{z} \ln (R+2 d)\right)
$$

A quick calculation using the Lyapunov property (C.3) of $R$ and again the chain rule (A.1) and (A.2) for $L_{t}$ reveals

$$
\begin{aligned}
L_{t} \ln (R+2 d) & =\frac{L_{t} R}{R+2 d}-\frac{\left(\nabla_{z} R\right)^{\top} A \nabla_{z} R}{(R+2 d)^{2}} \\
& \leq \frac{-c T_{t} R+d}{R+2 d}-\frac{A_{c}\left|\nabla_{z} R\right|^{2}}{(R+2 d)^{2}} \\
& \leq \frac{-c T_{t}(R+d)+c T_{t} d+d}{R+2 d}-\frac{A_{c}\left|\nabla_{z} R\right|^{2}}{(R+2 d)^{2}} \\
\left(\nabla_{z} \ln (R+2 d)\right)^{\top} A \nabla_{z} \ln (R+2 d) & \leq|A|\left|\nabla_{z} \ln (R+2 d)\right|^{2} \\
& =|A|\left|\frac{\nabla_{z} R}{R+2 d}\right|^{2}
\end{aligned}
$$

which are bounded over $\mathbb{R}^{2 n+m+1}$ considering $R$ grows quadratically in space.
The proof of Proposition C. 8 follows very closely to Lemma 19 of [47] and its preceding lemmas.
Remark C.4. As explained in Remark C.3, Lemma C. 7 uses a different smooth compact function for the truncation than the one found in Section 5.2 in [47]. Lemma C. 7 is designed to establish (C.21) below, which is pivotal for the proof.

Proposition C.8. For any $\alpha>0$, there exists some constant $B>0$ and some $t_{H}>0$ both depending on $\left|D_{x}^{2} U\right|_{\infty}$, $A, \lambda, T_{t}, \alpha, U_{M}, U_{m}, U_{g}, E, r_{2}$ and $a_{m}$, such that for all $t>t_{H}$,

$$
\begin{equation*}
H(t) \leq B\left(\frac{1}{t}\right)^{1-\frac{U_{M}-U_{m}}{E}-2 \alpha} \tag{C.19}
\end{equation*}
$$

Proof. Consider the auxiliary distorted entropies

$$
\begin{aligned}
H_{\eta_{m}}(t)= & \int \eta_{m}\left(\frac{\left|2 \nabla_{x} h_{t}+8 S_{0}\left(\nabla_{y} h_{t}+\lambda^{-1} \nabla_{z} h_{t}\right)\right|^{2}}{h_{t}}+\frac{\left|\nabla_{y} h_{t}+S_{1} \lambda^{-1} \nabla_{z} h_{t}\right|^{2}}{h_{t}}\right. \\
& \left.+\beta\left(T_{t}^{-1}\right) h_{t} \ln \left(h_{t}\right)\right) d \mu_{T_{t}} \\
= & \int \eta_{m}\left(\Phi_{1}\left(h_{t}\right)+\Phi_{2}\left(h_{t}\right)+\beta\left(T_{t}^{-1}\right) \Phi_{3}\left(h_{t}\right)\right) d \mu_{T_{t}}=\int \eta_{m} \Psi_{T_{t}}\left(h_{t}\right) d \mu_{T_{t}}
\end{aligned}
$$

where recall $h_{t}=m_{t} \mu_{T_{t}}^{-1}$ is the ratio (2.3) between the law of the process and its instantaneous equilibrium. The point of Lemma C. 7 was so that a time derivative can be pushed under the integral:

$$
\begin{equation*}
\frac{d}{d t} H_{\eta_{m}}(t)=\int \eta_{m} \partial_{t}\left(\Psi_{T_{t}}\left(h_{t}\right)\right) d \mu_{T_{t}}+T_{t}^{\prime} \int \eta_{m} \partial_{T_{t}}\left(\Psi_{T_{t}}\left(h_{t}\right) \mu_{T_{t}}\right) d x d y d z \tag{C.20}
\end{equation*}
$$

where $\partial_{t}$ is fixed with respect to $T_{t}$ and vice versa for $\partial_{T_{t}}$.
Consider the terms separately. Since $m_{t}$ is the law of the process (1.3) and $L_{t}^{*}$ is the $L^{2}\left(\mu_{T_{t}}\right)$ adjoint of $L_{t}$,

$$
\int f \partial_{t} m_{t}=\int f L_{t}^{T} m_{t}=\int L_{t} f m_{t}=\int L_{t} f \frac{m_{t}}{\mu_{T_{t}}} \mu_{T_{t}}=\int f L_{t}^{*}\left(\frac{m_{t}}{\mu_{T_{t}}}\right) \mu_{T_{t}}
$$

where $L_{t}^{T}$ is the $L^{2}\left(\mathbb{R}^{2 n+m}\right)$ adjoint of $L_{t}$, yielding $\partial_{t} m_{t}=L_{t}^{*} h_{t} \mu_{T_{t}}$. This allows the first term in (C.20) to be
bounded as follows.

$$
\begin{align*}
\int \eta_{m} \partial_{t}\left(\Psi_{T_{t}}\left(h_{t}\right)\right) d \mu_{T_{t}}= & \int \eta_{m} d \Psi_{T_{t}}\left(h_{t}\right) \cdot \partial_{t} h_{t} d \mu_{T_{t}} \\
= & \int \eta_{m} d \Psi_{T_{t}}\left(h_{t}\right) \cdot \frac{\partial_{t} m_{t}}{\mu_{T_{t}}} d \mu_{T_{t}} \\
= & \int \eta_{m} d \Psi_{T_{t}}\left(h_{t}\right) \cdot L_{t}^{*} h_{t} d \mu_{T_{t}} \\
= & -\int \eta_{m} 2 \Gamma_{L_{t}^{*}, \Psi_{T_{t}}}\left(h_{t}\right) d \mu_{T_{t}}+\int \eta_{m} L_{t}^{*}\left(\Psi_{T_{t}}\left(h_{t}\right)\right) d \mu_{T_{t}} \\
= & -\int \eta_{m} 2 \Gamma_{L_{t}^{*}, \Psi_{T_{t}}}\left(h_{t}\right) d \mu_{T_{t}} \\
& +\int L_{t} \eta_{m}\left(\Psi_{T_{t}}\left(h_{t}\right)+\beta\left(T_{t}^{-1}\right) e^{-1}\right) d \mu_{T_{t}} \\
\leq & -2 \int \eta_{m} \frac{\left|\nabla h_{t}\right|^{2}}{h_{t}} d \mu_{T_{t}}+\frac{C}{m} \int\left(\Psi_{T_{t}}\left(h_{t}\right)+\beta\left(T_{t}^{-1}\right) e^{-1}\right) d \mu_{T_{t}} \tag{C.21}
\end{align*}
$$

by Proposition C. 5 and Lemma C. 7 where $\beta\left(T_{t}^{-1}\right) e^{-1}$ is added to force

$$
\beta\left(T_{t}^{-1}\right)\left(h_{t} \ln h_{t}+e^{-1}\right) \geq 0 \Longrightarrow \Psi_{T_{t}}\left(h_{t}\right)+\beta\left(T_{t}^{-1}\right) e^{-1} \geq 0
$$

(C.21) is a satisfactory bound for now - after taking the limit $m \rightarrow \infty$, the log-Sobolev inequality (C.16) will bound the first term on the right hand side of (C.21) such that a Grönwall-type argument can be applied.

For the second term in (C.20), consider the $\Phi_{1}$ and $\Phi_{2}$ terms in the integrand $\eta_{m} \partial_{T_{t}}\left(\Psi_{T_{t}} \mu_{T_{t}}\right)=\eta_{m} \partial_{T_{t}}\left(\left(\Phi_{1}+\Phi_{2}+\right.\right.$ $\left.\beta\left(T_{t}^{-1}\right) \Phi_{3}\right) \mu_{T_{t}}$ ) of $H_{\eta_{m}}(t)$ as

$$
\eta_{m} \partial_{T_{t}}\left(\Phi_{i}\left(h_{t}\right) \mu_{T_{t}}\right)=\eta_{m} \partial_{T_{t}}\left|M_{i} \nabla \ln \left(\frac{m_{t}}{\mu_{T_{t}}}\right)\right|^{2} m_{t}, \quad i=1,2
$$

for the corresponding matrices $M_{1}$ and $M_{2}$. With the form (2.2) of the equilibrium $\mu_{T_{t}}$, this yields

$$
\begin{equation*}
\eta_{m} \partial_{T_{t}}\left(\Phi_{i}\left(h_{t}\right) \mu_{T_{t}}\right)=-2 \eta_{m}\left(M_{i} \nabla \ln h_{t} \cdot M_{i} \nabla \partial_{T_{t}} \ln \mu_{T_{t}}\right) m_{t} \tag{C.22}
\end{equation*}
$$

where the trickiest part, using $Z_{T_{t}}=\int_{\mathbb{R}^{2 n+m}} e^{-\frac{1}{T_{t}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)} d x d y d z$, gives

$$
\begin{align*}
\partial_{T_{t}} & \ln \mu_{T_{t}} \\
& =\mu_{T_{t}}^{-1} \partial_{T_{t}}\left(Z_{T_{t}}^{-1} e^{-\frac{1}{T_{t}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)}\right) \\
& =\mu_{T_{t}}^{-1}\left(-Z_{T_{t}}^{-2} \partial_{T_{t}} Z_{T_{t}}-\frac{Z_{T_{t}}^{-1}}{T_{t}^{2}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)\right) e^{-\frac{1}{T_{t}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)} \\
& =\mu_{T_{t}}^{-1}\left(-\mu_{T_{t}} Z_{T_{t}}^{-1} \partial_{T_{t}} Z_{T_{t}}-\frac{\mu_{T_{t}}}{T_{t}^{2}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)\right) \\
& =\int_{\mathbb{R}^{2 n+m}} \frac{1}{T_{t}^{2}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right) d \mu_{T_{t}}-\frac{1}{T_{t}^{2}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right) \tag{C.23}
\end{align*}
$$

Integrating by parts in each of the variables $x, y, z$, where for the $x$ variable $\nabla_{x} U(x) \cdot x$ is used, after some manipulation using the quadratic assumptions (2.5) and (2.6) for $U(x)$ :

$$
\begin{equation*}
p_{1}\left(T_{t}^{-1}\right) \leq \partial_{T_{t}} \ln \mu_{T_{t}}+\frac{1}{T_{t}^{2}}\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}-\frac{n+m}{2} T_{t}\right) \leq p_{2}\left(T_{t}^{-1}\right) \tag{C.24}
\end{equation*}
$$

where $p_{1}(x)=\frac{2 a_{m}^{2} n}{r_{2}+1} x-\left(\frac{a_{m}^{2} U_{g}}{r_{2}+1}+U_{m}\right) x^{2}$ and $p_{2}(x)=\frac{a_{M}^{2} n}{r_{1}} x+\left(\frac{a_{M}^{2} U_{g}}{r_{1}}+U_{M}\right) x^{2}$ are found from bounding the integral over $U(x)$.
Substituting (C.23) back into (C.22),

$$
\begin{align*}
\eta_{m} \partial_{T_{t}}\left(\Phi_{i}\left(h_{t}\right) \mu_{T_{t}}\right) & \leq \eta_{m}\left(\left|M_{i} \nabla \ln h_{t}\right|^{2}+T_{t}^{-4}\left|M_{i} \nabla\left(U(x)+\frac{|y|^{2}}{2}+\frac{|z|^{2}}{2}\right)\right|^{2}\right) m_{t} \\
& \leq \Phi_{i}\left(h_{t}\right) \mu_{T_{t}}+\widetilde{C} T_{t}^{-4}\left(1+|x|^{2}+|y|^{2}+|z|^{2}\right) m_{t} \tag{C.25}
\end{align*}
$$

for constant $\widetilde{C} \geq 0$ by the quadratic assumption (2.7) on $\left|\nabla_{x} U\right|^{2}$ and $\eta_{m} \leq 1$.
In the second term of (C.20), $\Phi_{3}\left(h_{t}\right)=\frac{m_{t}}{\mu T_{t}} \ln \frac{m_{t}}{\mu_{t}}$, using the left inequality of (C.24), gives

$$
\begin{align*}
& \eta_{m} \partial_{T_{t}}\left(\beta\left(T_{t}^{-1}\right) \Phi_{3}\left(h_{t}\right) \mu_{T_{t}}\right) \\
&=-\eta_{m} T_{t}^{-2} \beta^{\prime}\left(T_{t}^{-1}\right) \Phi_{3}\left(h_{t}\right) \mu_{T_{t}}-\eta_{m} \beta\left(T_{t}^{-1}\right) \partial_{T_{t}} \ln \mu_{T_{t}} m_{t} \\
&=-\eta_{m} T_{t}^{-2} \beta^{\prime}\left(T_{t}^{-1}\right)\left(\Phi_{3}\left(h_{t}\right)+e^{-1}\right) \mu_{T_{t}}+\eta_{m} T_{t}^{-2} \beta^{\prime}\left(T_{t}^{-1}\right) e^{-1} \mu_{T_{t}} \\
&-\eta_{m} \beta\left(T_{t}^{-1}\right) \partial_{T_{t}} \ln \mu_{T_{t}} m_{t} \\
& \leq T_{t}^{-2} \beta^{\prime}\left(T_{t}^{-1}\right) e^{-1} \mu_{T_{t}} \\
&+\beta\left(T_{t}^{-1}\right)\left|p_{1}\left(T_{t}^{-1}\right)+\frac{1}{T_{t}^{2}}\left(\frac{n+m}{2} T_{t}-U_{M}-|\bar{a} \circ x|^{2}-\frac{|y|^{2}}{2}-\frac{|z|^{2}}{2}\right)\right| m_{t} \tag{C.26}
\end{align*}
$$

where in the last step $\eta_{m} \leq 1$ and (2.5) have been used.
Putting together the bounds (C.25) and (C.26) and applying Corollary C. 3 yields

$$
\begin{align*}
\int \eta_{m} \partial_{T_{t}}\left(\Psi_{T_{t}}\left(h_{t}\right) \mu_{T_{t}}\right) d x d y d z & \leq q\left(T_{t}^{-1}\right)\left(H(t)+\mathbb{E}\left[1+\left|X_{t}\right|^{2}+\left|Y_{t}\right|^{2}+\left|Z_{t}\right|^{2}\right]\right) \\
& \leq p\left(T_{t}^{-1}\right)(H(t)+\hat{C}) \tag{C.27}
\end{align*}
$$

where $p$ and $q$ are some finite order polynomial with nonnegative coefficients and $\hat{C} \geq 0$.
Returning to (C.20), collecting (C.21) and (C.27) then integrating from any $s$ to $t$ gives

$$
\begin{aligned}
H_{\eta_{m}}(t)-H_{\eta_{m}}(s) \leq & \int_{s}^{t}\left(-2 \int \eta_{m} \frac{\left|\nabla h_{u}\right|^{2}}{h_{u}} d \mu_{T_{u}}+\left|T_{u}^{\prime}\right| p\left(T_{u}^{-1}\right)(H(u)+\hat{C})\right) d u \\
& +\mathcal{O}\left(m^{-1}\right)
\end{aligned}
$$

After taking $m \rightarrow \infty$ then applying Fatou's lemma for the first terms of both sides and $\eta_{m} \leq 1$ for the second term on the left, this becomes

$$
\begin{align*}
H(t)-H(s) & \leq \int_{s}^{t}-2 \int \frac{\left|\nabla h_{u}\right|^{2}}{h_{u}} d \mu_{T_{u}} d u+\int_{s}^{t}\left|T_{u}^{\prime}\right| p\left(T_{u}^{-1}\right)(H(u)+\hat{C}) d u  \tag{C.28}\\
& \leq \int_{s}^{t}\left(\left(\left|T_{u}^{\prime}\right| p\left(T_{u}^{-1}\right)-2 C_{u}^{-1}\right) H(u)+\hat{C}\left|T_{u}^{\prime}\right| p\left(T_{u}^{-1}\right)\right) d u \tag{C.29}
\end{align*}
$$

where the last inequality follows by the log-Sobolev inequality (C.16).
Taking $s \rightarrow t$, the derivative of $H$ can be seen to be bounded in such a way that Theorem 2.17 of [66] can be applied to see that

$$
\begin{equation*}
H(t)<\infty \quad \forall t \geq 0 \tag{C.30}
\end{equation*}
$$

After taking $s \rightarrow t$, it remains to show that

$$
\begin{equation*}
2 C_{t}^{-1} \gg\left|T_{t}^{\prime}\right| p\left(T_{t}^{-1}\right) \tag{C.31}
\end{equation*}
$$

for large $t$.
Here the assumptions on $T_{t}$ are imposed. Considering the form (C.17) of $C_{t}$ and that $t^{\alpha} \gg(\ln t)^{p}$ for any $p>0$ and $\alpha>0$ given large enough $t$, this yields that for any $\alpha>0$, there exists $t_{1}>0$ and $c_{i}>0, i=1,2,3$ such that for all $t \geq t_{1}$,

$$
\frac{d}{d t} H(t) \leq\left(c_{1}\left(\frac{1}{t}\right)^{1-\alpha}-c_{2}\left(\frac{1}{t}\right)^{\frac{U_{M}-U_{m}}{E}+\alpha}\right) H(t)+c_{3}\left(\frac{1}{t}\right)^{1-\alpha}
$$

Since $E>U_{M}-U_{m}$ by assumption, taking $\alpha$ small enough, there exists $c_{4}>0$ and $t_{2}>0$ such that for all $t \geq t_{2}$,

$$
\begin{equation*}
\frac{d}{d t} H(t) \leq-c_{4} H(t)\left(\frac{1}{t}\right)^{\frac{U_{M}-U_{m}}{E}+\alpha}+c_{3}\left(\frac{1}{t}\right)^{1-\alpha} \tag{C.32}
\end{equation*}
$$

Setting

$$
\gamma_{1}(t):=c_{4}\left(\frac{1}{t}\right)^{\frac{U_{M}-U_{m}}{E}+\alpha}, \quad \gamma_{2}(t):=c_{3}\left(\frac{1}{t}\right)^{1-\alpha}
$$

and following the argument as per [47] from Lemma 6 in [46], (C.32) becomes

$$
\begin{equation*}
\frac{d}{d t}\left(H(t) e^{\int_{t_{2}}^{t} \gamma_{1}(s) d s}\right) \leq e^{\int_{t_{2}}^{t} \gamma_{1}(s) d s} \gamma_{2}(t)=\frac{\gamma_{2}(t)}{\gamma_{1}(t)} \gamma_{1}(t) e^{\int_{t_{2}}^{t} \gamma_{1}(s) d s} \tag{C.33}
\end{equation*}
$$

but for any $t \geq t_{2}$ and $t^{*} \leq t$, taking $\alpha \leq \frac{1}{2}\left(1-\frac{U_{M}-U_{m}}{E}\right)$,

$$
\begin{equation*}
\frac{\gamma_{2}(t)}{\gamma_{1}(t)}=\frac{c_{3}}{c_{4}}\left(\frac{1}{t}\right)^{1-\frac{U_{M}-U_{m}}{E}-2 \alpha} \leq \frac{c_{3}}{c_{4}}\left(\frac{1}{t^{*}}\right)^{1-\frac{U_{M}-U_{m}}{E}-2 \alpha} \tag{C.34}
\end{equation*}
$$

which allows (C.33) to give

$$
H(t) \leq H\left(t_{2}\right) e^{-\int_{t_{2}}^{t} \gamma_{1}(s) d s}+\frac{c_{3}}{c_{4}}\left(\frac{1}{t^{*}}\right)^{1-\frac{U_{M}-U_{m}}{E}-2 \alpha}\left(1-e^{-\int_{t_{2}}^{t} \gamma_{1}(s) d s}\right)
$$

The proof is finished after substituting back in $t^{*}=t$ since $H\left(t_{2}\right)$ is finite due to (C.30) and the expression

$$
\int_{t_{2}}^{t} \gamma_{1}(s) d s=c_{5}\left(\left(\frac{1}{t}\right)^{-1+\frac{U_{M}-U_{m}}{E}+\alpha}-\left(\frac{1}{t_{2}}\right)^{-1+\frac{U_{M}-U_{m}}{E}+\alpha}\right)
$$

grows to infinity as $t \rightarrow \infty$, where $c_{5}>0$ is a constant.
Remark C.5. The annealing schedule $T_{t}$ is chosen to satisfy the relationship (C.31) between $C_{t}^{-1}$ and $\left|T_{t}^{\prime}\right| p\left(T_{t}^{-1}\right)$. (C.31) has two purposes - for the coefficient of $H(u)$ on the right hand side of (C.29) to be negative and for this coefficient to be much stronger in magnitude for large $t$ than the last term in (C.29). These allow (C.34) and consequently for (C.33) to be a fruitful step.

## Appendix D Additional Results

We conclude the appendices by presenting the analog of Proposition C. 8 for the $T_{t}=T>0$ sampling case and a result about the choice of the annealing schedule.

Proposition D.1. Let 1. and 3. of Assumption 1 hold and let $T_{t}=T$ for all $t$ for some constant $T>0$. It holds that

$$
\int\left|h_{t}-1\right| d \mu_{T} \leq \sqrt{\frac{2 H(0)}{\beta(T)}} e^{-\frac{C_{*}^{-1}}{2} t}
$$

where $C_{*}$ is the log-Sobolev constant

$$
\begin{equation*}
C_{*}=A_{*}+\beta\left(T^{-1}\right) e^{\left(U_{M}-U_{m}\right) T^{-1}} \frac{T}{4} \max \left(2, a_{m}^{-2}\right) \tag{D.1}
\end{equation*}
$$

Proof. After Pinsker's inequality (2.9) and consideration of the definition (C.8) of $H$, what remains is the partial time derivative part of the proof of Proposition C.8. The proof concludes by the same calculations, keeping in mind $T_{t}^{\prime}=0$, until (C.29) followed by the Grönwall argument.

Proposition D.2. The schedule $T_{t}=\frac{E}{\ln (e+t)}, E>U_{M}-U_{m}$ is optimal in the sense that for any differentiable $f: \mathbb{R}_{+} \rightarrow \mathbb{R}_{+}$, if

$$
\begin{equation*}
T_{t}=\frac{1}{f(t)}\left(\frac{U_{M}-U_{m}}{\ln (e+t)}\right) \tag{D.2}
\end{equation*}
$$

$C_{t}$ is the log-Sobolev factor (C.17) and $p$ is the polynomial with nonnegative coefficients from the proof of Proposition C.8, then the relation (C.31):

$$
2 C_{t}^{-1} \gg\left|T_{t}^{\prime}\right| p\left(T_{t}^{-1}\right)
$$

holds for large times only if $\lim \sup _{t \rightarrow \infty} f(t) \leq 1$.
Proof. Suppose there exists a constant $\delta>0$ and times $\left(t_{i}\right)_{i \in \mathbb{N}}$ such that $0<t_{i} \rightarrow \infty$ and

$$
f\left(t_{i}\right) \geq 1+\delta \quad \forall i
$$

From its definition (C.17),

$$
C_{t}^{-1} \sim \mathcal{O}\left(e^{-\left(U_{M}-U_{m}\right) T_{t}^{-1}} T_{t}^{-1}\right)
$$

which after substituting in the form (D.2) for $T_{t}$ is

$$
\begin{equation*}
e^{-\left(U_{M}-U_{m}\right) T_{t}^{-1}} T_{t}^{-1}=(e+t)^{-f(t)} \frac{f(t) \ln (e+t)}{U_{M}-U_{m}} \sim \mathcal{O}\left(t^{-f(t)} f(t) \ln t\right) \tag{D.3}
\end{equation*}
$$

Compare this to

$$
\begin{equation*}
\left|T_{t}^{\prime}\right| p\left(T_{t}^{-1}\right) \propto \frac{p(f(t) \ln (e+t))}{(f(t) \ln (e+t))^{2}}\left(\frac{f(t)}{e+t}+\left|f^{\prime}(t)\right| \ln (e+t)\right) \tag{D.4}
\end{equation*}
$$

which has order at least $(t f(t))^{-1}(\ln t)^{-2}$. For $t=t_{i}$ large enough, $f(t) \geq 1+\delta$ and so

$$
\begin{equation*}
t^{-f(t)} f(t) \ln t \ll(t f(t))^{-1}(\ln t)^{-2} \tag{D.5}
\end{equation*}
$$

violating (C.31).
Remark D.1. One can strengthen the proposition by making precise the form of $p$ from Proposition C.8, which will determine how slowly $f(t)$ is allowed to converge to 1 ; in fact $p$ should be at least fifth order. This appears inconsequential with respect to optimality and so is omitted.


[^0]:    ${ }^{1}$ under more restrictive conditions on the objective function or the initial condition than those considered here

[^1]:    ${ }^{2}$ To the best of the authors' knowledge, there is no known direct translation between (1.4) and (1.3) for a non-constant $T_{t}$; such a translation quite possibly exists and at the very least the intuition here is useful.

[^2]:    ${ }^{3}$ This work caters to the unbounded coefficients in the stochastic differential equation, the time-dependence of one of the coefficients and the need for using the general Hörmander's condition involving the ' $X_{0}$ ' operator rather than the restrictive one, see [45] for definitions. Indeed, the constant assumption 2.(iii) on the annealing schedule for a small interval at the beginning is used so that this theorem can be applied.

